title: Graphs, Algorithms, and Optimization Discrete Mathematics and Its Applications
author: Kocay, William.; Kreher, Donald L.
publisher: CRC Press
isbn10 | asin: 0203489055
print isbn13: 9780203620892
ebook isbn13: 9780203489055
language:
subject English Graph algorithms.
publication date: 2005
Icc: QA166.245.K63 2005eb
ddc: 511/.5
subject: Graph algorithms.
cover
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DISCRETE MATHEMATICS AND ITS APPLICATIONS
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GRAPHS, ALGORITHMS, AND OPTI MI ZATI ON

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Series Editor KENNETH H.ROSEN
GRAPHS, ALGORITHMS, AND OPTI MI ZATI ON
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CHAPMAN \& HALLICRC
A CRC Press Company
Boca Raton London NewYork Washington, D.C.

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This edition published in the Taylor \& Francis e-Library, 2005.
To purchase your own copy of this or any of Taylor \& Francis or Routledge's collection of thousands of eBooks please go to www.eBookstore.tandf.co.uk.
Library of Congress Cataloging-in-Publication Data
Kocay, William.
Graphs, algorithms, and optimization/William Kocay, Donald L.Kreher.
p. cm.-(Discrete mathematics and its applications) Includes bibliographical references and index.
ISBN 1-58488-396-0 (alk. paper)

1. Graph algorithms. I. Kreher, Donald L. II. Title. III. Series.

QA166.245.K63 2004
511'.5-dc22 2004056153
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© 2005 by Chapman \& Hall/CRC Press
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ISBN 0-203-48905-5 Master e-book ISBN
ISBN 0-203-62089-5 (OEB Format)
International Standard Book Number 1-58488-396-0 (Print Edition)
Library of Congress Card Number 2004056153
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Page vii
The authors would like to take this opportunity to express their appreciation and gratitude to the following
people who have had a very significant effect on their mathematical development:
Adrian Bondy, Earl Kramer, Spyros Magliveras, Ron Read, and Ralph Stanton.
This book is dedicated to the memory of
William T.Tutte, (1917-2002)
"the greatest of the graphmen"
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## Preface

Our objective in writing this book is to present the theory of graphs from an algorithmic viewpoint. We present the graph theory in a rigorous, but informal style and cover most of the main areas of graph theory.
The ideas of surface topology are presented from an intuitive point of view. We have also included a discussion on linear programming that emphasizes problems in graph theory. The text is suitable for students in computer science or mathematics programs.
Graph theory is a rich source of problems and techniques for programming and data structure development, as well as for the theory of computing, including NP-completeness and polynomial reduction.
This book could be used a textbook for a third or fourth year course on graph algorithms which contains a programming content, or for a more advanced course at the fourth year or graduate level. It could be used in a course in which the programming language is any major programming language (e.g., C, C ++ , J ava). The algorithms are presented in a generic style and are not dependent on any particular programming language. The text could also be used for a sequence of courses like "Graph Algorithms I" and "Graph Algorithms II". The courses offered would depend on the selection of chapters included. A typical course will begin with Chapters 1, 2, 3, and 4. At this point, a number of options are available.
A possible first course would consist of Chapters 1, 2, 3, 4, 6, 8, 9, 10, 11, and 12, and a first course stressing optimization would consist of Chapters $1,2,3,4,8,9,10,14,15$, and 16 . Experience indicates that the students consider these substantial courses. One or two chapters could be omitted for a lighter course. We would like to thank the many people who provided encouragement while we wrote this book, pointed out typos and errors, and gave useful suggestions. In particular, we would like to convey our thanks to Ben Li and J ohn van Rees of the University of Manitoba for proofreading some chapters.
William Kocay
Donald L.Kreher

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Donald L.Kreher obtained his Ph.D. from the University of Nebraska in 1984. He has held academic positions at Rochester Institute of Technology and the University of Wyoming. He is currently a University Professor of Mathematical Sciences at Michigan Technological University, where he teaches and conducts research in combinatorics and combinatorial algorithms. He has published numerous research papers and is a co-author of the internationally acclaimed text "Combinatorial Algorithms: Generation Enumeration and Search", CRC Press, 1999. He serves on the editorial boards of two journals.
Professor Kreher is the sole recipient of the 1995 Marshall Hall Medal, awarded by the Institute of Combinatorics and its Applications.

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## Page 1

1
Graphs and Their Complements

### 1.1 I ntroduction

The diagram in Figure 1.1 illustrates a graph. It is called the graph of the cube. The edges of the geometric cube correspond to the line segments connecting the nodes in the graph, and the nodes correspond to the corners of the cube where the edges meet. They are the vertices of the cube.


FI GURE 1.1

## The graph of a cube

This diagram is drawn so as to resemble a cube, but if we were to rearrange it, as in Figure 1.2, it would still be the graph of the cube, although it would no longer look like a cube. Thus, a graph is a graphical representation of a relation in which edges connect pairs of vertices.
DEFINITION 1.1: A simple graph $G$ consists of a vertex set $V(G)$ and an edge set $E(G)$, where each edge is a pair $\{u, u\}$ of vertices $u, v \in V(G)$.
We denote the set of all pairs of a set $V$ by $\binom{V}{2}$. Then $E(G) \subseteq\binom{V(G)}{2}$. In

Page 2


## FI GURE 1.2

## The graph of the cube

the example of the cube, $V(G)=\{1,2,3,4,5,6,7,8\}$, and $E(G)=\{12,23,34,14,15,26,37,48,56,67,78$, 58\}, where we have used the shorthand notation $u u$ to stand for the pair $\{u, u\}$. If $u, v \in V(G)$, then $u \longrightarrow v$ means that $u$ is joined to $u$ by an edge. We say that $u$ and $u$ are adjacent. We use this notation to remind us of the linked list data structure that we will use to store a graph in the computer. Similarly, $u \nrightarrow v$ means that $u$ is not joined to $u$. We can also express these relations by writing $u v \in E(G)$ or $u v \notin E(G)$, respectively. Note that in a simple graph if $u \longrightarrow v$, then $v \longrightarrow u$. If $u$ is adjacent to each of $u 1, u 2, \ldots, u k$, then we write $u \longrightarrow\left\{u_{1}, u_{2}, \ldots, u_{k}\right\}$.
These graphs are called simple graphs because each pair $u, u$ of vertices is joined by at most one edge. Sometimes we need to allow several edges to join the same pair of vertices. Such a graph is also called a multigraph. An edge can then no longer be defined as a pair of vertices, (or the multiple edges would not be distinct), but to each edge there still corresponds a pair $\{u, u\}$. We can express this formally by saying that a graph $G$ consists of a vertex set $V(G)$, an edge set $E(G)$, and a correspondence $\psi: E(G) \rightarrow\binom{V(G)}{2}$. Given an edge $e \in E(G), \psi(e)$ is a pair $\{u, u\}$ which are the endpoints of e. Different edges can then have the same endpoints. We shall use simple graphs most of the time, which is why we prefer the simpler definition, but many of the theorems and techniques will apply to multigraphs as well.
This definition can be further extended to graphs with loops as well. A loop is an edge in which both endpoints are equal. We can include this in the general definition of a graph by making the mapping $\psi: E(G) \rightarrow\binom{V(G)}{2} \cup V(G)$. An edge $e \in E(G)$ for which $\psi(e)=u \in V(G)$ defines a loop. Figure 1.2 shows a graph with multiple edges and loops. However, we shall use simple graphs most of the time, so that an
edge will be considered to be a pair of vertices.
The number of vertices of a graph $G$ is denoted $|G|$. It is called the order of $G$.

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## FI GURE 1.3

## A multigraph

The number of edges is $\varepsilon(G)$. If $G$ is simple, then obviously $\varepsilon(G) \leq\binom{|G|}{2}$, since $E(G) \subseteq\binom{V(G)}{2}$. We shall often use node or point as synonyms for vertex.
Many graphs have special names. The complete graph $K n$ is a simple graph with $|K n|=n$ and $\varepsilon=\binom{n}{2}$. The empty graph $\bar{K}_{n}$ is a graph with $\left|\bar{K}_{n}\right|=n$ and $\varepsilon=0 . \bar{K}_{n}$ is the complement of $K n$.


FI GURE 1.4
The complete graph K5
DEFINITION 1.2: Let $G$ be a simple graph. The complement of $G$ is $\bar{G}$, where $V(\bar{G})=V(G)$ and $E(\bar{G})=\binom{V(G)}{2} \backslash E(G)$.
$E(\bar{G})$ consists of all those pairs uu which are not edges of $G$. Thus, $u v \in E(\bar{G})$
page_3
Page 4
if and only if $u v \notin E(G)$. Figure 1.5 show a graph and its complement.


## FI GURE 1.5

## A graph and its complement

Figure 1.6 shows another graph and its complement. Notice that in this case, when $\bar{G}$ is redrawn, it looks identical to $G$.
In a certain sense, this $G$ and $\bar{G}$ are the same graph. They are not equal, since $E(G) \neq E(\bar{G})$, but it is clear that they have the same structure. If two graphs have the same structure, then they can only differ in the names of the vertices. Therefore, we can rename the vertices of one to make it exactly equal to the other graph. In the example above, we can rename the vertices of $G$ by the mapping $\theta$ given by

$$
\begin{array}{rrrrrr}
k: & 1 & 2 & 3 & 4 & 5 \\
\hline \theta(k): & 1 & 3 & 5 & 2 & 4
\end{array},
$$

then $\theta(G)$ would equal $\bar{G}$. This kind of equivalence of graphs is known as isomorphism. Observe that a one-to-one mapping $\theta$ of the vertices of a graph $G$ can be extended to a mapping of the edges of $G$ by defining $\theta(\{u, u\})=\{\theta(u), \theta(u)\}$.


G

$\bar{G}$

$\bar{G}$

FI GURE 1.6
Another graph and its complement
page_4

## Page 5

DEFINITION 1.3: Let $G$ and $H$ be simple graphs. $G$ and $H$ are isomorphic if there is a one-to-one correspondence $\theta: V(G) \rightarrow V(H)$ such that $\theta(E(G))=E(H)$, where $\theta(E(G))=\{\theta(u v): u v \in E(G)\}$.
We write $G \cong H$ to denote isomorphism. If $G \cong H$, then $u v \in E(G)$ if and only if $\theta(u v) \in E(H)$. One way to determine whether $G \cong H$ is to try and redraw $G$ so as to make it look identical to $H$. We can then read off the mapping $\theta$ from the diagram. However, this is limited to small graphs. For example, the two graphs $G$ and $H$ shown in Figure 1.7 are isomorphic, since the drawing of $G$ can be transformed into $H$ by first moving vertex 2 to the bottom of the diagram, and then moving vertex 5 to the top. Comparing the two diagrams then gives the mapping

$$
\begin{array}{rrrrrrr}
k: & 1 & 2 & 3 & 4 & 5 & 6 \\
\hline \theta(k): & 6 & 4 & 2 & 5 & 1 & 3
\end{array}
$$

as an isomorphism.


## FI GURE 1.7

## Two isomorphic graphs

It is usually more difficult to determine when two graphs $G$ and $H$ are not isomorphic than to find an isomorphism when they are isomorphic. One way is to find a portion of $G$ that cannot be part of H . For example, the graph H of Figure 1.7 is not isomorphic to the graph of the prism, which is illustrated in Figure 1.8, because the prism contains a triangle, whereas $H$ has no triangle. A subgraph of a graph $G$ is a graph $X$ such that $V(X) \subseteq V(G)$ and $E(X) \subseteq E(G)$. If $\theta: G \rightarrow H$ is a possible isomorphism, then $\theta(X)$ will be a subgraph of $H$ which is isomorphic to $X$. A subgraph $X$ is an induced subgraph if for every $u$, $v \in V(X) \subseteq V(G), u v \in E(X)$ if and only if $u v \in E(G)$.
The degree of a vertex $u \in V(G)$ is $\operatorname{DEG}(u)$, the number of edges which contain $u$. If $k=\operatorname{DEG}(u)$ and $u \longrightarrow\left\{u_{1}, u_{2}, \ldots, u_{k}\right\}$, then $\theta(u) \longrightarrow$

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## FI GURE 1.8

## The graph of the prism

$\left\{\theta\left(u_{1}\right), \theta\left(u_{2}\right), \ldots, \theta\left(u_{k}\right)\right\}$, so that $\operatorname{DEG}(u)=\operatorname{DEG}(\theta(u))$. Therefore a necessary condition for $G$ and $H$ to be isomorphic is that they have the same set of degrees. The examples of Figures 1.7 and 1.8 show that this is not a sufficient condition.
In Figure 1.6, we saw an example of a graph $G$ that is isomorphic to its complement. There are many such graphs.
DEFINITION 1.4: A simple graph $G$ is self-complementary if $G \cong \bar{G}$.
LEMMA 1.1 If $G$ is a self-complementary graph, then $|G| \equiv 0$ or $1(\bmod 4)$.
PROOF If $G \cong \bar{G}$, then $\varepsilon(G)=\varepsilon(\bar{G})$. But $E(\bar{G})=\binom{V(G)}{2} \backslash E(G)$, so that $\varepsilon(\bar{G})=\binom{|G|}{2}-\varepsilon(G)=\varepsilon(G)$, so $\varepsilon(G)=\frac{1}{2}\binom{|G|}{2}=|G|(|G|-1) / 4$. Now $|G|$ and $|G|-1$ are consecutive integers, so that one of them is odd. Therefore $|G| \equiv 0(\bmod 4)$ or $|G| \equiv 1(\bmod 4)$.
So possible orders for self-complementary graphs are $4,5,8,9,12,13, \ldots, 4 k, 4 k+1, \ldots$

## Exercises

1.1.1 The four graphs on three vertices in Figure 1.9 have $0,1,2$, and 3 edges, respectively. Every graph on three vertices is isomorphic to one of these four. Thus, there are exactly four different isomorphism types of graph on three vertices.
Find all the different isomorphism types of graph on 4 vertices (there are 11 of them). Hint: Adding an edge to a graph with $\varepsilon=m$, gives a graph with $\varepsilon=m+1$. Every graph with $\varepsilon=m+1$ can be obtained in this way. Table 1.1 shows the number of isomorphism types of graph up to 10 vertices.
$G_{1}$


FIGURE 1.9
Four graphs on three vertices TABLE 1.1
Graphs up to $\mathbf{1 0}$ vertices

| $n$ | No. graphs |
| :---: | :---: |
| 1 | 1 |
| 2 | 2 |
| 3 | 4 |
| 4 | 11 |
| 5 | 34 |
| 6 | 156 |
| 7 | 1,044 |
| 8 | 12,346 |
| 9 | 247,688 |
| 10 | $12,005,188$ |

1.1.2 Determine whether the two graphs shown in Figure 1.10 are isomorphic to each other or not. If they are isomorphic, find an explicit isomorphism.


FI GURE 1.10

## Two graphs on eight vertices

1.1.3 Determine whether the three graphs shown in Figure 1.11 are isomorphic to each other or not. If they are isomorphic, find explicit isomorphisms.
1.1.4 Find a self-complementary graph on four vertices.
1.1.5 Figure 1.6 illustrates a self-complementary graph, the pentagon, with five vertices. Find another selfcomplementary graph on five vertices.
1.1.6 We have seen that the pentagon is a self-complementary graph. Let $G$ be the pentagon shown in Figure 1.6 , with $V(G)=\{u 1, u 2, u 3, u 4, u 5\}$. Notice that

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## FI GURE 1.11

## Three graphs on 10 vertices

$\theta=(u 1)(u 2, u 3, u 5, u 4)$ is a permutation which maps $G$ to $\bar{G}$; that is, $\theta(G)=\bar{G}$, and $\theta(\bar{G})=G \cdot \theta$ is called a complementing permutation. Since $u 2 u 3 \in E(G)$, it follows that $\theta\left(u_{2} u_{3}\right)=u_{3} u_{5} \in E(\bar{G})$. Consequently, $\theta\left(u_{3} u_{5}\right)=u_{5} u_{4} \in E(G)$ again. Applying $\theta$ twice more gives $\theta\left(u_{5} u_{4}\right)=u_{4} u_{2} \in E(\bar{G})$ and $\theta(u 4 u 2)=u 2 u 3$, which is where we started. Thus, if we choose any edge uiuj and successively apply $\theta$ to it, we alternately get edges of $G$ and $\bar{G}$. It follows that the number of edges in the sequence so-obtained must be even. Use the permutation ( $1,2,3,4$ ) ( $5,6,7,8$ ) to construct a self-complementary graph on eight vertices.
1.1.7 Can the permutation $(1,2,3,4,5)(6,7,8)$ be used as a complementing permutation? Can ( $1,2,3,4$, $5,6)(7,8)$ be? Prove that the only requirement is that every sequence of edges obtained by successively applying $\theta$ be of even length.
1.1.8 If $\theta$ is any permutation of $\{1,2, \ldots, n\}$, then it depends only on the cycle structure of $\theta$ whether it can be used as a complementing permutation. Discover what condition this cycle structure must satisfy, and prove it both necessary and sufficient for $\theta$ to be a complementing permutation.

### 1.2 Degree sequences

THEOREM 1.2 For any simple graph $G$ we have

$$
\sum_{u \in V(G)} \operatorname{DEG}(u)=2 \varepsilon(G) .
$$

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## Page 9

PROOF An edge $u u$ has two endpoints. Therefore each edge will be counted twice in the summation, once for $u$ and once for $u$.
We use $\delta(G)$ to denote the minimum degree of $G$; that is, $\delta(G)=\operatorname{Min}\{\operatorname{DEG}(u) \mid u \in V(G)\}$. $\Delta(G)$ denotes the maximum degree of $G$. By Theorem 1.2, the average degree equals $2 \varepsilon /|G|$, so that $\delta \leq 2 \varepsilon /|G| \leq \Delta$.
COROLLARY 1.3 The number of vertices of odd degree is even.
PROOF Divide $V(G)$ into Vodd=\{ulDEG(u) is odd\}, and Veuen=\{ul deg $(u)$ is even\}. Then $2 \varepsilon=\sum_{u \in V_{\text {odd }}} \operatorname{DEG}(u)+\sum_{u \in V_{\text {even }}} \operatorname{DEG}(u)$. Clearly $2 \varepsilon$ and $\sum_{u \in V_{\text {even }}} \operatorname{DEG}(u)$ are both even. Therefore, so is $\sum_{u \in V_{\text {odd }}} \operatorname{DeG}(u)$, which means that |Vodd| is even.
DEFINITION 1.5: A graph $G$ is a regular graph if all vertices have the same degree. $G$ is $k$-regular if it is regular, of degree $k$.
For example, the graph of the cube (Figure 1.1) is 3 -regular.
LEMMA 1.4 If $G$ is simple and $|G| \geq 2$, then there are always two vertices of the same degree.
PROOF In a simple graph, the maximum degree $\Delta \leq|G|-1$. If all degrees were different, then they would be
$0,1,2, \ldots,|G|-1$. But degree 0 and degree $|G|-1$ are mutually exclusive. Therefore there must be two
vertices of the same degree.
Let $V(G)=\{u 1, u 2, \ldots, u n\}$. The degree sequence of $G$ is
DEG(G)=(DEG(u1), DEG(u2),..., DEG(un))
where the vertices are ordered so that
DEG $(u 1) \geq$ DEG $(u 2) \geq \ldots \geq$ DEG $(u n)$.
Sometimes it's useful to construct a graph with a given degree sequence. For example, can there be a simple graph with five vertices whose degrees are (4, 3, 3, 2, 1)? Since there are three vertices of odd degree,
Corollary 1.3 tells us that there is no such graph. We say that a sequence
$D=(d 1, d 2, \ldots, d n)$,
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is graphic if
$d 1 \geq d 2 \geq \ldots \geq d n$,
and there is a simple graph $G$ with $\operatorname{DEG}(G)=D$. So $(2,2,2,1)$ and $(4,3,3,2,1)$ are not graphic, whereas ( 2 , $2,1,1),(4,3,2,2,1)$ and $(2,2,2,2,2,2,2)$ clearly are.

Problem 1.1: Graphic
Instance: a sequence $D=(d 1, d 2, \ldots, d n)$.
Question:is $D$ graphic?
Find: a graph $G$ with $\operatorname{DEG}(G)=D$, if $D$ is graphic.
For example, $(7,6,5,4,3,3,2)$ is not graphic; for any graph $G$ with this degree sequence has $\Delta(G)=|G|=7$,
which is not possible in a simple graph. Similarly, $(6,6,5,4,3,3,1)$ is not graphic; here we have $\Delta(G)=6$, $|G|=7$ and $\delta(G)=1$. But since two vertices have degree $|G|-1=6$, it is not possible to have a vertex of degree one in a simple graph with this degree sequence.
When is a sequence graphic? We want a construction which will find a graph $G$ with $D E G(G)=D$, if the sequence $D$ is graphic.
One way is to join up vertices arbitrarily. This does not always work, since we can get stuck, even if the sequence is graphic. The following algorithm always produces a graph $G$ with $D E G(G)=D$, if $D$ is graphic.
procedure GRAPHGEN(D)
Create vertices $u 1, u 2, \ldots$, un
comment: upon completion, ui will have degree $D[i]$
graphic $\leftarrow$ false "assume not graphic"
$i \leftarrow 1$
while $D[i]>0$
do $\left\{\begin{array}{l}\mathrm{k} \leftarrow \mathrm{D}[\mathrm{i}] \\ \text { if there are at least } k \text { vertices with DEG }>0\end{array} \quad \begin{array}{l}\text { then }\left\{\begin{array}{l}\text { join } u_{i} \text { to the } k \text { vertices of largest degree } \\ \text { decrease each of these degrees by } 1 \\ D[i] \leftarrow 0 \\ \text { comment: vertex } u_{i} \text { is now completely joined }\end{array}\right. \\ \text { else exit " } u_{i} \text { cannot be joined" } \\ i \leftarrow i+1\end{array}\right.$
graphic $\leftarrow$ true
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This uses a reduction. For example, given the sequence

$$
D=(3,3,3,3,3,3) \text {, }
$$

the first vertex will be joined to the three vertices of largest degree, which will then reduce the sequence to (*, 3, 3, 2, 2, 2), since the vertex marked by an asterisk is now completely joined, and three others have had their degree reduced by 1. At the next stage, the first remaining vertex will be joined to the three vertices of largest degree, giving a new sequence (*, *, 2, 2, 1, 1). Two vertices are now completely joined. At the next step, the first remaining vertex will be joined to two vertices, leaving (*, *, *, 1, 1, 0). The next step joins the two remaining vertices with degree one, leaving a sequence $(*, *, *, *, 0,0)$ of zeroes, which we know to be graphic.
In general, given the sequence
where

$$
\begin{gathered}
D=(d 1, d 2, \ldots, d n) \\
d 1 \geq d 2 \geq \ldots \geq d n
\end{gathered}
$$

the vertex of degree $d 1$ is joined to the $d 1$ vertices of largest degree. This leaves the numbers

$$
d_{2}-1, d_{3}-1, \ldots, d_{d_{1}+1}-1, d_{d_{1}+2}, \ldots, d_{n}
$$

in some order. If we rearrange them into descending order, we get the reduced sequence $D^{\prime}$. Write

$$
D^{\prime}=\left(d_{2}^{\prime}, d_{3}^{\prime} \ldots, d_{n}^{\prime}\right)
$$

where the first vertex $u 1$ has been deleted. We now do the same calculation, using $D^{\prime}$ in place of $D$.
Eventually, after joining all the vertices according to their degree, we either get a graph $G$ with $\operatorname{Deg}(G)=D$ or else at some stage, it is impossible to join some vertex ui.
An excellent data structure for representing the graph $G$ for this problem is to have an adjacency list for each vertex $v \in V(G)$. The adjacency list for a vertex $v \in V(G)$ is a linked list of the vertices adjacent to $u$. Thus it is a data structure in which the vertices adjacent to $u$ are arranged in a linear order. A node $x$ in a linked list has two fields: data $\langle x\rangle$, and next $\langle x\rangle$.


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Given a node $x$ in the list, data $\langle x\rangle$ is the data associated with $x$ and next $\langle x\rangle$ points to the successor of $x$ in the list or $\operatorname{next}\langle x\rangle=$ NIL if $x$ has no successor. We can insert data $u$ into the list pointed to by $L$ with procedure LISTINSERT(), and the first node on list $L$ can be removed with procedure LISTREMOVEFIRST().
procedure LISITINSERT(L, u)
$x \leftarrow$ NEWNODE()
data $\langle x\rangle \leftarrow u$
$\operatorname{next}\langle x\rangle \leftarrow L$
$L \leftarrow x$
procedure LISTREMOVEFIRST(L)
$x \leftarrow L$
$L \leftarrow \operatorname{next}\langle x\rangle$
FREENODE ( $x$ )
We use an array AdjList[•] of linked lists to store the graph. For each vertex $v \in V(G), \operatorname{AdjList[v]~points~to~}$ the head of the adjacency lists for $v$. This data structure is illustrated in Figure 1.12.


FI GURE 1.12
Adjacency lists of a graph
We can use another array of linked lists, Pts[k], being a linked list of the vertices ui whose degree-to-be $d i=k$. With this data structure, Algorithm 1.2.1 can be written as follows:

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Algorithm 1.2.1: GRAPHGEN(D)
comment: $\left\{\begin{array}{l}\text { Assume } D \text { is not graphic. } \\ \text { Create and initialize the linked lists Pts }[k] .\end{array}\right.$
graphic $\leftarrow$ false
for $k \leftarrow 0$ to $n-1$ do Pts $[k] \leftarrow \mathrm{NIL}$ for $k \leftarrow 1$ to $n$ do LISTINSERT(Pts[D[k]], $k$ )
comment: Begin with vertex of largest degree.
for $k \leftarrow n$ downto 1
do while Pts $[k] \neq$ NIL
comment: These points are to have degree $k$.
$x \leftarrow P t s[k]$
$u \leftarrow \operatorname{data}\langle x\rangle$
LISTREMOVEFIRST(Pts [ $k]$ )
comment: $\left\{\begin{array}{l}\text { Join } u \text { to the next } k \text { vertices } v \text { of largest degree. } \\ \text { If this is not possible, then } \mathrm{D} \text { is not graphic so exit. }\end{array}\right.$
$i \leftarrow k$
for $j \leftarrow 1$ to $k$
$\left\{\right.$ while Pts $[i]=$ NIL do $\left\{\begin{array}{l}i \leftarrow i-1 \\ \text { if } i=0 \text { exit }\end{array}\right.$
$x=P t s[i]$
do $\left\{\begin{array}{l}v=\text { data }\langle x\rangle \\ \text { LISTREMOV }\end{array}\right.$
ListRemoveFirst(Pts[i])
ListInsert(AdjList [u], v)
Listinsert(AdjList $[v], u$ )
ListInsert(TempList $[i], v$ )
comment: $\left\{\begin{array}{l}\text { For each such } \mathrm{v} \text { joined to } \mathrm{u} \text { if } v \text { is on list Pts }[j], \\ \text { then transfer } v \text { to Pts }[j-1]\end{array}\right.$
for $j \leftarrow k$ downto 1
do $\left\{\begin{array}{l}\text { while TempList }[j] \neq \text { NIL } \\ \text { do }\left\{\begin{array}{l}x=\text { TempList }[j] \\ v=\operatorname{data}\langle x\rangle \\ \operatorname{LiSTREMOVEFIRST}(\text { TempList }[j]) \\ \operatorname{LiSTINSERT}(\text { Pts }[j-1], v)\end{array}\right.\end{array}\right.$
comment: $u$ is now completely joined. Choose the next point.
comment: Now every vertex has been successfully joined.
graphic $\leftarrow$ true
This program is illustrated in Figure 1.13 for the sequence $D=(4,4,2,2,2,2)$, where $n=6$. The diagram shows the linked lists before vertex 1 is joined to vertices $2,3,4$, and 5 , and the new configuration after joining. Care must be
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| Pts $[6] \rightarrow$ | $D=(4,4,2,2,2,2)$ |  |  |
| :--- | :--- | :--- | :--- |
| Pts $[5] \rightarrow$ |  | 1 <br> $\rightarrow$ | ${ }^{5} \bullet$ |
|  |  | $\bullet 2$ |  |

$\operatorname{Pts}[4] \rightarrow \sqrt{1} \rightarrow \sqrt{2} \rightarrow$
$\operatorname{Pts}[3] \rightarrow$
$\operatorname{Pts}[2] \rightarrow \sqrt[3]{3} \rightarrow \sqrt[4]{4} \rightarrow \sqrt{5} \rightarrow$
$\operatorname{Pts}[1] \rightarrow$
$\operatorname{Pts}[0] \rightarrow$
(a)

(b)

## FI GURE 1.13

The linked lists Pts[k]. (a) Before 1 is joined to 2, 3, 4, and 5. (b) After 1 is joined to 2, 3, 4, and 5.
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used in transferring the vertices $u$ from Pts[j] to Pts[j-1], since we do not want to join $u$ to $u$ more than once. The purpose of the list Pts[0] is to collect vertices which have been transferred from Pts[1] after having been joined to $u$. The degrees $d 1, d 2, \ldots, d n$ need not necessarily be in descending order for the program to work, since the points are placed in the lists Pts[ $k$ ] according to their degree, thereby sorting them into buckets. Upon completion of the algorithm vertex $k$ will have degree $d k$. However, when this algorithm is done by hand, it is much more convenient to begin with a sorted list of degrees; for example, $D=(4,3,3,3,2,2$, $2,2,1$ ), where $n=9$. We begin with vertex $u 1$, which is to have degree four. It will be joined to the vertices $u 2, u 3$, and $u 4$, all of degree three, and to one of $u 5, u 6, u 7$, and $u 8$, which have degree two. In order to keep the list of degrees sorted, we choose $u 8$. We then have $u_{1} \longrightarrow\left\{u_{2}, u_{3}, u_{4}, u_{8}\right\}$, and $D$ is reduced to $\left(^{*}\right.$, $2,2,2,2,2,2,1,1$ ). We then choose $u 2$ and join it to $u 6$ and $u 7$, thereby further reducing $D$ to (*, *, 2, 2, $2,2,1,1,1,1)$. Continuing in this way, we obtain a graph $G$.
In general, when constructing $G$ by hand, when $u k$ is to be joined to one of $u i$ and $u j$, where $d i=d j$ and $i<j$, then join $u k$ to $u j$ before $u i$, in order to keep $D$ sorted in descending order.
We still need to prove that Algorithm 1.2.1 works. It accepts a possible degree sequence $D=(d 1, d 2, \ldots . . d n)$,
and joins $u 1$ to the $d 1$ vertices of largest remaining degree. It then reduces $D$ to new sequence

$$
D^{\prime}=\left(d_{2}^{\prime}, d_{3}^{\prime}, \ldots d_{n}^{\prime}\right) .
$$

THEOREM 1.5 (Havel-Hakimi theorem) $D$ is graphic if and only if $D^{\prime}$ is graphic.

PROOF Suppose $D^{\prime}$ is graphic. Then there is a graph $G^{\prime}$ with degree sequence $D^{\prime}$, where $V\left(G^{\prime}\right)=\{u 2, u 3, \ldots$, $u n\}$ with $\operatorname{DEG}\left(u_{i}\right)=d_{i}^{\prime}$. Furthermore

$$
D^{\prime}=\left(d_{2}^{\prime}, d_{3}^{\prime}, \ldots, d_{n}^{\prime}\right)
$$

consists of the degrees

$$
\left\{d_{2}-1, d_{3}-1, \ldots, d_{d_{1}+1}-1, d_{d_{1}+2}, \ldots, d_{n}\right\}
$$

arranged in descending order. Create a new vertex $u 1$ and ioin it to vertices of degree $d_{2}-1, d_{3}-1, \ldots, d_{d_{1}+1}-1$.
Then $\operatorname{DEG}(u 1)=d 1$. Call the new graph $G$. Clearly the degree sequence of $G$ is $D=(d 1, d 2, \ldots, d n)$.

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FI GURE 1.14
Vertices adjacent to $u \mathbf{l}$
Therefore $D$ is graphic.
Now suppose $D$ is graphic. Then there is a graph $G$ with degree sequence

$$
D=(d 1, d 2, \ldots, d n),
$$

where $V(G)=\{u 1, u 2, \ldots, u n\}$, with $\operatorname{DEG}(u i)=d i$. If $u 1$ is adjacent to vertices of degree $d_{2}, d_{3}, \ldots, d_{d_{1}+1}$, then $G$ $'=G-u 1$ has degree sequence $D^{\prime}$, in which case $D^{\prime}$ is qraphic.
Otherwise, $u 1$ is not adjacent to vertices of degree $d_{2}, d_{3}, \ldots, d_{d_{1}+1}$. Let $u k$ (where $k \geq 2$ ) be the first vertex such that $u 1$ is not joined to $u k$, but is joined to $u 2, u 3, \ldots, u k-1$. (Maybe $k=2$.)
Now $\operatorname{DEG}(u 1)=d 1 \geq k$, so $u 1$ is joined to some vertex $x \neq u 2, u 3, \ldots u k-1$. $u k$ is the vertex of next largest degree, so DEG $(u k) \geq \operatorname{DEG}(x)$. Now $x$ is joined to $u 1$, while $u k$ is not. Therefore, there is some vertex $y$ such that $u_{k} \longrightarrow y$ but $x \nrightarrow y$. Set $G \leftarrow G+x y+u 1 u k-u 1 x-u k y$.
The degree sequence of $G$ has not changed, and now $u_{1} \longrightarrow\left\{u_{2}, u_{3}, \ldots, u_{k}\right\}$. Repeat until $u_{1} \longrightarrow\left\{u_{2}, u_{3}, \ldots, u_{d_{1}+1}\right\}$. Then $G^{\prime}=G-u 1$ has degree sequence $D^{\prime}$, so that $D^{\prime}$ is graphic.
Therefore we know the algorithm will terminate with the correct answer, because it reduces $D$ to $D^{\prime}$. So we have an algorithmic test to check whether $D$ is graphic and to generate a graph whenever one exists. There is another way of determining whether $D$ is graphic, without constructing a graph.
THEOREM 1.6 (Erdós-Gallai theorem) Let $D=(d 1, d 2, \ldots, d n)$, where $d 1 \geq$
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$d 2 \geq \ldots \geq d n$. Then $D$ is graphic if and only if

1. $\sum_{i=1}^{n} d_{i}$ is even; and
2. $\sum_{i=1}^{k} d_{i} \leq k(k-1)+\sum_{i=k+1}^{n} \operatorname{MiN}\left(k, d_{i}\right)$, for $k=1,2, \ldots, n$.

PROOF Suppose $D$ is graphic. Then $\sum_{i=1}^{n} d_{i}=2 \varepsilon$, which is even. Let $V 1$ contain the $k$ vertices of largest degree, and let $V 2=V(G)-V 1$ be the remaining vertices. See Figure 1.15.

## $\varepsilon_{1}$ <br> $V_{1}$ <br> $V_{2}$

## FI GURE 1.15

## The vertices V 1 of largest degree and the remaining vertices $u 2$

Suppose that there are $\varepsilon 1$ edges within $V 1$ and $\varepsilon 2$ edges from $V 1$ to $V 2$. Then $\sum_{i=1}^{k} d_{i}=2 \varepsilon_{1}+\varepsilon_{2}$, since each edge within $V 1$ is counted twice in the sum, once for each endpoint, but edges between $V 1$ and $V 2$ are counted once only. Now $\varepsilon_{1} \leq\binom{ k}{2}$, since $V 1$ can induce a complete subgraph at most. Each vertex $v \in V_{2}$ can be joined to at most $k$ vertices in $V 1$, since $|V 1|=k$, but $u$ can be joined to at most $D E G(u)$ vertices in $V 1$, if $\operatorname{DEG}(u)<k$. Therefore $\varepsilon 2$, the number of edges between $V 1$ and $V 2$, is at most $\sum_{v \in V_{2}} \operatorname{MIN}(k, \operatorname{DEG}(v))$, which equals $\sum_{i=k+1}^{n} \operatorname{MIN}\left(k, d_{i}\right)$. This now gives $\sum_{i=1}^{k} d_{i}=2 \varepsilon_{1}+\varepsilon_{2} \leq k(k-1)+\sum_{i=k+1}^{n} \operatorname{MIN}\left(k, d_{i}\right)$ The proof of the converse is quite long, and is not included here. A proof by induction can be found in the book by HARARY [59].
Conditions 1 and 2 of the above theorem are known as the Erdös-Gallai conditions.

## Exercises

1.2.1 Prove Theorem 1.2 for arbitrary graphs. That is, prove
page_17
Page 18
THEOREM 1.7 For any graph $G$ we have

$$
\sum_{u \in V(G)} \operatorname{Deg}(u)+\ell=2 \varepsilon(G)
$$

where $\ell$ is the number of loops in $G$ and $\operatorname{DEG}(u)$ is the number of edges incident on $u$. What formula is obtained if loops count two toward DEG(u)?
1.2.2 If $G$ has degree sequence $D=(d 1, d 2, \ldots, d n)$, what is the degree sequence of $\bar{G}$ ?
1.2.3 We know that a simple graph with $n$ vertices has at least one pair of vertices of equal degree, if $n \geq 2$.

Find all simple graphs with exactly one pair of vertices with equal degrees. What are their degree sequences?
Hint: Begin with $n=2,3,4$. Use a recursive construction. Can degree 0 or $n-1$ occur twice?
1.2.4 Program the GRAPHGEN() algorithm. Input the sequence $D=(d 1, d 2, \ldots, d n)$ and then construct a graph with that degree sequence, or else determine that the sequence is not graphic. Use the following input data:
(a) 44444
(b) 333333
(c) 33333333
(d) 333333333
(e) 2222222222
(f) 76665521
1.2.5 Let $D=(d 1, d 2, \ldots, d n)$, where $d 1 \geq d 2 \geq \ldots \geq d n$. Prove that there is a multigraph with degree sequence $D$ if and only if $\sum_{i=1}^{n} d_{i}$ is even, and $d 1 \leq \sum_{i=2}^{n} d_{i}$.

### 1.3 Analysis

Let us estimate the number of steps that Algorithm 1.2.1 performs. Consider the loop structure
for $k \leftarrow n$ downto 1
do while Pts[k] $=$ NI L
do $\{\ldots$
The for-loop performs $n$ iterations. For many of these iterations, the contents of the while-loop will not be executed, since Pts $[k]$ will be NIL. When the contents of the loop are executed, vertex $u$ of degree-to-be $k$ will be joined to $k$ vertices. This means that $k$ edges will be added to the adjacency lists of the graph $G$ being constructed. This takes $2 k$ steps, since an edge $u u$ must be added to both GraphAdj[u] and GraphAdj[u]. It also makes $D E G(u)=k$. When edge $u v$ is added, $u$ will be transferred from Pts[ [] to Pts[ $i-1]$, requiring additional $k$ steps. Once $u$ has been joined, it is removed from the list. Write $\varepsilon=\frac{1}{2} \sum_{i} d_{i}$,

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the number of edges of $G$ when $D$ is graphic. Then, in all, the combination for-while-loop will perform exactly $2 \varepsilon$ steps adding edges to the graph and a further $\varepsilon$ steps transferring vertices to other lists, plus $n$ steps for the $n$ iterations of the for-loop. This gives a total of $3 \varepsilon+n$ steps for the for-while-loop. The other work that the algorithm performs is to create and initialize the lists Pts[•], which takes $2 n$ steps altogether. So we can say that in total, the algorithm performs $3 \varepsilon+3 n$ steps.
Now it is obvious that each of these "steps" is composed of many other smaller steps, for there are various comparisons and assignments in the algorithm which we have not explicitly taken account of (they are
subsumed into the steps we have explicitly counted). Furthermore, when compiled into assembly language, each step will be replaced by many smaller steps. Assembly language is in turn executed by the
microprogramming of a computer, and eventually we come down to logic gates, flip-flops, and registers.
Because of this fact, and because each computer has its own architecture and machine characteristics, it is customary to ignore the constant coefficients of the graph parameters $\varepsilon$ and $n$, and to say that the algorithm has order $\varepsilon+n$, which is denoted by $O(\varepsilon+n)$, pronounced "big Oh of $\varepsilon+n$ ". A formal definition is provided by Definition 1.6. Even though the actual running time of a given algorithm depends on the architecture of the machine it is run on, the programmer can often make a reasonable estimate of the number of steps of some constant size (e.g., counting one assignment, comparison, addition, multiplication, etc. as one step), and thereby obtain a formula like $3 \varepsilon+3 n$. Such an algorithm will obviously be superior to one which takes
$15 \varepsilon+12 n$ steps of similar size. Because of this fact, we shall try to obtain formulas of this form whenever possible, as well as expressing the result in a form like $O(\varepsilon+n)$.
The complexity of an algorithm is the number of steps it must perform, in the worst possible case. That is, it is an upper bound on the number of steps. Since the size of each step is an unknown constant, formulas like $5 n 2 / 6$ and $25 n 2$ are both expressed as $O(n 2)$. We now give a formal definition of this notation.
DEFINITION 1.6: Suppose $f: \mathbb{Z}^{+} \rightarrow \mathbb{R}$ and $g: \mathbb{Z}^{+} \rightarrow \mathbb{R}$. We say that $f(n)$ is $O(g(n))$ provided that there exist constants $c>0$ and $n 0 \geq 0$ such that $0 \leq f(n) \leq c \cdot g(n)$ for all $n \geq n 0$.
In other words, $f(n)$ is $O(g(n))$ provided that $f(n)$ is bounded above by a constant factor times $g(n)$ for large enough $n$. For example, the function $5 n 3+2 n+1$ is $O(n 3)$, because for all $n \geq 1$, we have
$5 n 3+2 n+1 \leq 5 n 3+2 n 3+n 3=8 n 3$.
Hence, we can take $c=8$ and $n 0=1$, and Definition 1.6 is satisfied.
The notation $f(n)$ is $\Omega(g(n)$ ) ("big omega") is used to indicate that $f(n)$ is bounded below by a constant factor times $g(n)$ for large enough $n$.
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DEFINITION 1.7: Suppose $f: \mathbb{Z}^{+} \rightarrow \mathbb{R}$ and $g: \mathbb{Z}^{+} \rightarrow \mathbb{R}$. We say that $f(n)$ is $\Omega(g(n))$ provided that there exist constants $c>0$ and $n 0 \geq 0$ such that $f(n) \geq c \cdot g(n) \geq 0$ for all $n \geq n 0$.
We say that $f(n)$ is $\Theta(g(n))$ ("big theta") when $f(n)$ is bounded above and below by constant factors times $g(n)$. The constant factors may be different. More precisely:
DEFINITION 1.8: Suppose $f: \mathbb{Z}^{+} \rightarrow \mathbb{R}$ and $g: \mathbb{Z}^{+} \rightarrow \mathbb{R}$. We say that $f(n)$ is $\Theta(g(n))$ provided that there exist constants $c, c^{\prime}>0$ and $n 0 \geq 0$ such that $0 \leq c \cdot g(n) \leq f(n) \leq c^{\prime} \cdot g(n)$ for all $n \geq n 0$.
If $f(n)$ is $\Theta(g(n))$, then we say that $f$ and $g$ have the same growth rate.
The big $O$-notation is a method of indicating the qualitative nature of the formula, whether quadratic, linear, logarithmic, exponential, etc. Notice that "equations" involving $O(\cdot)$ are not really equations, since $O(\cdot)$ can only be used in this sense on the right hand side of the equals sign. For example, we could also have shown that $10 n 2+4 n-4$ is $O(n 3)$ or that $10 n 2+4 n-4$ is $O(2 n)$, but these expressions are not equal to each other. Given a complexity formula like $10 n 2+4 n-4$, we want the smallest function $f(n)$ such that $10 n 2+4 n-4$ is $O(f(n))$. Among the useful rules for working with the $O$-notation are the following sum and product rules. THEOREM 1.8 Suppose that the two functions $f 1(n)$ and $f 2(n)$ are both $O(g(n))$. Then the function $f 1(n)+f 2(n)$ is $O(g(n))$.
THEOREM 1.9 Suppose that $f 1(n)$ is $O(g 1(n))$ and $f 2(n)$ is $O(g 2(n))$. Then the function $f 1(n) f 2(n)$ is $O(g 1(n)$ $g 1(n))$.
As examples of the use of these notations, we have that $n 2$ is $O(n 3), n 3$ is $\Omega(n 2)$, and $2 n 2+3 n-\sin n+1 / n$ is $\Theta(n 2)$.
Several properties of growth rates of functions that arise frequently in algorithm analysis follow. The first of these says that a polynomial of degree $d$, in which the high-order coefficient is positive, has growth rate nd. THEOREM 1.10 Suppose that $a d>0$. Then the function $a 0+a 1 n+\ldots+$ adndis $\Theta$ (nd).
The next result says that logarithmic growth does not depend on the base to which logarithms are computed.

It can be proved easily using the formula loga $n=\operatorname{loga} b \cdot \log b n$.
THEOREM 1.11 The function logan is $\Theta$ (logbn) for any $a, b>1$.
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The next result can be proved using Stirling's formula. It gives the growth rate of the factorial function in terms of exponential functions.
THEOREM 1.12 The function $n!$ is $\Theta\left(n^{n+1 / 2} e^{-n}\right)$.

## Exercises

1.3.1 Show that if $G$ is a simple graph with $n$ vertices and $\varepsilon$ edges, then $\log \varepsilon=O(\log n)$.
1.3.2 Consider the following statements which count the number of edges in a graph, whose adjacency matrix is $A d j$.

```
                                    Edges \(\leftarrow 0\)
                        for \(u \leftarrow 1\) to \(n-1\)
do for \(u \leftarrow u+1\) to \(n\)
    do if \(\operatorname{Adj}[u, u]=1\)
then Edges \(\leftarrow\) Edges +1
```

Calculate the number of steps the algorithm performs. Then calculate the number of steps required by the following statements in which the graph is stored in adjacency lists:
Edges $\leftarrow 0$
for $u \leftarrow 1$ to $n-1$
do for each $v \longrightarrow u$
do if $u<u$
then Edges $\leftarrow$ Edges +1
What purpose does the condition $u<u$ fulfill, and how can it be avoided?
1.3.3 Use induction to prove that the following formulas hold:
(a) $1+2+3+\cdots+n=\binom{n+1}{2}$
(b) $\binom{2}{2}+\binom{3}{2}+\binom{4}{2}+\cdots+\binom{n}{2}=\binom{n+1}{3}$.
(c) $\binom{t}{t}+\binom{t+1}{t}+\binom{t+2}{t}+\cdots+\binom{n}{t}=\binom{n+1}{t+1}$.
1.3.4 Show that $3 n 2+12 n=O(n 2)$; that is, find constants $A$ and $N$ such that $3 n 2+12 n<A n 2$ whenever $n \geq N$.
1.3.5 Show that $\log (n+1)=O(\log n)$, where the logarithm is to base 2.
1.3.6 Use the answer to the previous question to prove that $(n+1) \log (n+1)=O(n \log n)$.
1.3.7 Prove that if $f 1(n)$ and $f 2(n)$ are both $O(g(n))$, then $f 1(n)+f 2(n)$ is $O(g(n))$.
1.3.8 Prove that if $f 1(n)$ is $O(g 1(n))$ and $f 2(n)$ is $O(g 2(n))$, then $f 1(n) f 2(n)$ is $O(g 1(n) g 1(n))$.
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### 1.4 Notes

Some good general books on graph theory are BERGE [14], BOLLOBÁS [18], BONDY and MURTY [19], CHARTRAND and LESNIAK [24], CHARTRAND and OELLERMANN [25], DIESTEL [35], GOULD [53], and WEST [123]. A very readable introductory book is TRUDEAU [115]. GIBBONS [51] is an excellent treatment of graph algorithms. A good book discussing the analysis of algorithms is PuRDOM and BROWN [96]. AHO, HOPCROFT, and ULLMAN [1], SEDGEWICK [108] and WEISS [122] are all excellent treatments of data structures and algorithm analysis.
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2
Paths and Walks

### 2.1 I ntroduction

Let $u$ and $v$ be vertices of a simple graph $G$. A path $P$ from $u$ to $v$ is a sequence of vertices $u 0, u 1, \ldots, u k$ such that $u=u 0, v=u k, u_{i} \longrightarrow u_{i+1}$, and all the ui are distinct vertices. The length of a path $P$ is $l(P)$, the number of edges it uses. In this example, $I(P)=k$, and $P$ is called a $u v$-path of length $k$. A $u v$-path of length 4 is illustrated in Figure 2.1, with dashed edges.
A cycle $C$ is a sequence of vertices $u 0, u 1, \ldots, u k$ forming a $u 0 u k$-path, such that $u_{k} \longrightarrow u_{0}$. The length of $C$ is $I(C)$, the number of edges that it uses. In this case, $I(C)=k+1$.

A uv-path $P$ connects $u$ to $v$. The set of all vertices connected to any vertex $u$ forms a subgraph $C u$, the connected component of $G$ containing $u$. It will often be the case that $C u$ contains all of $G$, in which case $G$ is a connected graph. $w(G)$ denotes the number of distinct connected components of $G$. The graph of Figure 2.1 is disconnected, with $w=3$.
There are several ways of finding the connected components of a graph G. One way to find the sets Cu for a graph $G$ is as follows:
procedure COMPONENTS(G)
for each $u \in V(G)$
do initialize Cu to contain only $u$
for each $u \in V(G)$
do $\left\{\begin{array}{l}\text { for each } v \longrightarrow u \\ \text { do if } C_{u} \neq C_{v}\end{array}\right.$ then $\operatorname{Merge}\left(C_{u}, C_{v}\right)$
The inner for-loop ensures that, upon completion, if $u \longrightarrow v$, then $C u=C v$, for any vertices $u$ and ?. Therefore, if $P=(u 0, u 1, \ldots, u k)$ is any path, we can
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## FI GURE 2.1

## A graph with three components

be sure that $C_{u_{0}}=C_{u_{1}}=\cdots=C_{u_{k}}$, so that when the algorithm terminates, each $C u$ will contain all the vertices connected to $u$ by any path; that is, Cu will be the connected component containing $u$.
The complexity of the algorithm naturally depends upon the data structures used to program it. This algorithm is a perfect example of the use of the merge-find data structure. Initially, each $C u=\{u\}$ and $C u=\{u\}$. When the edge $u u$ is examined, $C u$ and $C u$ are merged, so that now $C u=C u=\{u, u\}$. The two operations which need to be performed are to determine whether $C u=C u$, and to merge $C u$ and $C u$ into one.
This can be done very efficiently by choosing a vertex in each Cu as component representative.
$u R e p \leftarrow \operatorname{COMPREP}(C u)$
$v R e p \leftarrow C O M P R E P(C u)$
if $u R e p \neq v R e p$
then MERGE(Cu,Cu)
Initially, $C u=\{u\}$, so that $u$ begins as the representative of $C u$. Associated with each vertex $u$ is a pointer toward the representative of the component containing $u$. To find the representative of $C u$, we start at $u$ and follow these pointers, until we come to the component representative. The component representative is marked by a pointer that is negative. The initial value is-1. The pointers are easily stored as an array, CompPtr.
$\operatorname{COMPREP}()$ is a recursive procedure that follows the component pointers until a negative value is reached.
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procedure COMPREP(u)
if CompPtr[u]<0
then return ( $u$ )
else $\left\{\begin{array}{l}\text { theRep } \leftarrow \operatorname{ComPREP}(\text { CompPtr }[u]) \\ \text { CompPtr }[u] \leftarrow \text { theRep } \\ \text { return (theRep) }\end{array}\right.$

The assignment
is called path compression. It ensures that the next time CompPtr( $u$ ) is computed, the representative will be found more quickly. The algorithm COMPONENTS() can now be written as follows:
Algorithm 2.1.1: COMPONENTS(G)
for $u \leftarrow 1$ to $n$
do CompPtr[u] $\leftarrow-1$
for $u \leftarrow 1$ to $n$

$$
\text { do }\left\{\begin{array}{l}
\text { for each } v \longrightarrow u \\
\text { do }\left\{\begin{array}{l}
u R e p \leftarrow \operatorname{ComPREP}(u) \\
v R e p \leftarrow \operatorname{ComPREP}(v) \\
\text { if } u \operatorname{Rep} \neq v R e p \\
\text { then } \operatorname{MERGE}(u R e p, v R e p)
\end{array}\right.
\end{array}\right.
$$

The essential step in merging $C u$ and $C u$ is to assign either
CompPtr[vRep] $\leftarrow u \operatorname{Rep}$
or
CompPtr[uRep] $\leftarrow$ vRep
The best one to choose is that which merges the smaller component onto the larger. We can determine the size of each component by making use of the negative values of CompPtr[uRep] and CompPtr[vRep]. Initially, CompPtr[u]=-1, indicating a component of size one.

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procedure MERGE(uRep,vRep)
uSize $\leftarrow-C o m p P t r[u R e p] ~$
$v$ Size $\leftarrow-$ CompPtr[ vRep]
if uSize<vSize
then $\left\{\begin{array}{l}\text { CompPtr }[u \text { Rep }] \leftarrow v R e p \\ \text { CompPtr }[v R e p] \leftarrow-(u S i z e+v \text { Size })\end{array}\right.$
CompPtr[vRep] $\leftarrow u R e p$
else $\left\{\begin{array}{l}\text { CompPtr }[v R e p] \leftarrow u R e p \\ \text { CompPtr }[u R e p] \leftarrow-(u S i z e ~+v \text { Size })\end{array}\right.$
When Cu and Cv are merged, the new component representative (either uRep or vRep) has its CompPtr••] assigned equal to-(uSize+vSize). The component pointers can be illustrated graphically. They are shown in Figure 2.2 as arrows. The merge operation is indicated by the dashed line.


CompPtr $[u R e p]=-8$

## FI GURE 2.2 <br> Component representatives

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### 2.2 Complexity

The components algorithm is very efficient. The for-loop which initializes the CompPtr array requires $n$ steps. If adjacency lists are used to store $G$, then the total number of times that the body of the main loop is executed is

$$
\sum \operatorname{DEG}(u)=2 \varepsilon
$$

Thus COMPREP() is called $4 \varepsilon$ times. How many times is MERGE() called? At each merge, two existing components are replaced by one, so that at most $n-1$ merges can take place. Each merge can be performed using four assignments and a comparison. It takes $n$ steps to initialize the CompPtr array. Thus the total number of steps is about $6 n+4 \varepsilon$ (number of steps per call to COMPREP()). The number of steps each call to COMPREP() requires depends on the depth of the trees which represent the components. The depth is changed by path compression, and by merging. It is proved in AHO, HOPCROFT, and ULLMAN [1], that if there are a total of $n$ points involved, the number of steps required is $O(a(n))$, where $a(n)$ is the inverse of the function $A(n)$, defined recursively as follows.

$$
A(1)=1
$$

$A(k)=2 A(k-1)$
Thus, $A(2)=21=2, A(3)=22=4, A(4)=24=16, A(5)=216=65536$, etc. It follows that $a(n) \leq 5$, for all $n \leq 65536$. So the complexity of Algorithm 2.1.1 is almost linear, namely, $O(n+\varepsilon a(n))$, where $a(n) \leq 5$, for all practical values of $n$.

## Exercises

2.2.1 Assuming the data structures described above, program the COMPONENTS() algorithm, merging the smaller component onto the larger. Include an integer variable NComps which contains the current number of components. Upon completion, its value will equal $w(G)$.
2.2.2 Algorithm 2.1.1 computes the connected components Cu using the array CompPtr. If we now want to print the vertices of each distinct $C u$, it cannot be done very efficiently. Show how to use linked lists so that for each component, a list of the vertices it contains is available. Rewrite the MERGE() procedure to include this. Is the complexity thereby affected?
2.2.3 In the Algorithm 2.1.1 procedure, the for-loop
for $u \leftarrow 1$ to $n$ do

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executes the statement $u$ Rep $\leftarrow \operatorname{COMPREP}(u)$ once for every $v \longrightarrow u$. Show how to make this more efficient by taking the statement $u$ Rep $\leftarrow \operatorname{COMPREP}(u)$ out of the $u$-loop, and modifying the MERGE() procedure slightly. Calculate the new complexity.
2.2.4 Let $n=|G|$. Show that if $\varepsilon>\binom{n-1}{2}$, then $G$ is connected. Hint: If $G$ is disconnected, there is a component of size $x<n$. What is the maximum number of edges $G$ can then have?
2.2.5 Show that if $\delta>\lfloor(n-1) / 2\rfloor$, then $G$ is connected.
2.2.6 Show that if $G$ is disconnected, then $\bar{G}$ is connected.
2.2.7 Show that if $G$ is simple and connected but not complete, then $G$ has three vertices $u, u$, and $w$ such that $u \longrightarrow v, w$, but $v \nrightarrow w$.
2.2.8 A longest path in a graph $G$ is any path $P$ such that $G$ contains no path longer than $P$. Thus a graph can have several different longest paths (all of the same length, though). Show that $\ell(P) \geq \delta(G)$, for any longest path. Hint: Consider an endpoint of $P$.
2.2.9 Show that every graph $G$ has a cycle of length at least $\delta(G)+1$, if $\delta(G) \geq 2$. Hint: Consider a longest path.
2.2.10 Prove that in a connected graph, any two longest paths have at least one vertex in common.

### 2.3 Walks

Paths do not contain repeated vertices or edges. A walk in $G$ is any sequence of vertices $u 0, u 1, \ldots, u k$ such that $u_{i} \longrightarrow u_{i+1}$. Thus, in a walk, edges and vertices may be repeated. Walks are important because of their connection with the adjacency matrix of a graph. Let $A$ be the adjacency matrix of $G$, where $V(G)=\{u 1, u 2 \ldots$, un\}, such that row and column $i$ of $A$ correspond to vertex ui.
THEOREM 2.1 Entry [i,j] of Akis the number of walks of length $k$ from vertex ui to uj.
PROOF By induction on $k$. When $k=1$, there is a walk of length 1 from ui to uj if and only if $u_{i} \longrightarrow u_{j}$, in which case entry $A[i, j]=1$. Assume it's true whenever $k \leq t$ and consider $A t+1$. Let $W$ be a uiuj-walk of length $t+1$, where $t \geq 2$. If $u l$ is the vertex before $u j$ on $W$, then $W$ can be written as ( $\left.W^{\prime}, u l, u j\right)$, where $W^{\prime}$ is a uiulwalk of length $t$. Furthermore, every uiul-walk of length $t$ gives a uiuj-walk of length $t+1$ whenever $u_{l} \longrightarrow u_{j}$. Therefore the number of
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uiuj-walks of length $t+1$ is

$$
\sum_{l}\left(\text { the number of } u_{i} u_{l}-\text { walks of length } t\right)(A[l, j]) .
$$

But the number of uiul-walks of length $t$ is $A t[i, I]$, so that the number of uiuj-walks of length $t+1$ is

$$
\sum_{l=1}^{n} A^{t}[i, l] A[l, j],
$$

which equals $A t+1[i, j]$. Therefore the result is true when $k=t+1$. By induction, it's true for all values of $k$. Notice that this result is also true for multigraphs, where now $A[i, j]$ is the number of edges joining ui to uj. For multigraphs, a walk $W$ must be specified by giving the sequence of edges traversed, as well as the sequence of vertices, since there can be more than one edge joining the same pair of vertices.

## Exercises

2.3.1 Show that $A 2[i, j]$ equals the number of uiuj-paths of length 2 , if $i \neq j$, and that $A 2[i, i]=\operatorname{DEG}(u i)$.
2.3.2 Show that $A 3[i, i]$ equals the number of triangles containing vertex ui. Find a similar interpretation of A3[i,j], when $i \neq j$. (A triangle is a cycle of length 3 .)
2.3.3 Ak contains the number of walks of length $k$ connecting any two vertices. Multiply $A k$ by $x k$, the $k t h$ power of a variable $x$, and sum over $k$, to get the matrix power series $I+A x+A 2 \times 2+A 3 \times 3+\ldots$, where $I$ is the identity matrix. The sum of this power series is a matrix whose ijth entry is a function of $x$ containing the number of uiuj-walks of each length, as the coefficient of $x k$. Since the power series expansion of $(1-a)-1$ is $1+a+a 2+a 3+\ldots$, , we can write the above matrix as $(I-A x)-1$. That is, the inverse of the matrix $(I-A x)$ is the walk generating matrix. Find the walk generating matrix for the graph of Figure 2.3.

## FI GURE 2.3

Compute the number of walks in this graph.

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### 2.4 The shortest-path problem

The distance from vertex $u$ to $v$ is $\operatorname{DIST}(u, v)$, the length of the shortest $u v$-path. If $G$ contains no $u v$-path, then $\operatorname{DIST}(u, v)=\infty$. In this section we study the following two problems.

## Problem 2.1:

Shortest Path
Instance: a graph $G$ and a vertex $u$.
Find: $\quad \operatorname{DIST}(u, v)$, for all $u, v \in V(G)$.

## Problem 2.2:

All Paths
Instance: a graph $G$.
Find: $\quad \operatorname{DIST}(u, v)$, for all $u, v \in V(G)$.
Given a vertex $u$, one way of computing DIST( $u, v$ ), for all $v$, is to use a breadth-first search (BFS), as is done in procedure BFS().
procedure $\operatorname{BFS}(G, u)$
comment: $\left\{\begin{array}{l}\text { ScanQ is a queue of vertices } \\ \text { dist }[v] \text { will equal } \operatorname{DIST}(u, v), \text { upon completion }\end{array}\right.$
for each $v \in V(G)$
do dist $[V] \leftarrow 8$
dist[u] $\leftarrow 0$
place $u$ on ScanQ
repeat
select $v$ for the head of ScanQ
for each $w \longrightarrow v$
do if $w$ not on ScanQ
then $\left\{\begin{array}{l}\operatorname{add} w \text { to the end of } \operatorname{Scan} Q \\ \operatorname{dist}[w] \leftarrow \operatorname{dist}[v]+1\end{array}\right.$
advance ScanQ
until all of ScanQ has been processed
Procedure BFS() uses a type of data structure called a queue. A queue is an ordered list in which we usually access only the first or the head of the list and new items are only placed at the end or tail of the list. This is similar to one's
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experience of waiting in line at the checkout counter of a store. The person at the head of the line is processed first by the checker and the new customers enter at the end of the line. One of the most convenient ways to store a queue is as an array. For when an algorithm builds a queue on an array, all the vertices visited are on the array when the algorithm completes, ready for input to the next procedure. BFS() works in this way.


FI GURE 2.4

## A breadth-first search

The breadth-first search (BFS) algorithm is a fundamental algorithm in graph theory. It appears in various guises wherever shortest paths are useful (e.g., network flows, matching theory, coset enumeration, etc.). Figure 2.4 shows the result of applying a BFS to the Petersen graph, where the vertices are numbered according to the order in which they were visited by the algorithm, and shaded according to their distance from vertex 1. The thicker edges show the shortest paths found by the algorithm.
Notice that the first vertex on the $S c a n Q$ is $u$, whose dist $[u]=\operatorname{DIST}(u, u)=0$. The next vertices to be placed on the queue will be those adjacent to $u$, that is, those at distance 1 . When they are placed on the queue, their distance will be computed as

$$
\operatorname{dist}[\cdot] \leftarrow \operatorname{dist}[u]+1 .
$$

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So we can say that initially, that is, up to vertices of distance one, vertices are placed on the queue in order of their distance from $u$; and that when each vertex $w$ is placed on $\operatorname{ScanQ}$, dist[$[w]$ is made equal to DIST $(u, w)$. Assume that this is true for all vertices of distance $k$ or less, where $k \geq 1$. Consider when $u$ is chosen as the first vertex of distance $k$ on ScanQ. The for-loop examines all vertices $w \longrightarrow v$. If $w$ on ScanQ already, then there is a $u w$-path of length $\leq k$, and $w$ is ignored. If $w$ is not on $\operatorname{ScanQ,\text {thenDIST}(u,w)>k\text {.The}}$ $u w$-path via $u$ has length $k+1$, so $w$ is added to the queue, and dist[ $w]$ is set equal to dist[ $u]+1=k+1$. Since every vertex at distance $k+1$ is adjacent to a vertex at distance $k$, we can be sure that when all vertices $u$ on ScanQ at distance $k$ have been scanned, all vertices at distance $k+1$ will be on the queue. Thus the assertion that vertices are placed on the queue in order of their distance from $u$, and that when each vertex $w$ is placed on ScanQ, dist $[w]$ is made equal to DIST $(u, w)$, is true up to distance $k+1$. By induction, it is true for all distances.
This proof that the BFS() algorithm works illustrates how difficult and cumbersome it can be to prove that even a simple, intuitively "obvious" algorithm works correctly. Nevertheless, it is important to be able to prove that algorithms work correctly, especially the more difficult algorithms. Writing down a proof for an "obvious" algorithm will often reveal hidden bugs that it contains. This proof also illustrates another feature, namely, proofs that algorithms work tend to use induction, often on the number of iterations of a main loop. The complexity of the BFS() is very easy to calculate. The main operations which are performed are 1. Scan all $w \longrightarrow v$.
2. Select the next $v \in S c a n Q$.
3. Determine whether $w \in S c a n Q$.

The first operation is most efficiently done if $G$ is stored in adjacency lists. We want the second and third operations to take a constant number of steps. We store ScanQ as an integer array, and also store a boolean
array onScanQ to tell whether $w \in \operatorname{ScanQ}$. The revised algorithm is Algorithm 2.4.1.

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Algorithm 2.4.1: $\operatorname{BFS}(G, u)$ global $n$
for $u \leftarrow 1$ to $n$
do $\left\{\begin{array}{l}\text { dist }[v] \leftarrow \infty \\ \text { onScanQ } Q[v] \leftarrow \text { false }\end{array}\right.$ dist[u] $\leftarrow 0$
ScanQ[1] $\leftarrow u$
onScanQ[u]<true
QSize -1
$k \leftarrow 1$
repeat
$u \leftarrow S c a n Q[k]$
for each $w \longrightarrow v$
do if not onScanQ[w]
then $\left\{\begin{array}{l}\text { Scan } Q[Q S i z e] \leftarrow w \\ \text { cis }\end{array}\right.$
onScanQ $[w] \leftarrow$ true
$\operatorname{dist}[w] \leftarrow \operatorname{dist}[v]+1$
$k \leftarrow k+1$
until $k>Q$ Size
The initialization takes $2 n$ steps. The repeat-loop runs at most $n$ times. At most $n$ vertices are placed on the queue. The for-loop over all $w \longrightarrow v$ requires

$$
\sum_{v} \operatorname{DEG}(v)=2 \varepsilon
$$

steps, all told. This assumes that the adjacent vertices are stored in a linked list-the for-loop traverses the adjacency list. Therefore the total number of steps executed is at most

$$
3 n+2 \varepsilon=O(n+\varepsilon)=O(\varepsilon) \text {. }
$$

Notice that in this program we could have dispensed with the array onScanQ, by using instead dist[ w]=m to determine $w$ is on ScanQ. Because a breath-first search always uses a queue but not always a dist[•] array, we have kept the boolean array, too.

### 2.5 Weighted graphs and Dijkstra's algorithm

A breath-first search calculates $\operatorname{DIST}(u, U)$ correctly because in a simple graph, each edge has "length" one; that is, the length of a path is the number of edges
page_33
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it contains. In a more general application where graphs are used to model a road network, or distribution network, etc., we may want to assign a length $\geq 1$ to each edge. This is illustrated in Figure 2.5.


FI GURE 2.5
A weighted graph
This is an example of a weighted graph. Each edge $u v \in E(G)$ is assigned a positive integral weight WT(uu). WT (uu) may represent length, weight, cost, capacity, etc., depending on the application. In a weighted graph, the length of a path $P=(u 0, u 1, \ldots, u k)$ is

$$
\ell(P)=\sum_{i=0}^{k-1} \mathrm{~W} \mathrm{~T}\left(u_{i} u_{i+1}\right)
$$

The distance between two vertices is now defined as
DIST $(u, u)=\mathrm{MIN}\{\ell(P): P$ is a $u u$-path $\}$
A breath-first search will not compute DIST $(u, U)$ correctly in a weighted graph, because a path with more edges may have the shorter length. There are many algorithms for computing shortest paths in a weighted graph. Dijkstra's algorithm is one.
page_34
Page 35
procedure DIJKSTRA(u)
comment: $\left\{\begin{array}{l}\text { Compute } \operatorname{DiST}(u, v), \text { for all } v \in V(G) \\ \text { dist }[v] \text { will equal DIST }(u, v) \text { upon completion. } \\ \text { Vertices are chosen as } u_{1}, u_{2}, \ldots, u_{n}, \\ \text { in order of their distance from } u .\end{array}\right.$ $u 1 \leftarrow u$ "the nearest vertex to $u$."
for $k \leftarrow 1$ to $n-1$
comment: $\left\{\begin{array}{l}u_{1}, u_{2}, \ldots, u_{k} \text { are currently known } \\ \text { in this iteration, } u_{k+1} \text { is selected }\end{array}\right.$
do
select $v$, the nearest vertex to $u_{1}$, such that $v \notin\left\{u_{1}, u_{2}, \ldots, u_{k}\right\}$
$u_{k+1} \leftarrow v$
assign dist $\left[u_{k+1}\right]$
comment: $u_{1}, u_{2}, \ldots, u_{k+1}$ are now known
comment: all dist[ui] are now known
Dijkstra's algorithm is an example of a so-called "greedy" or "myopic" algorithm, that is, an algorithm which always selects the next nearest, or next best, etc., on each iteration. Many problems can be solved by greedy algorithms.
We need to know how to choose $v$, the next nearest vertex to $u 1$, in each iteration. On the first iteration, it will be the vertex $v$ adjacent to $u 1$ such that $W T(u 1 v)$ is minimum. This will $\{u 1, u 2\}$ such that DIST( $u 1, u 1$ ) and $\operatorname{DIST}(u 1, u 2)$ are known. On the next iteration, the vertex ? chosen will be adjacent to one of $u 1$ or $u 2$. The distance to $u l$ will then be either

$$
\operatorname{DIST}(u 1, u 1)+\mathrm{WT}(u 1 v)
$$

and $v$ will be the vertex for which this sum is minimum.
In general, at the beginning of iteration $k$, vertices $u 1, u 2 \ldots, u k$ will have been chosen, and for these vertices, DIST[ui]=DIST(u1, ui).
The next nearest vertex $v$ must be adjacent to some ui, so that the shortest $u 1 v$-path will have length
dist[ui]+WT(uiv), for some $i$. $v$ is chosen as the vertex for which this value is a minimum. This is illustrated in Figure 2.6. The refined code for Dijkstra's algorithm is Algorithm 2.5.1.

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$u_{1}$


FI GURE 2.6
A shortest $u 1 u$-path, via vertex $u k$
Algorithm 2.5.1: DIJ KSTRA(u)
Compute $\operatorname{Dist}(u, v)$, for all $v \in V(G)$
comment:
$\operatorname{dist}[v]$ will equal $\operatorname{DIST}(u, v)$ upon completion.
Vertices are chosen as $u_{1}, u_{2}, \ldots, u_{n}$,
in order of their distance from
for each $u$
do $\operatorname{dist}[U] \leftarrow \infty$
$u 1 \leftarrow u$ "the nearest vertex to $u$ " $\operatorname{dist}[u] \leftarrow 0$
for $k \leftarrow 1$ to $n-1$
do
comment: $\left\{\begin{array}{l}u_{1}, u_{2}, \ldots, u_{k} \text { are currently known } \\ \text { in this iteration, } u_{k+1} \text { is selected }\end{array}\right.$ for each $v \longrightarrow u_{k}$ such that $v \notin\left\{u_{1}, u_{2}, \ldots, u_{k}\right\}$
do $\operatorname{dist}[v] \leftarrow \operatorname{MIN}\left(\operatorname{dist}[v], \operatorname{dist}\left[u_{k}\right]+\mathrm{WT}\left(u_{k} v\right)\right)$
pick $v \notin\left\{u_{1}, u_{2}, \ldots, u_{k}\right\}$ such that $\operatorname{dist}[v]$ is minimum
$u_{k+1} \leftarrow v$
comment: $\left\{\begin{array}{l}\operatorname{dist}\left[u_{k+1}\right] \text { now equals } \operatorname{DIST}\left(u_{1}, u_{k+1}\right), \text { and } \\ u_{1}, u_{2}, \ldots, u_{k+1} \text { are now known }\end{array}\right.$

## Exercises

2.5.1 Prove that Dijkstra's algorithm works. Use induction on the number $k$ of iterations to prove that at the beginning of iteration $k$, each dist[ui]=DIST ( $u 1, u i$ ), and that for all $u \neq u i$, for any $i$, dist[ $u$ ] equals the length of a shortest $u 1 u$-path using only the vertices $\{u 1, u 2, \ldots, u k-1, u\}$. Conclude that after $n-1$ iterations, all distances dist[u]=DIST(ul,u).
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2.5.2 Assuming that $G$ is stored in adjacency lists, and that the minimum dist[v] is computed by scanning all $n$ vertices, show that the complexity of Dijkstra's algorithm is $O(e+n 2)$.

### 2.6 Data structures

When computing the distances $\operatorname{DIST}(u 1, v)$, it would also be a good idea to store a shortest $u l v$-path. All the $u 1 v$-paths can easily be stored using a single array
PrevPt[v]: the previous point to $v$ on a shortest ulv-path.
I nitially, PrevPt[u] $\leftarrow 0$. When dist[v] and dist[uk]+WT(ukv) are compared, if the second choice is smaller, then assign PrevPt $[v] \leftarrow u k$. The shortest $u l v$-path can then be printed by the following loop:
repeat
output (v)
$v \leftarrow$ PrevPt[v]
until $v=0$
The complexity of Dijkstra's algorithm was calculated in Exercise 2.5.2 above as $O(n 2+e)$. The term $O(n 2)$ arises from scanning up to $n$ vertices in order to select the minimum vertex ?. This scanning can be eliminated if we store the vertices in a partially ordered structure in which the minimum vertex is always readily available. A heap is such a structure. In a heap $H$, nodes are stored so that the smallest element is always at the top.
A heap is stored as an array, but is viewed as the partially ordered structure shown above. Its elements are not sorted, but satisfy the heap property, namely that $H[i]=H[2 i]$ and $H[i]=H[2 i+1]$; that is, the value stored in each node is less than or equal to that of either of its children. Therefore, $H[1]$ is the smallest entry in the array.
The heap shown above has depth four; that is, there are four levels of nodes. A heap of depth $k$ can contain up to $2 k-1$ nodes, so that the depth needed to store $N$ values is the smallest value of $k$ such that $2 k-1=N$, namely, $k=\lceil\log (N+1)\rceil$, where the log is to base 2 .
If the value stored in a node is changed so that the heap property is no longer satisfied, it is very easy to update the array so that it again forms a heap. For example, if $H$ [10] were changed to 4 , then the following loop will return $H$ to heap form. The movement of data is shown in Figure 2.8.

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## FI GURE 2.7

A heap
procedure FLOATUP(k)
comment: Element $H[K]$ floats up to its proper place in the heap
temp $\leftarrow H[k]$
$j \leftarrow k / 2$
while temp $<H[j]$ and $j>0$
do $\left\{\begin{array}{l}H[k] \leftarrow H[j] \\ k \leftarrow j \\ j \leftarrow k / 2\end{array}\right.$
$H[k] \leftarrow$ temp
Notice the circular movement of data when an altered element floats up to its proper place in the heap. If some entry in the heap were made larger, say H[1] became equal to 10, then a similar loop (Procedure FLOATDOWN) would cause the new value to float down to its proper place. Since the depth of a heap containing $N$ items is $\lceil\log (N+1)\rceil$, the number of items moved is at most $1+\lceil\log (N+1)\rceil$.

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## FI GURE 2.8 Updating a heap with FLOATUP. procedure FLOATDOWN $(k)$

 the entry at $H[k]$ has been increased - it now floats down the heap to its correct location. There are currently $n$ entriescomment: $\{$ in the heap, $H[1]$ to $H[n]$. The array $H[\cdot]$ has been dimensioned so that $H[0]$ is also available as a sentinel. $H[0]$ contains a large value bigger than any valid heap entry.
temp $\leftarrow H[k]$
while $k+k=n$
$\left\{\begin{array}{l}i \leftarrow k+k \quad \text { "the left child of } H[k] \text { ". } \\ j \leftarrow i+1 \quad \text { "the right child of } H[k] \text { ". } \\ \text { if } j>n \\ \text { then } j \leftarrow 0 \quad \text { "the sentinel at } H[0] \text { ". }\end{array}\right.$
if $H[i]>H[j]$
then $i \leftarrow j$
comment: $H[i]$ is now the smaller child
if temp $\leq H[i]$
then break "break out of loop"
$H[k] \leftarrow H[i]$
$k \leftarrow i$
$H[k] \leftarrow$ temp
In order to extract the smallest item from a heap of $N$ elements, we take its value from $H[1]$, and then perform the following steps:

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$H[1] \leftarrow H[N]$
$\mathrm{N} \leftarrow \mathrm{N}-1$
FLOATDOWN(1)
The new $H 1]$ floats down at most $1+\lceil\log N\rceil$ steps.
There are two ways of building a heap.
procedure BUILDHEAPTOPDOWN(H,N)
comment: $\left\{\begin{array}{l}\text { The array } H \text { contains } N \text { entries } \\ \text { transform it into a heap }\end{array}\right.$

```
        comment: the first \(k\) values in \(H\) already form a heap
do
\(k \leftarrow k+1\)
    \(\operatorname{FloatUp}(k)\)
```

Using this method, the elements in entries $1,2, \ldots, k$ already form a subheap with $k$ entries. On each iteration, a new entry is allowed to float up to its proper position so that the first $k+1$ values now form a heap. There are two nodes on level two of the heap. The FLOATUP() operation for each of these may require up to $1+2=3$ data items to be moved. On level three there are four nodes. FLOATUP() may require up to $1+3=4$ data items to be moved for each one. In general, level $k$ contains $2 k-1$ nodes, and FLOATUP() may need up to $1+k$ data items to be moved for each. The total number of steps to create a heap with $d=\lceil\log (N+1)\rceil$ levels in this way is therefore at most

$$
S=3 \cdot 2^{1}+4 \cdot 2^{2}+5 \cdot 2^{3}+\ldots+(1+d) 2^{d-1}=\sum_{k=2}^{d-1}(1+k) 2^{k-1}
$$

Therefore

$$
2 S=3.22+4.22+5.23+\ldots+(1+d) 2 d,
$$

so that

$$
\begin{gathered}
2 S-S=(1+d) 2 d-3 \cdot 21-(22+23+\ldots+2 d-1) \\
=(1+d) 2 d-5-(1+2+22+23+\ldots+2 d-1) \\
=(1+d) 2 d-5-(2 d-1) \\
=d 2 d-4
\end{gathered}
$$

Thus, it takes $O(N \log N)$ steps to build a heap in this way.
The second way of building a heap is to use FLOATDOWN().
page_40
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procedure BUILDHEAPBOTTOMUP $(H, N)$
comment: $\left\{\begin{array}{l}\text { The array } H \text { contains } N \text { entries } \\ \text { transform it into a heap }\end{array}\right.$
$k \leftarrow N / 2$
while $k \geq 1$
do $\left\{\begin{array}{l}\text { comment: }\left\{\begin{array}{l}\text { the substructures at nodes } H[2 k] \text { and } H[2 k+1] \\ \text { already form subheaps }\end{array}\right.\end{array}\right.$
$\left\{\begin{array}{l}\text { FloatDown }(k) \\ k \leftarrow k-1\end{array}\right.$
This way is much more efficient, requiring only $O(N)$ steps, as is proved in Exercise 2.7.1 below.
We can use a heap $H$ to store the values dist[v] in Dijkstra's algorithm. The main loop now looks like this. $u 1 \leftarrow u$ "the nearest vertex to $u$."
for $k \leftarrow 1$ to $n-1$

$$
\text { do }\left\{\begin{array}{l}
\text { comment: } u_{1}, u_{2}, \ldots, u_{k} \text { are currently known } \\
\text { for each } v \longrightarrow u_{k} \text { such that } v \notin u_{1}, u_{2}, \ldots, u_{k} \\
\text { do }\left\{\begin{array}{c}
\text { if } \operatorname{dist}[v]>\operatorname{dist}\left[u_{k}\right]+\mathrm{W}\left(u_{k} v\right) \\
\text { then }\left\{\begin{array}{l}
\text { dist }[v] \leftarrow \operatorname{dist}\left[u_{k}\right]+\mathrm{WT}\left(u_{k} v\right) \\
\mathrm{FLOATUP}(v)
\end{array}\right. \\
\text { "which entry corresponds to } v ? \text { ?" } \\
\text { choose } u_{k+1} \text { using } H[1] \\
H[1] \leftarrow H[n-k] \\
\text { remove last entry from } H
\end{array}\right. \\
\text { FLOATDown }(1) \\
\text { comment: } u_{1}, u_{2}, \ldots, u_{k+1} \text { are now known }
\end{array}\right.
$$

Notice that the FLOATUP(v) operation requires that we also know which node $H[k]$ in the heap corresponds to
the vertex $v$, and vice versa. This can be done with an array mapping vertices into the heap. Let us work out the complexity of Dijkstra's algorithm using this data structure. It is not possible to get an accurate estimate of the number of steps performed in this case, but only an upper bound. The initialization of the heap and dist[•] array take $O(n)$ steps. The inner for-loop executes a total of at most $2 e$ if-statements, so that at most $2 e$ FLOATUP()'s are performed, each requiring at most $1+\lceil\log (n+1)\rceil$ steps. There are also $n-1$ FLOATDOWN()'S performed. Thus the complexity is now

$$
O((2 \varepsilon+n)(1+\lceil\log (n+1)\rceil))=O(\varepsilon \log n)
$$

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This may be better or worse that the previous estimate of $O(n 2)$ obtained when the minimum vertex is found by scanning up to $n$ vertices on each iteration. If the graph has few edges, say $\Sigma \leq \Delta n / 2$, where the maximum degree? is some fixed constant, or a slowly growing function of $n$, then Dijkstra's algorithm will certainly be much more efficient when a heap is used. Furthermore, it must be remembered that the complexity estimate using the heap is very much an upper bound, whereas the other method will always take at least $O(n 2)$ steps. If the number of edges is large, say $\Sigma=O(n 2)$, then the heap-version of Dijkstra's algorithm can spend so much time keeping the heap up-to-date, that no increase in efficiency is obtained.

### 2.7 Floyd's algorithm

Floyd's algorithm solves the All Paths Problem, computing a matrix of values Dist[ $u, v]=\operatorname{DIST}(u, v)$, for all $u, v$ $v \in V(G)$. Initially, Dist $[\cdot \cdot \cdot]$ equals the weighted adjacency matrix $A$, where

$$
A[u, v]= \begin{cases}\mathrm{WT}(u, v), & \text { if } u \longrightarrow v \\ \infty, & \text { if } u \not v v \\ 0, & \text { if } u=v\end{cases}
$$

Floyd's algorithm is extremely simple to program.
procedure FLOYD(Dist)
comment: Dist[ $u, v$ ] will equal DIST ( $u, v$ ), upon completion
for $k \leftarrow 1$ to $n$
do $\left\{\begin{array}{l}\text { for } v \leftarrow 1 \text { to } n-1 \\ \quad \operatorname{do}\left\{\begin{array}{l}\text { for } w \leftarrow v+1 \text { to } n \\ \text { do } \operatorname{Dist}[v, w] \leftarrow \operatorname{MIN}\left(\operatorname{Dist}[v, w], \operatorname{Dist}\left[v, u_{k}\right]+\operatorname{Dist}\left[u_{k}, w\right]\right)\end{array}\right.\end{array}\right.$
The for-loops for $v$ and $w$ together examine $\binom{n}{2}$ pairs $v w$ for each value of $u$, so the complexity of the algorithm is

$$
n\binom{n}{2}=\frac{1}{2} n^{3}-\frac{1}{2} n^{2}=O\left(n^{3}\right)
$$

The graph is stored as a weighted adjacency matrix, in which non-adjacent vertices $v$, $w$ can be considered to be joined by an edge of weight $\infty$. Figure 2.9 shows a weighted graph on which the reader may like to work Floyd's algorithm by hand.
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|  | $u_{1}$ | $u_{2}$ | $u_{3}$ | $u_{4}$ | $u_{5}$ |
| :---: | :---: | :---: | :---: | :---: | :---: |
| $u_{1}$ | 0 | 5 | $\infty$ | 3 | 2 |
| $u_{2}$ | 5 | 0 | 4 | $\infty$ | 2 |
| $u_{3}$ | $\infty$ | 4 | 0 | 1 | 4 |
|  | 3 | $\infty$ | 1 | 0 | 6 |
| $u_{5}$ | 2 | 2 | 4 | 6 | 0 |

FI GURE 2.9

## A complete weighted graph and its weighted adjacency matrix.

Let the vertices of $G$ be named $u 1, u 2, \ldots$, un. In order to prove that Floyd's algorithm works, we prove by induction, that at the end of kth iteration of the for-loop for $u, \operatorname{Dist}[v, w]$ is the length of the shortest $u w$-path which uses only vertices $u, w$, and $u 1, u 2, \ldots, u k$. When $k=0$, that is, before the first iteration, $\operatorname{Dist}[u, w]$ is the length of the edge $u w$, that is, the length of the shortest path using only vertices $u$ and $w$. At the end of the first iteration, Dist $[u, w]=\mathrm{MIN}(\mathrm{WT}(u, w), \mathrm{WT}(u, u 1)+\mathrm{WT}(u 1, w))$. This is the length of the shortest $u w$-path using only vertices $u, w$, and $u 1$, since that path either uses $u 1$, or else consists only of the edge $u w$. Thus, the statement is true when $k=1$.

$w$
FI GURE 2.10
A path viaut+1.
Assume that it is true whenever $k \leq t$, and consider iteration $t+1$. At the end
page_43
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of the iteration, each
$\operatorname{Dist}[v, w]=\mathrm{MIN}(\operatorname{Dist}[v, w], \operatorname{Dist}[v, u t+1]+\operatorname{Dist}[u t+1, w])$.
If the shortest $v w$-path using only vertices $v, w, u 1, u 2, \ldots, u t+1$ does not use $u t+1$, then its length is the previous value of $\operatorname{Dist}[v, w]$ from iteration $t$. If the path does use $u t+1$, then the length is given by the second term above. Therefore, at the end of the iteration, the value of Dist $[v, w]$ is as required. By induction, it follows that at the end of the nth iteration, $\operatorname{Dist}[v, w]=\operatorname{DIST}(v, w)$, for all $v$ and $w$. Floyd's algorithm finds all distances in the graph. It always takes $n\binom{n}{2}=O\left(n^{3}\right)$ steps, irrespective of the number of edges of $G$. When there are few edges, it is faster to use Dijkstra's algorithm $n$ times, once for every starting vertex $u$. This gives a complexity of $O(e n \log n)$, using a heap, which can be less than $O(n 3)$.

## Exercises

2.7.1 Calculate the number of steps needed to construct a heap using the BUILDHEAPBOTTOMUP() procedure.
2.7.2 The repeat-loop of the FLOATUP() procedure described above requires $k+2$ data items to be moved when an entry floats up $k$ nodes in the heap. If FLOATUP() is programmed by swapping adjacent elements instead of moving them in a cycle, calculate the number of items moved when an entry floats up $k$ nodes. Which is more efficient?
2.7.3 The type of heap discussed above is called a binary heap, since each node $H[k]$ has two children, $H[2 k]$ and $H[2 k+1]$. The depth of a binary heap with $N$ elements is $\lceil\log (N+1)\rceil$. In a ternary heap, node $H[k]$ has three children, $H[3 k], H[3 k+1]$, and $H[3 k+2]$. What is the depth of a ternary heap with $N$ nodes? Calculate the number of steps needed to construct it using the BUILDHEAPBOTTOMUP() procedure.
2.7.4 Program Dijkstra's algorithm using a binary heap.
2.7.5 Show how to store a complete set of shortest paths in Floyd's algorithm, using a matrix PrevPt[ $v, w]$, being the previous point to $v$ on a shortest $w w$-path. What should the initial value of $\operatorname{PrevPt}[v, w]$ be, and how and when should it be modified?
2.7.6 Ford's algorithm. Consider the following algorithm to find DIST $(u, v)$, for a given vertex $u \in V(G)$ and all vertices $v \in V(G)$.
procedure FORD (u)
for each $v \in V(G)$
do $\operatorname{dist}[V] \leftarrow \infty$
dist[u] $\leftarrow 0$
while there is an edge $v w$ such that dist[ $[w]>\operatorname{dist}[v]+\mathrm{WT}[v w]$
do dist[ w$] \leftarrow \operatorname{dist}[\mathrm{v}]+\mathrm{WT}[\mathrm{w} \mathrm{w}]$

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Prove that Ford's algorithm correctly computes DIST $(u, u)$. What data structures are necessary for an efficient implementation of Ford's algorithm? Analyze the complexity of Ford's algorithm. Give a numerical estimate of the number of steps, as well as a formula of the form $O(\cdot)$.

### 2.8 Notes

WEISS [122] contains an excellent treatment of the merge-find data structure and heaps. Dijkstra's shortestpath algorithm and Floyd's algorithm are described in most books on algorithms and data structures.
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3
Some Special Classes of Graphs

### 3.1 Bipartite graphs

A graph $G$ is said to be bipartite if $V(G)$ can be divided into two sets $X$ and $Y$ such that each edge has one end in $X$ and one end in $Y$. For example, the cube is a bipartite graph, where the bipartition ( $X, Y$ ) is illustrated by the coloring of the nodes in Fiqure 3.1.

(a)

(b)

FI GURE 3.1

## Two bipartite graphs

The maximum number of edges in a simple bipartite graph in which $X$ and $Y$ are the two sides of the bipartition is clearly $|X| \cdot|Y|$. The complete bipartite graph $K m, n$ has $|X|=m,|Y|=n$, and $\varepsilon=m n$. For example, $K 3,3$ is illustrated in Figure 3.1.
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LEMMA 3.1 A simple, bipartite graph $G$ has at most $|G| 2 / 4$ edges.
PROOF Let $G$ have bipartition $(X, Y)$, where $|X|=x$ and $|Y|=n-x$, where $n=|G|$. Then $\varepsilon \leq x(n-x)=n x-x 2=n 2 / 4-(n / 2-x) 2 \leq n 2 / 4$.
If $C=(x 1, y 1, x 2, y 2, \ldots)$ is a cycle in a bipartite graph $G$, then consecutive vertices of $C$ must be alternately in $X$ and $Y$, the two sides of the bipartition. It follows that $\ell(C)$ is even. In fact, any graph in which all cycles have even length must be bipartite.
THEOREM 3.2 G is bipartite if and only if all cycles of $G$ have even length.
PROOF Let $G$ be a connected graph in which all cycles have even length. Pick any $x \in V(G)$ and set $X=\{u$ : $\operatorname{DIST}(x, u)$ is even\}, and $Y=V(G)-X$. Clearly $X$ and $Y$ partition $V(G)$ into two parts. We must show that there are no edges with both endpoints in $X$ or $Y$. Suppose that uu is an edge with $u, v \in X$. Let Pu be a shortest $x u$-path, that is, a path of length DIST $(x, u)$, and let $P u$ be a shortest $x u$-path. Then $\ell(P u)$ and $\ell(P u)$ are both even. Say $\ell(P u) \leq \ell(P u)$. Pu and Pu both begin at point $x$. They do not both contain $u$, or Puuu would be a shortest $x u$-path of length $\ell(P u)+1$, an odd number. So let $z$ be the last point in common to $P u$ and $P u$. This defines the cycle $C=P u[z, u] u u P u[u, z]$. Here $P u[z, u]$ denotes the portion of $P u$ from $z$ to $u$ and $P u[u, z]$ denotes
the portion of $P u$ from $u$ to $z$. The length of $C$ is then $\ell(P u[z, u])+\ell(P u[u, z])+1=\ell(P u)+\ell(P u)-2 D I S T(x, z)+1$, which is odd, a contradiction. Therefore no edge $u u$ has both endpoints in $X$. Similarly, no edge uu has both endpoints in $Y$. Since a graph is bipartite if and only if every component is bipartite, this completes the proof.


FI GURE 3.2
Two paths in a bipartite graph

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LEMMA 3.3 If $G$ is a $k$-regular bipartite graph, where $k>0$, with bipartition $(X, Y)$, then $|X|=|Y|$.
PROOF Since each edge has one end in $X$, we can write $\varepsilon=\Sigma_{x \in X} \operatorname{DEG}(x)=k \cdot|X|$. Similarly, $\varepsilon=\Sigma_{y \in Y} \operatorname{DEG}(y)=k \cdot|Y|$. Therefore $k \cdot|X|=k \cdot|Y|$. Since $k>0$, it follows that $|X|=|Y|$.

## Exercises

3.1.1 The $k$-cube $Q k$ is a graph whose vertex set consists of all binary vectors of length $k$ :

$$
V\left(Q_{k}\right)=\left\{\left(a_{1}, a_{2}, \ldots, a_{k}\right): a_{i} \in\{0,1\}\right\}
$$

Thus there are $2 k$ vertices. The edges of $Q k$ are formed by joining two vertices $\vec{a}=\left(a_{1}, a_{2}, \ldots, a_{k}\right)$ and $\vec{b}=\left(b_{1}, b_{2}, \ldots, b_{k}\right)$ if $\vec{a}$ and $\vec{b}$ differ in exactly one coordinate, that is, ai=bi for all $i$ but one. $Q 3$ is displayed in Fiqure 3.3. Prove that $Q k$ is bipartite. Describe a bipartition of $Q k$.


FI GURE 3.3
The 3-cube, Q3
3.1.2 Prove that $\varepsilon(Q k)=k 2 k-1$.
3.1.3 Describe in pseudo-code an algorithm to find a bipartition of $G$, or to determine that $G$ is not bipartite. Describe the data-structures needed, and calculate the complexity (should be $O(\varepsilon)$ ).
3.1.4 Let $G$ be a bipartite simple graph with bipartition $(X, Y)$ and n vertices. Let $\delta x$ be the minimum degree among the vertices of $X$, and $\delta Y$ be the minimum degree

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among the vertices of $Y$. Show that if $\delta X+\delta Y>n / 2$, then $G$ is connected, where $\delta X, \delta Y>0$.

### 3.2 Line graphs

Two edges of a graph $G$ are adjacent if they share a common endpoint. The linegraph of $G$ is a graph $L(G)$ which describes the adjacencies of the edges of $G$. Thus, every vertex of $L(G)$ corresponds to an edge $u u$ of $G$, so that $|L(G)|=\varepsilon(G)$. This is illustrated in Figure 3.4.


FI GURE 3.4

## Constructing a line-graph

A line-graph can always be decomposed into complete subgraphs. For a vertex $v \in V(G)$ lies on DEG( $u$ ) distinct edges all of which share the endpoint $u$. The DEG(U) corresponding vertices of $L(G)$ form a complete subgraph containing $\left(\begin{array}{c}\mathrm{DEG}(u)\end{array}\right)$ edges. Every edge of $L(G)$ is contained in exactly one such complete subgraph.


## $\operatorname{DEG}(v)$ edges

## FI GURE 3.5

Complete subgraph in a line-graph
This gives the following theorem:

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THEOREM $3.4 \quad \sum_{u \in V(G)} \quad 2$

## Exercises

3.2.1 Find the line-graph of the cube.
3.2.2 Construct $\overline{L\left(K_{5}\right)}$ and show that it is isomorphic to the Petersen graph.
3.2.3 Let $G$ be any graph. If we insert a vertex of degree two into each edge, we obtain a new graph $S(G)$, called the subdivision graph of $G$. For example, $S(K 4)$ is illustrated in Figure 3.6. Prove that $S(G)$ is always bipartite, and find a formula for $\varepsilon(S(G))$.

$K_{4}$

$S\left(K_{4}\right)$

FI GURE 3.6

## Subdivision graph of $K \mathbf{4}$

3.2.4 The graph $P$ in Figure 3.7 is called the 3 -prism. Find the line-graphs of the subdivision graphs of $K 4$ and $P$. Draw them as neatly as possible. What can you say in general about constructing the line-graph of the subdivision graph of a 3 -regular graph?


FI GURE 3.7
The 3-prism
3.2.5 We know that

$$
\sum_{u} \operatorname{DEG}(u)=2 \varepsilon(G),
$$

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and that

$$
\sum_{u}\binom{\operatorname{DEG}(u)}{2}=\varepsilon(L(G)) .
$$

Can you find a similar way of interpreting

$$
\sum_{u}\binom{\operatorname{DEG}(u)}{3} ?
$$

Assume first that there are no triangles in $G$.
3.2.6 Suppose that a graph $G$ is represented by its adjacency matrix. Write a program to print out the adjacency lists of $L(G)$, but do not store either adjacency lists or an adjacency matrix for $L(G)$; just print it out. Also print out a list of the edges of $G$, in order to give a numbering to the vertices of $L(G)$.
3.2.7 Notice that $L\left(K_{3}\right) \cong L\left(K_{1,3}\right) \cong K_{3}$ (see Figure 3.8). Prove that if $G$ and $H$ are any other graphs, then $G \cong H$ if $L(G) \cong L(H)$.


## FI GURE 3.8

## Two graphs with isomorphic line-graphs

### 3.3 Moore graphs

The length of the shortest cycle in a graph $G$ is called its girth, denoted $\gamma(G)$. For example, the cube has girth four. Graphs with fixed degree $k$ and fixed girth often have interesting properties. For example, let $G$ be a $k$ regular graph of girth four, and pick any vertex $u$ in $G$. There are $k$ vertices at distance one from $u$. Since $G$ has no triangles, there are at least $k-1$ vertices at distance two from $u$, as shown in Figure 3.9. Therefore, $|G| \geq 1+k+(k-1)=2 k$. There is only one such graph with $|G|=2 k$, and that is the complete bipartite graph $K k$, k.

Now let $G$ be a $k$-regular graph of girth five, and let $u$ be any vertex, There are $k$ vertices at distance one from $u$. Since $G$ has no 4-cycles, each point at distance one is adjacent to $k-1$ more vertices at distance two, so that $|G| \geq 1+k+k(k-1)=k 2+1$.
Problem. Are there any $k$-regular graphs $G$ of girth five with $|G|=k 2+1$ ?
These graphs are called Moore graphs. Let $n=|G|$. A l-regular graph cannot have $\gamma=5$, so $k \geq 2$. If $k=2$, then $n=22+1=5 . G$ is a cycle of length five. This is the unique Moore graph of degree two.

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$\geq k-1$ points
$k$ points

1 point

## FI GURE 3.9

## Kk,k

If $k=3$, then $n=32+1=10$. There are three vertices at distance one from $u$, and six at distance two, as illustrated in Figure 3.10. Consider vertex u6. $u_{6} \nrightarrow u_{5}$, since this would create a triangle, whereas $\gamma=5$. Without loss of generality, we can take $u 6 \rightarrow u 8$. Were we now to join $u 6 \rightarrow u 7$, this would create a 4-cycle ( $u 6$, $u 7, u 3, u 8)$, which is not allowed. Therefore, without loss of generality, we can take $u 6 \rightarrow u 9$. This is shown in Figure 3.10.


FI GURE 3.10
A Moore graph of degree three
There is now only one way of completing the graph so as to maintain $\gamma=5$. Vertex $u 9$ cannot be joined to $u 5$, $u 8$, or $u 10$. Therefore $u 9 \rightarrow u 7$. Similarly $u 8 \rightarrow u 10$, etc. The completed graph is shown in Figure 3.10 , and has been redrawn in Figure 3.11 (check that this is the same graph).
Thus, we have proved the following.
THEOREM 3.5 The Petersen graph is the unique Moore graph of degree three.
There is a very elegant theorem proving that Moore graphs can exist only for

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## FI GURE 3.11

## The Petersen graph

special values of $k$.
THEOREM 3.6 A Moore graph of degree $k$ can exist only if $k=2,3,7$, or 57.
PROOF Let $G$ be a Moore graph with adjacency matrix $A$ and consider $A 2$. Entry $[i, j]$ of $A 2$ is the number of uiuj-paths of length two. If $u_{i} \longrightarrow u_{j}$, then there is no 2 -path from $u i$ to $u j$, since $\gamma=5$. Therefore, [A2]ij=0 if $[A] i i=1$.
If $u_{i} \nrightarrow u_{j}$ then $\operatorname{DIST}(u i, u j)>1$. It is shown in Exercise 3.3.3 that DIST ( $u i, u j$ ) is always at most 2 .
Therefore, if $u_{i} \nrightarrow u_{j}$ there must be a 2 -path connecting ui to $u j$. There cannot be two such 2 -paths, for that would create a 4 -cycle containing ui and $u j$. Therefore, $[A 2] i j=1$ if $[A] i j=0$.
It follows that the matrix $A 2+A$ consists of all 1's off the diagonal. The diagonal elements all equal $k$, the degree of $G$. The number of vertices is $n=k 2+1$.

$$
A^{2}+A=\left[\begin{array}{cccc}
k & & & 1 \\
& k & & \\
1 & & \ddots & \\
& & & k \times n \\
\end{array}\right]_{n \times n}
$$

We can find the eigenvalues of this matrix. Write $B=A 2+A$. If x is an eigenvector of $A$ with eigenvalue $a$, then

$$
B x=(A 2+A) x=A A x+A x=a A x+a x=(a 2+a) x
$$

so that $\beta=a 2+a$ is an eigenvalue of $B$. To find the eigenvalues of $B$, we solve $\operatorname{det}(\lambda I-B)=0$ for $\lambda$.

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$$
\operatorname{det}(\lambda I-B)=\left[\begin{array}{cccc}
\lambda-k & & & -1 \\
& \lambda-k & & \\
-1 & \ddots & \\
& & & \lambda-k
\end{array}\right]_{n \times n}
$$

Adding rows 2 to $n$ onto row 1 gives

$$
\begin{aligned}
& \left|\begin{array}{cccc}
\lambda-k-n+1 & \lambda-k-n+1 & \cdots & \\
-1 & \lambda-k & & \\
& & \ddots & \\
& & & \lambda-k
\end{array}\right|_{n \times n} \\
& =(\lambda-k-n+1)\left|\begin{array}{cccc}
1 & 1 & \cdots & \\
-1 & \lambda-k & & \\
& & \ddots & \lambda-k
\end{array}\right|_{n \times n}
\end{aligned}
$$

Now add the first row to each row to get

$$
\begin{gathered}
(\lambda-k-n+1)\left|\begin{array}{cccc}
1 & 1 & \cdots & \\
0 & \lambda-k+1 & & \\
& & \ddots & \\
& & & \lambda-k+1
\end{array}\right|_{n \times n} \\
=(\lambda-k-n+1)(\lambda-k+1)^{n-1}=0 .
\end{gathered}
$$

Therefore the eigenvalues of $B$ are

$$
\lambda= \begin{cases}\beta_{1}=k+n-1 & \text { (once) }, \\ \beta_{2}=k-1 & (n-1 \text { times }) .\end{cases}
$$

Since $\beta=a 2+a$, we can solve for $\alpha=\frac{1}{2}(-1 \pm \sqrt{1+4 \beta})$. Should we take the plus or minus sign? Since $n=k 2+1$, the value $\beta 1=k 2+k+1$ qives

$$
\alpha=\frac{1}{2}\left(-1 \pm \sqrt{4 k^{2}+4 k+1}\right)=\frac{1}{2}\{-1 \pm(2 k+1)\}=k \text { or }-k-1 .
$$

Now $\beta 1$ occurs only once as an eigenvalue, so we must choose only one of these. $G$ is $k$-regular, so that the rows of $A$ all sum to $k$. Thus, if $x$ is the vector of all 1's, then $A x=k x$, so that $k$ is in fact an eigenvalue of $A$. Consider now $\beta 2$. The corresponding eigenvalues of $A$ are

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$$
\alpha= \begin{cases}\alpha_{1}=\frac{1}{2}(-1+\sqrt{4 k-3}) & \left(m_{1} \text { times }\right), \\ \alpha_{2}=\frac{1}{2}(-1-\sqrt{4 k-3}) & \left(m_{2} \text { times }\right) .\end{cases}
$$

The total multiplicity is $m 1+m 2=n-1=k 2$. Since the trace of $A$, that is, the sum of its diagonal elements, also equals the sum of its eigenvalues, we can write

$$
a 1 m 1+a 2 m 2+k=0 \text { (sum of eigenvalues) }
$$

Solving these equations for $m 1$ and $m 2$ gives

$$
m_{1}=\frac{-1}{2}\left\{\frac{-k^{2}+2 k}{\sqrt{4 k-3}}-k^{2}\right\}
$$

and

$$
m_{2}=\frac{1}{2}\left\{\frac{-k^{2}+2 k}{\sqrt{4 k-3}}+k^{2}\right\}
$$

The multiplicities $m 1$ and $m 2$ are integers. Consider the fraction

$$
\frac{-k^{2}+2 k}{\sqrt{4 k-3}}
$$

If $k=2$, the numerator is 0 . If $k \neq 2$, then $\sqrt{4 k-3}$ must be an integer, so that $4 k-3$ is a perfect square, say $4 k-3=s 2$. Then

$$
k=\frac{1}{4}\left(s^{2}+3\right),
$$

and

$$
-k^{2}+2 k=\frac{1}{16}\left(-s^{4}+2 s^{2}+15\right) .
$$

This expression must be divisible by $\sqrt{4 k-3}=s$. If $s$ does not divide 15 , it cannot be an integer, since the other 2 terms have no $s$ in the denominator. Therefore $s=1,3,5$, or 15 . The corresponding values of $k, m 1$, $m 2, a 1$, and $a 2$ are shown in the following table:

| $s$ | $k$ | $n$ | $m 1$ | $m 2$ | $a 1$ | $a 2$ |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| 1 | 1 | 2 | 1 | 0 | 0 | -1 |
| 3 | 3 | 10 | 4 | 5 | 1 | -2 |
| 5 | 7 | 50 | 21 | 28 | 2 | -3 |
| 15 | 57 | 3250 | 1520 | 1729 | 7 | -8 |

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The value $k=1$ does not correspond to a graph. $k=3$ gives the Petersen graph. There is a unique Moore graph with $k=7$ and $n=50$, called the Hoffman-Singleton graph. It is not known whether a Moore graph with $k=57$ and $n=3250$ exists. The 5 -cycle is a Moore graph with $k=2$. Its eigenvalues are

$$
\alpha_{1}=\frac{1}{2}(-1+\sqrt{5})
$$

and

$$
\alpha_{2}=\frac{1}{2}(-1-\sqrt{5}),
$$

with multiplicities $m 1=m 2=2$.
The diameter of a graph is the maximum distance between any two vertices,

$$
\operatorname{diam}(G)=\max \{\operatorname{DIST}(u, v): u, v \in V(G)\} .
$$

Thus, Moore graphs have diameter two.

## Exercises

3.3.1 Let $G$ be a Moore graph of degree $k$, with $n=k 2+1$ vertices. Let $u$ be any vertex of $G$. Prove that there are exactly

$$
\frac{k(k-1)^{2}}{2}
$$

pentagons containing $u$. Conclude that $G$ contains

$$
\frac{k\left(k^{2}+1\right)(k-1)^{2}}{10}
$$

pentagons, so that $k \not \equiv 4(\bmod 5)$.
3.3.2 Let $G$ be as above. Prove that every $v \in V(G)$ is contained in exactly

$$
\frac{k(k-1)^{2}(k-2)}{2}
$$

hexagons and in

$$
\frac{k(k-1)^{2}(k-2)(k-3)}{2}
$$

heptagons.
3.3.3 Show that in a $k$-regular graph of girth five, with $n=k 2+1$ vertices, the distance $\operatorname{DIST}(u, u)$ between any two vertices is at most two. Hint: Show that DIST $(u, u)=3$ implies the existence of a 4-cycle.
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### 3.4 Euler tours

Figure 3.12 shows a drawing of $K 5$ illustrating a walk in which each edge is covered exactly once.


FI GURE 3.12
A traversal of $K 5$
A walk which covers each edge of a graph $G$ exactly once is called an Euler trail in G. A closed Euler trail is called an Euler tour. The example above shows that $K 5$ has an Euler tour. Therefore we say that $K 5$ is Eulerian. It is easy to prove that a graph is Eulerian when all its degrees are even.
THEOREM 3.7 A connected graph $G$ has an Euler tour if and only if all degrees of $G$ are even.
PROOF Let $W$ be a closed Euler trail in $G$, beginning at vertex $u$. Each time that $W$ enters a vertex $u$, it also must exit it. Therefore $W$ uses an even number of edges at each vertex $u \neq u$. Since the trail is closed, the same is true of $u$. Since $W$ covers every edge of $G$ exactly once, all degrees must be even.
Conversly suppose that all degrees of $G$ are even. The proof is by induction on the number of edges of $G$. The smallest connected graphs with even degrees are $K 1$ and $K 3$, and both of these are Eulerian (for $K 1, W=\varnothing$ is an Euler trail). If the theorem is not true, let $G$ be the smallest graph (i.e., smallest $\varepsilon$ ) with even degrees with no closed Euler trail. Clearly $\delta(G) \geq 2$, so that $G$ contains a cycle, which is an Eulerian subgraph. Let $C$ be the largest Eulerian subgraph which $G$ contains. Then $\varepsilon(C)<\varepsilon(G)$. The complementary subgraph $G-C$ also has even

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degrees, and since it is smaller than $G$, each component of it must be Eulerian. Furthermore, $C$ intersects each component of $G-C$. We can now make an Euler trail in $G$ from $C$, by inserting into $C$ Euler trails of each component $K$ of $G-C$, as the walk in $C$ reaches each $K$ in turn. Therefore $G$ is Eulerian. By induction, all connected graphs with even degrees are Eulerian.
Notice that this theorem is true for multigraphs as well as simple graphs. If a connected graph $G$ has exactly two vertices, $u$ and $u$, of odd degree, then we can add an extra edge $u u$ to $G$ to get $G^{\prime}$, which will then have all even degrees. $G^{\prime}$ may now have multiple edges. If we now choose an Euler tour $W$ in $G^{\prime}$ beginning at $u$, we can number the edges so that the new edge $u u$ is the last edge traversed. Then $W$ - $u v$ will be an Euler trail in $G$ beginning at $u$ and ending at $u$. This is illustrated in Figure 3.13.


FI GURE 3.13


## An Euler trail

If there are more than two vertices of odd degree, then it is clear that $G$ cannot have an Euler trail.

### 3.4.1 An Euler tour algorithm

The proof of Theorem 3.7 is essentially an algorithm to find an Euler tour in a connected graph $G$. The algorithm works by building a walk from a starting vertex $u$. It takes the first edge $e 0=u v$ incident on $u$ and follows it. Then it takes the first edge el incident on $u$ and follows it, and so forth. Because the degrees are all even, it must eventually return to $u$. At this point, it will have found a closed walk in $G$ that is a sub-tour of an Euler tour. All the vertices visited in the

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sub-tour are stored on an array called the ScanQ. It then repeats this process at $u$, finding another sub-tour. The sub-tours are then linked together to create one sub-tour. It continues like this until all the edges at $u$ have been used up. It then moves to the next vertex on the ScanQ and builds a sub-tour there, always linking the sub-tours found into the existing tour. When the algorithm completes, all the vertices of $G$ are in the ScanQ array, because $G$ is connected. Therefore we have an Euler tour.
The Euler tour is stored as a linked list of edges. This makes it easy to insert a sub-tour into the list at any location. If $e=u u$ is an edge of $G$, then we write nextEdge $\langle e\rangle$ and prevEdge $\langle e\rangle$ for the next and previous edges in a tour, respectively. The adjacency list for vertex $u$ is denoted by Graph[u]. This is a linked list of incident edges.
When the algorithm begins building a sub-tour at vertex $u$, it needs to know an edge at $u$ currently in the Euler tour, if there is one. This is stored as EulerEdge[u]. It allows the algorithm to insert a sub-tour into the existing tour at that location in the linked list.
Algorithm 3.4.1 is very efficient. For each vertex $u$, all incident edges are considered. Each edge is linked into the Euler tour. This takes DEG( $u$ ) steps. Several sub-tours at $u$ may be linked into the Euler tour being constructed. There are at most DEG(u)/2 sub-tours at $u$. If follows that the complexity is determined by

$$
\sum_{u} \operatorname{DEG}(u)=2 \varepsilon(G)=O(\varepsilon) .
$$

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Algorithm 3.4.1: EULERTOUR(G)
comment: Construct an Euler tour in $G$
ScanQ[1] $\leftarrow 1$
QSize $\leftarrow 1$
$k \leftarrow 1$
while $k \leq$ QSize

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## Exercises

3.4.1 Program Algorithm 3.4.1 EULERTOUR(G).
3.4.2 Let $G$ be a connected graph in which $2 k$ of the vertices are of odd degree. Show that there are $k$ trails $W 1, W 2, \ldots, W k$ such that, taken together, $W 1, W 2, \ldots, W k$ cover each edge of $G$ exactly once.
3.4.3 Let $W$ be an Euler tour in $G$. To what subgraph of the line-graph $L(G)$, does $W$ correspond?
3.4.4 Show that any Euler tour of a graph $G$ can be written as a union of cycles.
3.4.5 What does the following algorithm do when input a connected graph $G$ ? What is its complexity? procedure TESTGRAPH(G)
ScanQ $[1] \leftarrow 1$
QSize $\leftarrow 1$
$\operatorname{Tag}[1] \leftarrow 1$
$k \leftarrow 1$
while $k \leq$ QSize
for all $v \longrightarrow u$
if $v \notin S c a n Q$
then $\left\{\begin{array}{l}\text { QSize } \leftarrow \text { QSize }+1 \\ \text { ScanQ }[\text { QSize }] \leftarrow v \\ \operatorname{Tag}[v] \leftarrow-\operatorname{Tag}[u]\end{array}\right.$
else if $\operatorname{Tag}[v]=\operatorname{Tag}[u]$
then return (false)
$k \leftarrow k+1$
The theorem on Moore graphs is due to HOFFMANN and SI NGLETON [63]. The application of algebraic
methods to graph theory is treated in BIGGS [9] and GODSIL and ROYLE [52]. The eigenvalues of graph
adjacency matrices is a vast topic. See the surveys by HOFFMAN [62], SCHWENK and WILSON [107], or the
book by CVETKOVIC, DOOB, and SACHS [30]. An excellent description of Euler tour algorithms can be found
in GOULD [53].
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4
Trees and Cycles

### 4.1 I ntroduction

A tree is a connected graph that has no cycles. Figure 4.1 shows a number of trees.

## FI GURE 4.1

## Several trees

Trees are the smallest connected graphs; remove any edge from a tree and it becomes disconnected. As well as being an important class of graphs, trees are important in computer science as data structures, and as
objects constructed by search algorithms. A fundamental property of trees is that all trees on $n$ vertices have the same number of edges.
THEOREM 4.1 If $G$ is a tree, then $\varepsilon(G)=|G|-1$.
PROOF The proof is by induction on $|G|$. If $|G|=1$, then $G=K 1$, which is a connected graph with no cycle, so that $\varepsilon=0$. Similarly, if $|G|=2$, then $G=K 2$, which has $\varepsilon=1$. Assume that the result is true whenever $|G| \leq t$, and page_63

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consider a tree $G$ with $|G|=t+1$. Now $G$ must have a vertex of degree one, or it would contain a cycle, so let $v \in V(G)$ have degree one. Then $G^{\prime}=G-u$ is still connected, and has no cycle, so it is a tree on $t$ vertices. Therefore $\varepsilon\left(G^{\prime}\right)=\left|G^{\prime}\right|-1=|G|-2$. It follows that $\varepsilon(G)=|G|-1$, so that the result is true when $|G|=t+1$. By induction, it holds for all values of $|G|$.
We saw in this proof that a tree $G$ with $\varepsilon>0$ must have a vertex of degree one. Consider a longest path $P$ in $G$. The two endpoints of $P$ can only be joined to vertices of $P$. Since $G$ does not contain any cycles, we can conclude that the endpoints of a longest path have degree one. Therefore a tree has at least two vertices of degree one.
In a connected graph, any two vertices are connected by some path. A fundamental property of trees is that any two vertices are connected by a unique path. For if there were two $u u$-paths $P$ and $Q$, where $P \neq Q$, then traveling from $u$ to $u$ on $P$ we could find the first point of $P$ which is not on $Q$. Continuing on $P$ until we come to the first point which is again on both $P$ and $Q$, we could now follow $Q$ back toward $u$ and so find a cycle, which, however, is not possible in a tree.
Every graph $G$ has subgraphs that are trees. The most important of these are the spanning trees, that is, trees which span all the vertices of $G$.

LEMMA 4.2 Every connected graph has a spanning tree.
PROOF Let $G$ be a connected graph. If $G$ has no cycles, then $G$ is a spanning tree. Otherwise choose a cycle $C$, and remove any edge $x y \in C$ from $G$. $G$ is still connected, since any uu-path which uses $x y$ can now be replaced by a path using $C-x y$, so that every $u$ and $u$ are still connected by some path after $x y$ has been removed. We repeat this as many times as necessary until the resulting graph has no cycles. It is a spanning tree of the original $G$.

## Exercises

4.1.1 Describe in pseudo-code an algorithm to find an edge on a cycle, if one exists.
4.1.2 Make a list of all isomorphism types of trees on $1,2,3,4,5$, and 6 vertices.
4.1.3 Show that there is a tree on $n$ vertices with degree sequence ( $d 1, d 2, \ldots, d n$ ) if and only if

$$
\begin{gathered}
\sum_{i=1}^{n} d_{i}=2(n-1) \\
\text { page_64 }
\end{gathered}
$$

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### 4.2 Fundamental cycles

Figure 4.2 shows a spanning tree of the Petersen graph. If $T$ is a spanning tree of $G$, let $G-T$ stand for the graph whose edges are $E(G)-E(T)$. Notice that if any edge $x y \in E(G-T)$ is added to $T$, then $T+x y$ contains a unique cycle Cxy. This is because $x$ and $y$ are connected by a unique path Pxy in $T$. Pxy+xy creates a cycle, Cxy, called the fundamental cycle of $x y$ with respect to $T$.


FI GURE 4.2
A spanning tree

## Exercises

4.2.1 Show that $G$ has $\varepsilon-\mid G+1$ fundamental cycles with respect to any spanning tree $T$.
4.2.2 Let $T$ be a spanning tree of $G$, and let $C$ be a cycle of $G$ containing exactly two edges $x y$ and $u u$ of $G-T$. Prove that $C=C_{x y} \oplus C_{u v}$, where $\oplus$ denotes the operation of exclusive OR.
Every cycle of $G$ can be formed from the fundamental cycles of $G$ with respect to any spanning tree $T$.
THEOREM 4.3 Let $T$ be a spanning tree of $G$. Let $C$ be any cycle containing $k$ edges $u 1 u 1, u 2 u 2, . .$, ukuk of $G-T$, where $k \geq 1$. Then $C=C_{u_{1} v_{1}} \oplus C_{u_{2} v_{2}} \oplus \cdots \oplus C_{u_{k} v_{k}}$.
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PROOF The proof is by induction on $k$. The result is certainly true when $k=1$. Suppose that the edges uivi occur on $C$ in the order $i=1,2, \ldots, k$. These edges divide the remaining edges of $C$ into a number of paths $P 1$, $P 2, \ldots, P k$, where $P i$ connects ui to $u i+1$. This is shown in Figure 4.3.


FI GURE 4.3

## Decomposition into fundamental cycles

Let $C_{i}=C_{u_{i} v_{i}}$ denote the ith fundamental cycle. Consider $C 1$. It consists of the edge $u 1 u 1$ and the unique path $P$ of $T$ connecting $u 1$ to $u 1 . P$ and $P 1$ both begin at vertex $u 1$. As we travel on $P$ from $u 1$ toward $u 1$, we eventually come to the first vertex of $P$ which is not on $P 1$. This is also the last vertex in common with $P 1$, because $P$ and $P 1$ are both contained in $T$, which has no cycles. $P$ may intersect several of the paths Pi. In each case the intersection must consist of a single segment, that is, a consecutive sequence of vertices of Pi , since $T$ contains no cycles. The last path which $P$ intersects is $P k$, since both $P$ and $P k$ end with $u 1$. This is illustrated in Fiqures 4.3 and 4.4.
Consider now $H=C_{1} \oplus C$. It is a subgraph of $G$. It consists of that part of $C$ which is not contained in $C 1$, plus that part of $P$ which is not contained in $C$. Thus, the portions common to $P$ and each $P i$ are discarded, but the new segments of $P$ which are now added create one or more new cycles. Thus, $H$ consists of one or more edge-disjoint cycles constructed from edges of $T$, plus the edges $u 2 u 2, u 3 u 3, \ldots$, ukuk. Since each of these cycles contains fewer than $k$
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## FI GURE 4.4

## Decomposition into fundamental cycles

edges of $G-T$, we can say that $H=C_{u_{2} v_{2}} \oplus C_{u_{3} v_{3}} \oplus \cdots \oplus C_{u_{k} v_{k}}$. We then have
$C_{1} \oplus H=\left(C_{1} \oplus C_{1}\right) \oplus C=C=C_{u_{1} v_{1}} \oplus C_{u_{2} v_{2}} \oplus \cdots \oplus C_{u_{k} v_{k}}$. Therefore the result is true when $C$ contains $k$ edges of $G-T$. By induction the result is true for all values of $k$.
Thus the fundamental cycles of $G$ with respect to any spanning tree $T$ generate all the cycles of $G$.

### 4.3 Co-trees and bonds

Let $G$ be a graph. If $S \subset V(G)$, then $\bar{S}$ denotes $V(G)-S$. The edge-cut $[S, \bar{S}]$ consists of all edges of $G$ with one endpoint in $S$ and one endpoint in $\bar{S}$. Notice that $G-[S, \bar{S}]$ is a disconnected graph. See Figure 4.5.
If $T$ is a spanning tree of $G$, then the complementary graph $\widehat{T}=G-T$ is called the co-tree corresponding to $T$. Now a co-tree $\widehat{T}$ cannot contain any edge-cut of

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FI GURE 4.5

## An edge-cut

$G$. This is because $G-\widehat{T}=T$ which is connected. If $u u$ is any edge of $T$, then $T$ - $u \boldsymbol{u}$ consists of two components, Su, those vertices connected to $u$, and $S u$, those vertices connected to $u$. [ $S u, S u$ ] is an edgecut of $G$. It is contained in $\widehat{T}+u v$. This is illustrated in Figure 4.6.
[Su, Su] is a minimal edge-cut of $G$, that is, it does not contain any smaller edge-cuts. For if $x y$ is any edge, where $x \in S_{u}$ and $y \in S_{v}$, then $G-[S u, S U]+x y$ is connected. Therefore:

1. A co-tree $\widehat{T}$ contains no edge-cut.
2. If $u v$ is any edge of $G-\widehat{T}$, then $\widehat{T}+u v$ contains a unique minimal edge-cut $B u u=[S u, S u]$. Compare this with trees:
3. A tree $T$ contains no cycle.
4. If $u v$ is any edge of $G-T$, then $T+u u$ contains a unique fundamental cycle Cuu.

The unique edge-cut Buu contained in $\widehat{T}+u v$ is called the fundamental edge-cut of $u u$ with respect to $\widehat{T}$. Any minimal edge-cut of $G$ is called a bond. There is a duality between trees and co-trees, and cycles and bonds (bonds are sometimes called co-cycles). There is a linear algebra associated with every graph, in which cycles and bonds generate orthogonal vector spaces, called the cycle space and bond space of $G$. Theorem 4.3 above shows that the fundamental cycles with


FI GURE 4.6

## Co-trees and edge-cuts

respect to any spanning tree form a basis for the cycle space. Similarly, the fundamental edge-cuts form a basis for the bond space. See BONDY and MURTY [19] for more information.

## Exercises

4.3.1 How may fundamental edge-cuts does $G$ have with respect to a co-tree $\widehat{T}$ ? What are the dimensions of the cycle space and bond space of $G$ ?
4.3.2 Let $T$ be a spanning tree of $G$, and let edge $u v \notin T$. Let $x y$ be any edge of the fundamental cycle Cuu, such that $x y \neq u u$. Then $T+u u-x y$ is also a spanning tree of $G$. Thus, spanning trees of $G$ are adjacent via fundamental cycles. The tree graph of $G$ is $\operatorname{Tree}(G)$. Its vertices are the spanning trees of $G$, and they are adjacent via fundamental cycles. Show that $\operatorname{Tree}(G)$ is a connected graph.
4.3.3 Show that trees (co-trees) are also adjacent via fundamental edge-cuts.
4.3.4 Let $T$ be a spanning tree of $G$, and let $W$ be a closed walk in $G$ such that $W$ uses edges of $T$ and edges u1 U1, u2u2, $\ldots$, ukuk of $G-T$. Describe the subgraph $H=C_{u_{1} v_{1}} \oplus C_{u_{2} v_{2}} \oplus \cdots \oplus C_{u_{k} v_{k}}$. What is its relation to $W$ ?
4.3.5 Let $[S, \bar{S}]$ be an edge-cut of $G$. Prove that $[S, \bar{S}]$ is a bond if and only if $G[S]$ and $G[\bar{S}]$ are connected graphs.
4.3.6 Let $B_{1}=\left[S_{1}, \bar{S}_{1}\right]$ be an edge-cut of $G$, and let $B_{2}=\left[S_{2}, \bar{S}_{2}\right]$ be a bond contained in $\left[S_{1}, \bar{S}_{1}\right]$; that is, $B_{1} \subset B_{2}$. (Note: $S 2$ will generally not be a subset of $S 1$.) Prove

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that $\left[S_{1}, \bar{S}_{1}\right]-\left[S_{2}, \bar{S}_{2}\right]$ is also an edge-cut.
4.3.7 Use the previous question to prove that every edge-cut can be decomposed into a disjoint union of bonds.
4.3.8 Find a decomposition of the edge-cut $[S, \bar{S}]$ in the graph shown in Figure 4.7 into bonds. The set $S$ is marked by the shading. Is the decomposition unique? (Hint: Redraw the graph so that edges don't cross each other.)


FI GURE 4.7
Find a decomposition into bonds
4.3.9 Let $T$ be a spanning tree of $G$. Let $u v$ and $x y$ be edges of $T$, with corresponding bonds Buu=[Su,Su] and
$B x y=[S x, S y]$, where $B u u \cap B x y \neq \emptyset$. Prove that $B_{u v} \oplus B_{x y}$ is a bond.
4.3.10 Prove that any cycle and any bond must intersect in an even number of edges.

### 4.4 Spanning tree algorithms

One of the easiest ways to construct a spanning tree of a graph $G$ is to use a breadth-first search. The following code is adapted from Algorithm 2.4.1. The statements marked with ( $\star$ ) have been added.
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Algorithm 4.4.1: $\operatorname{BFSEARCH}(G, u)$
comment: build a breadth-first spanning tree of $G$
for $U \leftarrow 1$ to $|G|$ do OnScan $Q[U] \leftarrow$ false
ScanQ $[1] \leftarrow u$
OnScanQ[u] זtrue QSize $\leftarrow 1$ $k \leftarrow 1$
Parent[u] $\leftarrow 0 \star$ BFNum[u] $\leftarrow$ 1* Count $\leftarrow 1 \star$
Tree_empty list* repeat
$U \leftarrow S c a n Q[k]$
for each $w \longrightarrow v$
do if not OnScanQ[w]

```
        QSize \(\leftarrow\) QSize +1
        ScanQ[QSize \(] \leftarrow w\)
        OnScanQ \([w] \leftarrow\) true
then \(\{\operatorname{Parent}[w] \leftarrow v\)
Count \(\leftarrow\) Count +1
Count \(\leftarrow\) Count +1 (*)
BFNum \([w] \leftarrow\) Count
add edge \(v w\) to Tree
until \(k>Q\) Size
A number of arrays ScanQ, OnScanQ, Parent, and BFNum are used in this algorithm, as well as the counters QSize and Count, and the list of edges Tree of the spanning tree constructed.
\(\operatorname{BFSEARCH}(G, u)\) visits each vertex of the connected graph \(G\), beginning with \(u\). The order in which the vertices are visited defines a numbering of the vertices, called the breadth-first numbering. It is saved in BFNum[•]. This is illustrated in Figure 4.8.
The search begins at node \(u\), called the root of the spanning tree. In the example above, \(u=1\). As each node w is placed on the ScanQ, its parent in the search tree is saved in Parent[w]. This is represented by the arrows on the tree in the diagram. Thus, beginning at any node in the graph, we can follow the Parent[•] values up to the root of the tree. The breadth-first numbering defines a traversal of the tree, which goes level by level, and from left to right in the drawing. A great many graph algorithms are built around the breadthfirst search. The important property of breadth-first spanning trees is that the paths it constructs connecting page_71
\(u=1\)



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\section*{FI GURE 4.8}

A breadth-first tree
any vertex \(w\) to the root of the tree are shortest paths.
In a weighted graph, different spanning trees will have different weight, where
\[
\mathrm{WT}(T)=\sum_{u v \in T} \mathrm{WT}(u v)
\]

We now want to find a spanning tree \(T\) of minimum weight. This is called the minimum spanning tree problem. There are many algorithms which solve it. We present some of them here.
4.4.1 Prim's algorithm

The idea here is to pick any \(u \in V(G)\) and "grow" a tree on it; that is, at each iteration, we add one more edge to the current tree, until it spans all of \(V(G)\). We must do this in such a way that the resulting tree is minimum.
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Algorithm 4.4.2: PRIM(G)
comment: \(\left\{\begin{array}{l}\text { Tree is a list of edges in a minimum spanning tree. } \\ V T \text { are the vertices in the current tree being grown. }\end{array}\right.\) initialize Tree to contain no edges \(t \leftarrow 0\) "the number of edges in Tree"
\[
\text { choose any } u \in V(G)
\]
initialize \(V T\) to contain \(u\)
comment: the Tree now has 1 node and 0 edges
while \(t<|G|-1\)
do \(\left\{\begin{array}{l}\text { choose an edge } x y \text { of minimum weight, with } x \in V T \text { and } y \notin V T \\ \text { add } x y \text { to Tree } \\ \text { add } y \text { to } V T \\ t \leftarrow t+1\end{array}\right.\)

\(V T\)
\(\overline{V T}\)

\section*{Growing a tree with Prim's algorithm}

We first prove that Prim's algorithm does in fact produce a minimum spanning tree. Initially, \(V T\) contains one vertex, and Tree contains no edges. On each iteration an edge \(x y\) with \(x \in V T\) and \(y \notin V T\) is added to Tree, and \(y\) is added to \(V T\). Therefore, the edges of Tree always form a tree which spans \(V T\). After \(n-1\) iterations, it is a spanning tree of \(G\). Call the tree produced by Prim's algorithm \(T\), and suppose that it consists of edges el, e2,..., en-1, chosen in that order. If it is not a minimum spanning tree, then choose a minimum tree \(T^{*}\) which agrees with \(T\) on the first \(k\) iterations, but not on iteration \(k+1\), where \(k\) is as large as possible. Then el, e2,... \(e_{k} \in T^{*}\), but \(e_{k+1} \notin T^{*}\). Consider iteration
page
Page 74
\(k+1\), and let \(e k+1=x y\), where \(x \in V T\) and \(y \in \overline{V T}\). Then \(T^{*}+x y\) contains a fundamental cycle \(C x y\).


\section*{VT}

\section*{FI GURE 4.10 \\ A fundamental cycle in Prim's algorithm}

Cxy must contain another edge \(u v\) with \(u \in V T\) and \(v \in V T\). Since Prim's algorithm chooses edges by
weight, we know that \(\mathrm{WT}(x y) \leq W T(u v)\). Now \(T^{\prime}=T^{*}+x y-u v\) is also a spanning tree, and \(W T\left(T^{\prime}\right) \geq W T\left(T^{*}\right)\), since \(T^{*}\) is a minimum tree. But \(\mathrm{WT}\left(T^{\prime}\right)=\mathrm{W} T\left(T^{*}\right)+\mathrm{W} T(x y)-\mathrm{WT}(u v)=\mathrm{W} T\left(T^{*}\right)\). Therefore, \(\mathrm{WT}\left(T^{\prime}\right)=\mathrm{W} T\left(T^{*}\right)\) and \(\mathrm{WT}(x y)=\mathrm{WT}(u v)\). It follows that \(T^{\prime}\) is also a minimum tree, and that \(T^{\prime}\) contains el, e2,..., ek+1; that is, it agrees with \(T\) on \(k+1\) iterations, a contradiction. Consequently, Prim's tree \(T\) is also a minimum spanning tree. Data structures
The main operation performed in Prim's algorithm is to select the edge \(x y\), of minimum weight, with \(x \in V T\) and \(y \in \overline{V T}\). One way to do this is to store two values for each vertex \(y \in \overline{V T}\) :
\(\operatorname{MinWt}[y]\) : the minimum weight \(\mathrm{WT}(x y)\), over all \(x \longrightarrow y\), where \(x \in V T\)
MinPt[y]: that vertex \(x \in V T\) with \(\mathrm{WT}(x y)=\operatorname{MinWt}[y]\).
Then to select the minimum edge \(x y\), we need only perform the following steps: select \(y \in \overline{V T}\) with smallest \(\operatorname{MinWt}[y]\) value \(x \leftarrow M i n P t[y]\)
\[
\text { for each } w \longrightarrow y \text { do }\left\{\begin{array}{l}
\text { if } w \in \overline{V T} \\
\text { then update } \operatorname{MinWt}[w]
\end{array}\right.
\]
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Let \(n=|G|\). If we scan the set \(\overline{V T}\) in order to select the minimum vertex \(y\) on each iteration, then the first iteration requires scanning \(n-1\) vertices, the second iteration requires \(n-2\) steps, etc., requiring \(1+2+\cdots+(n-1)=\binom{n}{2}\) in total. The total number of steps needed to update the MinWt[•] values is at most \(\Sigma y \operatorname{DEG}(y)=2 \varepsilon\) steps, over all iterations. Thus, the complexity when Prim's algorithm is programmed like this is
\[
O\left(2 \varepsilon+\frac{n^{2}}{2}\right)=O\left(\varepsilon+n^{2}\right)
\]

In order to remove the \(O(n 2)\) term, we could store the vertices \(\overline{V T}\) in a heap \(H\). Selecting the minimum now requires approximately \(\log n\) steps. For each \(w \longrightarrow y\) we may also have to update \(H\), requiring at most an additional DEG(y) log \(n\) steps per iteration. The total number of steps performed over all iterations is now at most
\[
\sum_{k=1}^{n-1} \log n+\sum_{y} \operatorname{DEG}(y) \log n \leq n \log n+2 \varepsilon \log n
\]

The complexity of Prim's algorithm using a heap is therefore \(O(n \log n+\varepsilon \log n)\). If \(\varepsilon\) is small, this will be better than the previous method. But if \(\varepsilon\) is large, this can be worse, depending on how much time is actually spent updating the heap. Thus we can say that Prim's algorithm has complexity:
\(O(\varepsilon+n 2)\), if the minimum is found by scanning.
\(O(n \log n+\varepsilon \log n)\), if a heap is used.

\section*{Exercises}
4.4.1 Work Prim's algorithm by hand on the graph in Figure 4.11, starting at the shaded vertex.
4.4.2 Consider Dijkstra's shortest-path algorithm, which finds a shortest uu-path for all \(v \in V(G)\). For each \(u\), let \(P u\) be the shortest path found. Show that the collection of paths, \(\bigcup_{v} P_{v}\), defines a spanning tree of \(G\). Is it a minimum spanning tree? ( Hint: Use induction.)
4.4.3 Program Prim's algorithm, storing the vertices \(\overline{V T}\) in a heap.
4.4.4 Modify the breadth-first search algorithm to find the fundamental cycles of \(G\) with respect to a BF-tree.

Print out the edges on each fundamental cycle. What is the complexity of the algorithm?
4.4.5 Let \(G\) be a weighted graph in which all edge-weights are distinct. Prove that \(G\) has a unique minimum spanning tree.

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FI GURE 4.11

\section*{Find a spanning tree}

\subsection*{4.4.2 Kruskal's algorithm}

A forest is a graph which need not be connected, but whose every component is a tree. Prim's algorithm constructs a spanning tree by growing a tree from some initial vertex. Kruskal's algorithm is quite similar, but it begins with a spanning forest and adds edges until it becomes connected. Initially the forest has \(n=|G|\) components and no edges. Each component is a single vertex. On each iteration, an edge which connects two distinct components is added, and the two components are merged. When the algorithm terminates the forest has become a tree.

Algorithm 4.4.3: KRUSKAL(G)
comment: \(\left\{\begin{array}{l}\text { Tree is a list of edges in a minimum spanning tree. } \\ T_{u} \text { is the component of the forest which contains } u .\end{array}\right.\)
initialize Tree to contain no edges
for each \(u \in V(G)\) do initialize Tu to contain only \(u\) \(t \leftarrow 0\) "the number of edges in Tree"
comment: the forest currently has \(|G|\) nodes and 0 edges
while \(t<|G|-1\)

Select the next edge \(x y\) of minimum weight, and determine which components \(x\) and \(y\) are in, say \(x \in T_{u}\) and \(y \in T_{v}\)
if \(T_{u} \neq T_{v}\)
\[
\text { then }\left\{\begin{array}{l}
\text { merge } T_{u} \text { and } T_{v} \\
\text { add } x y \text { to Tree } \\
t \leftarrow t+1
\end{array}\right.
\]
page_76
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I nitially the forest has \(n\) components, and each one is a tree with no edges. Each edge that is added connects two distinct components, so that a cycle is never created. Whenever an edge is added the two components are merged, so that the number of components decreases by one. After \(n-1\) iterations, there is only one component left, which must be a tree \(T\). If \(T\) is not a minimum tree, then we can proceed as we did in Prim's algorithm. Let \(T\) consist of edges el, e2,.., en-1, chosen in that order. Select a minimum tree \(T^{*}\) which contains el, e2,.., ek, but not ek+1, where \(k\) is as large as possible. Consider the iteration in which ek+1= \(x y\) was selected. \(T^{*}+x y\) contains a fundamental cycle Cxy, which must contain another edge ab incident on \(T x\). Since Kruskal's algorithm chooses edges in order of their weight, \(\mathrm{WT}(x y) \leq \mathrm{WT}(a b)\). Then \(T^{\prime}=T^{*}+x y-a b\) is a spanning tree for which \(\mathrm{WT}\left(T^{\prime}\right) \leq \mathrm{W} T\left(T^{*}\right)\). But \(T^{*}\) is a minimum tree, so that \(\mathrm{WT}\left(T^{\prime}\right)=\mathrm{WT}\left(T^{*}\right)\), and \(T^{\prime}\) is also a minimum tree. \(T^{\prime}\) contains edges el, e2,..., ek+1, a contradiction. Therefore, Kruskal's tree \(T\) is a minimum spanning tree.


\section*{FI GURE 4.12}

\section*{Growing a forest with Kruskal's algorithm Data structures and complexity}

The main operations in Kruskal's algorithm are:
1. Choose the next edge \(x y\) of minimum weight.
2. Determine that \(x \in T_{u}\) and \(y \in T_{v}\).
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3. Merge Tu and Tu.

The edges could either be completely sorted by weight, which can be done in \(O(\varepsilon \log \varepsilon)\) steps, or they could be kept in a heap, which makes it easy to find the minimum edge. Since we may have a spanning tree \(T\) before all the edges have been considered, it is usually better to use a heap. The components Tu can easily be stored using the merge-find data structure described in Chapter 2.
Each time an edge \(x y\) is selected from the heap, it requires approximately \(\log \varepsilon\) steps to update the heap. In the worse case we may need to consider every edge of \(G\), giving a bound of \(\varepsilon \log \varepsilon\) steps. Similarly, \(O(\varepsilon a(n))\) steps are needed to build the components, where \(n=|G|\). Thus, Kruskal's algorithm can be programmed with a complexity of \(O(\varepsilon \log n+\varepsilon a(n))\), where we have used \(\log \varepsilon<2 \log n\). Notice that this can be slightly better than Prim's algorithm. This is because the term \(a(n)\) is essentially a constant, and because the heap does not need to be constantly updated as the MinWt[•] value changes.

\subsection*{4.4.3 The Cheriton-Tarjan algorithm}

The Cheriton-Tarjan algorithm is a modification of Kruskal's algorithm designed to reduce the \(O(\varepsilon \log \varepsilon)\) term.

It also grows a spanning forest, beginning with a forest of \(n=|G|\) components each consisting of a single node. Now the term \(O(\varepsilon \log \varepsilon)\) comes from selecting the minimum edge from a heap of \(\varepsilon\) edges. Since every component Tu must eventually be connected to another component, this algorithm keeps a separate heap PQu for each component \(T u\), so that initially \(n\) smaller heaps are used. Initially, PQu will contain only DEG(u) edges, since Tu consists only of vertex \(u\). When Tu and Tu are merged, \(P Q u\) and \(P Q u\) must also be merged. This requires a modification of the data structures, since heaps cannot be merged efficiently. This is essentially because merging heaps reduces to building a new heap. Any data structure in which a minimum element can be found efficiently is called a priority queue. A heap is one form of priority queue, in which elements are stored as an array, but viewed as a binary tree. There are many other forms of priority queue. In this algorithm, PQu will stand for a priority queue which can be merged. The Cheriton-Tarjan algorithm can be described as follows.
It stores a list Tree of the edges of a minimum spanning tree. The components of the spanning forest are represented as Tu and the priority queue of edges incident on vertices of \(T u\) is stored as \(P Q u\).
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Algorithm 4.4.4: CHERITONTARJ AN(G)
initialize Tree to contain no edges for each \(u \in V(G)\)
initialize \(T_{u}\) to contain only \(u\) create \(P Q_{u}\)
for each \(v \longrightarrow u\)
do add \(u v\) to \(P Q_{u}\)
comment: each edge will appear in two priority queue
comment: the forest currently has \(\mid G\) nodes and 0 edges
\(t \leftarrow 0\)
while \(t<|G|-1\)
select a component \(T_{u}\)
repeat
select the minimum edge \(x y \in P Q_{u}\) and determine which components \(x\) and \(y\) are in, say \(x \in T_{u}\) and \(y \in T_{v}\)
until \(T_{u} \neq T_{v}\)
do
comment: \(x y\) connects two different components
merge \(T_{u}\) and \(T_{v}\)
merge \(P Q_{u}\) and \(P Q_{v}\)
add \(x y\) to Tree
\(t \leftarrow t+1\)

\section*{Exercises}
4.4.1 Prove that the Cheriton-Tarjan algorithm constructs a minimum spanning tree.
4.4.2 Show that a heap is best stored as an array. What goes wrong when the attempt is made to store a heap with pointers?
4.4.3 Show that heaps cannot be merged efficiently. Describe an algorithm to merge two heaps, both with \(n\) nodes, and work out its complexity.
4.4.4 Program Kruskal's algorithm, using a heap to store the edges, and the merge-find data structure to store the components.

\subsection*{4.4.4 Leftist binary trees}

A leftist binary tree (LB-tree) is a modification of a heap which allows efficient merging. A node \(x\) in an LBtree has the following four fields:
1. Value \(\langle x\rangle\) : the value stored.
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2. Left \(\langle x\rangle\) : a pointer to the left subtree.
3. Right \(\langle x\rangle\) : a pointer to the right subtree.
4. \({ }^{\text {rPath }}\langle x\rangle\) : the right-path distance.

An LB-tree satisfies the heap property; namely the entry stored in any node has value less than or equal to that of its two children:
\[
\begin{gathered}
\operatorname{Value}\langle x\rangle \leq \operatorname{Value}\langle\operatorname{Left}\langle x\rangle\rangle \\
\operatorname{Value}\langle x\rangle \leq \operatorname{Value}\langle\operatorname{Right}\langle x\rangle\rangle
\end{gathered}
\]

Therefore the smallest entry in the tree occurs in the top node. Thus, a heap is a special case of an LB-tree. The distinctive feature of LB-trees is contained in field \(r\) Path \([x]\). If we begin at any node in an LB-tree and follow Left and Right pointers in any sequence, we eventually reach a nil pointer. In an LB-tree, the shortest such path is always the rightmost path. This is true for every node in the tree. The length of the path for a node \(x\) is the \(r\) Path \(\langle x\rangle\) value. In the tree shown in Figure 4.13, the rPath values are shown beside each node.


FI GURE 4.13

\section*{A leftist binary tree}

In summary, an LB-tree is a binary tree which satisfies the heap property, and whose shortest path to a nil pointer from any node is always the rightmost path.
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This means that LB-trees will tend to have more nodes on the left than the right; hence the name leftist binary trees.
The rightmost path property makes it possible to merge LB-trees efficiently. Consider the two trees \(A\) and \(B\) in Figure 4.14 which are to be merged into a tree \(T\).


\section*{Fl GORE 4.14}

\section*{Merging two leftist binary trees}

The top node of \(T\) is evidently taken from \(A\), since it has the smaller minimum. This breaks \(A\) into two subtrees, \(L\) and \(R\). The three trees \(B, L\), and \(R\) are now to be made into two subtrees of \(T\). The easiest way to do this is first to merge \(R\) and \(B\) into a single tree \(P\), and then take \(P\) and \(L\) as the new right and left subtrees of \(T\), placing the one with the smaller \(r\) Path value on the right. The recursive merge procedure is described in Algorithm 4.4.5, and the result of merging \(A\) and \(B\) of Figure 4.14 is shown in Figure 4.15.

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Algorithm 4.4.5: \(\operatorname{LBMERGE}(A, B)\)
comment: Merge non-null LB-trees \(A\) and \(B\) if Value \(\langle A\rangle>\) Value \(\langle B\rangle\) then swap \(A\) and \(B\) if \(\operatorname{Right}\langle A\rangle=\) null then \(P \leftarrow B\)
else \(P \leftarrow\) LBMERGE \((\operatorname{Right}\langle A\rangle\) B)
comment: choose the tree with the smaller \(r\) Bath as right subtree
if \(L e f t\langle A\rangle=\) null

return (A)
Notice that when the top node of \(A\) is removed, thereby splitting \(A\) into two subtrees \(L\) and \(R\), the left subtree \(L\) subsequently becomes one of the subtrees of \(T\). That is, \(L\) is not decomposed in any way, it is simply
transferred to \(T\). Furthermore, \(L\) is usually the larger of the two subtrees of \(A\). Let us estimate the number of steps necessary to merge \(A\) and \(B\), with \(r\) Path values \(r 1\) and \(r 2\), respectively. One step is needed to choose the smaller node \(A\), say, as the new top node. The right subtree \(R\) will have rightmost path length of \(r 1-1\). When \(R\) and \(B\) are merged, one of them will be similarly decomposed into a left and right subtree. The left subtree is never broken down. At each step in the recursion, the smaller value is chosen as the new top node, and its Right becomes the next subtree to be considered; that is, LBMERGE() follows the rightmost paths of \(A\) and \(B\), always choosing the smaller entry of the two paths. Thus, the rightmost paths of \(A\) and \(B\) are merged into a single path (see Figure 4.15). Therefore, the depth of the recursion is at most \(r 1+r 2\). At the bottom of the recursion the rPath values may both equal zero. It then takes about five steps to merge the two trees. Returning up through the recursion, LBMERGE() compares the rPath values of \(L\) and \(P\), and makes the smaller one into the new right subtree. Also, the new rPath value is assigned. All this takes about four steps per level of recursion, so that the total number of steps is at most \(5(r 1+r 2+1)\).
What is the relation between the rightmost path length of an LB-tree and the number of nodes it contains? If the rPath value of an LB-tree \(T\) is \(r\), then beginning at the top node, every path to a nil pointer has length at least \(r\). Therefore, \(T\) contains at least a full binary tree of \(r\) levels; that is, \(T\) has at least \(2(r+1)-1\)

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\section*{FI GURE 4.15}

\section*{The merged LB-tree}
nodes. The largest rightmost path length possible if \(T\) is to store \(n\) nodes is the smallest value of \(r\) such that \(2(r+1)-1 \geq n\), or \(r \leq\left\lceil\log \frac{n+1}{2}\right\rceil\).
If \(A\) and \(B\) both contain at most \(n\) nodes, then \(r_{1}, r_{2} \leq\left\lceil\log \frac{n+1}{2}\right\rceil\), and LBMERGE(A, B) takes at most \(5(r 1+r 2+1) \leq 10 .\left\lceil\log \frac{n+1}{2}\right\rceil+5=O(\log n)\) steps. Thus LB-trees can be merged quite efficiently. We can use this same method to extract the minimum entry from an LB-tree \(A\), using at most \(O(\log n)\) steps: select minimum as Value \(\langle A\rangle\)
\(A \leftarrow L B M E R G E(L e f t\langle A\rangle, \operatorname{Right}\langle A\rangle)\)
Consider now how to construct an LB-tree. In Chapter 2 we found that there are two ways of building a heap, one much more efficient than the other. A similar situation holds for LB-trees. The most obvious way to build one is to merge successively each new node into an existing tree:
page_83
get next value
create and initialize a new LB-tree, \(B\)
\(A \leftarrow\) LBMERGE \((A, B)\)
until all values have been inserted
However, this can easily create LB-trees that are really linear linked lists, as shown below. This algorithm then becomes an insertion sort, taking \(O(n 2)\) steps, where \(n\) is the number of nodes inserted.

\section*{FI GURE 4.16}

\section*{An LB-tree}

A better method is to create \(n\) LB-trees, each containing only one node, and then merge them two at a time, until only one tree remains. The trees are kept on a queue, called the MergeQ.
Algorithm 4.4.6: BUILDLBTREE (MergeQ)
repeat
select \(A\) and \(B\), the first two trees of MergeQ
\(A \leftarrow \operatorname{LBMERGE}(A, B)\)
put \(A\) at end of MergeQ
until MergeQ contains only one tree

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How many steps are needed to build an LB-tree in this way, if we begin with \(n\) trees of one node each? There will be \(\lfloor n / 2\rfloor\) merges of pairs 1-node trees, each of which takes at most 5 steps. This will leave \(\lceil n / 2\rceil\) trees on the Merge \(Q\), each with at most two nodes. These will be taken two at a time, giving \(\lfloor n / 4\rfloor\) merges of up to 2node trees. Similarly there will be \(\lfloor n / 8\rfloor\) merges of up to 4 -node trees, etc. This is summarized in the following table:
\begin{tabular}{clll} 
tree size & \# pairs & max rPath & max \(r 1+r 2+1\) \\
1 & \(\lfloor n / 2\rfloor\) & 0 & 1 \\
2 & \(\lfloor n / 4\rfloor\) & 0 & 1 \\
4 & \(\lfloor n / 8\rfloor\) & 1 & 3 \\
8 & \(\lfloor n / 16\rfloor\) & 2 & 4 \\
\(\vdots\) & \(\vdots\) & \(\vdots\) & \(\vdots\) \\
\(2 k\) & \(\left\lfloor n / 2^{k+1}\right\rfloor\) & \(k-1\) & \(2 k-1\)
\end{tabular}

The last step will merge two trees with roughly \(n / 2\) nodes each. The maximum \(r\) Path value for these trees will be \(\leq\left\lceil\log \frac{n / 2+1}{2}\right\rceil\), or approximately \(\lceil\log n\rceil-1\). The total number of steps taken to build the LB-tree is then at most
\[
5\left\lfloor\frac{n}{2}\right\rfloor+\sum_{k=1}^{\lfloor\log n\rfloor-1} 5(2 k-1)\left\lfloor\frac{n}{2^{k+1}}\right\rfloor \leq \frac{5 n}{2}+5 n \sum_{k=1}^{\lfloor\log n\rfloor-1} \frac{2 k-1}{2^{k+1}}
\]

We can sum this using the same technique as in the heap analysis of Chapter 2, giving a sum of \(10 n-\frac{5 n r}{2^{r}}\), where \(r=\lfloor\log n\rfloor-1\). Thus, an LB-tree can be built in \(O(n)\) steps.
We can now fill in the details of the Cheriton-Tarjan spanning tree algorithm. There are three different kinds of trees involved in the algorithm:
1. A minimum spanning tree is being constructed.
2. The components \(T u\) are merge-find trees.
3. The priority queues PQu are LB-trees.

At the beginning of each iteration, a component \(T u\) is selected, and the minimum edge \(x y \in P Q_{u}\) is chosen. How is Tu selected? There are several possible strategies. If we choose the same Tu on each iteration, then the algorithm grows a tree from \(u\); that is, it reduces to Prim's algorithm. If we choose the component \(T u\) incident on the minimum remaining edge, then the algorithm reduces to Kruskal's algorithm. We could choose the smallest component \(T u\), but this would add an extra level of complication, since we would now have to keep a heap of components in order to find the smallest component quickly. The method which Cheriton and Tarjan recommend is uniform selection; that is, we keep a queue, TreeQ, of components. Each entry on the TreeQ contains Tu and PQu. On each
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iteration, the component \(T u\) at the head of the queue is selected and the minimum \(x y \in P Q_{u}\) is chosen, say \(x \in T_{u}\) and \(y \in T_{v}\), where \(T u \neq T u\). Once \(T u\) and \(T u, P Q u\) and \(P Q u\) have been merged, they are moved to the end of the TreeQ. Thus, the smaller components will tend to be at the head of the queue. So the algorithm uses two queues, the MergeQ for constructing LB-trees and merging them, and the TreeQ for selecting components.
The complexity analysis of the Cheriton-Tarjan algorithm is beyond the scope of this book. If analyzed very carefully, it can be shown to be \(O(\varepsilon \log \log \varepsilon)\), if programmed in a very special way.
Minimum spanning tree algorithms are a good illustration of the process of algorithm development. We begin with a simple algorithm, like growing a spanning tree from an initial vertex, and find a complexity of \(O(n 2)\). We then look for a data structure or programming technique that will allow the \(n 2\) term to be reduced, and obtain a new algorithm, with complexity \(O(\varepsilon \log n)\), say. We then ask how the \(\varepsilon\) or \(\log n\) term can be reduced, and with much more effort and more sophisticated data structures, obtain something like \(O(\sqrt{\varepsilon} \log n)\) or \(O(\varepsilon \log \log n)\). Invariably, the more sophisticated algorithms have a higher constant of proportionality, so that improvements in running time are only possible when \(n\) and \(\varepsilon\) become very large. However, the sophisticated algorithms also indicate that there are theoretical limits of efficiency for the problem at hand.

\section*{Exercises}
4.4.1 Prove that the result of \(\operatorname{LBMERGE}(\mathrm{A}, \mathrm{B})\) is always an \(\operatorname{LB}\)-tree, where \(A\) and \(B\) are non-nil LB-trees.
4.4.2 Let \(A\) be an LB-tree with \(n\) nodes and let \(B\) be an arbitrary node in the tree. Show how to update \(A\) if:
(a) Value \(\langle B\rangle\) is increased.
(b) Value \(\langle B\rangle\) is decreased.
(c) Node \(B\) is removed.
4.4.3 Delayed Merge. When ( \(T u, P Q u\) ) is selected from the TreeQ, and merged with ( \(T u, P Q u\) ), the result is moved to the end of the queue. It may never come to the head of the TreeQ again. In that case, it would not really be necessary to perform the LBMERGE \((P Q u, P Q U)\). Cheriton and Tarjan delay the merging of the two by creating a new dummy node \(D\) and making \(P Q u\) and \(P Q u\) into its right and left subtrees. \(D\) can be marked as a dummy by setting \({ }^{I P a t h}\langle D\rangle\) to -1 . Several dummy nodes may accumulate at the top of the trees \(P Q u\). Should a tree with a dummy node come to the head of the queue, its dummy nodes must be removed before the minimum edge \(x y \in P Q_{u}\) can be selected. Write a recursive tree traversal which removes the dummy nodes from an LB-tree, and places its non-dummy subtrees on the MergeQ. We can then use BUILDLBTREE() to combine all the subtrees on the
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MergeQ into one.
4.4.4 Program the Cheriton-Tarjan algorithm, using leftist binary trees with delayed merge, to store the priority queues.

\subsection*{4.5 Notes}

An excellent description of the cycle space and bond space can be found in BONDY and MURTY [19].
Kruskal's and Prim's algorithms are standard algorithms for minimum spanning trees. They are described in most books on algorithms and data structures. The Cheriton-Tarjan algorithm is from CHERITON and TARJAN
[22]. Leftist binary trees are from KNUTH [75], and are also described in WEISS [122].

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5
The Structure of Trees

\subsection*{5.1 Introduction}

The structure of trees is naturally recursive. When trees are used as data structures, they are typically processed by recursive procedures. Similarly, exhaustive search programs working by recursion also construct trees as they follow their search paths. These trees are always rooted trees; that is, they begin at a distinguished node, called the root vertex, and are usually built outwards from the root. Figure 5.1 shows several rooted trees, where the root vertex is shaded black.


\section*{FI GURE 5.1}

\section*{Several rooted trees}

If \(T\) is a tree, then any vertex \(u\) can be chosen as the root, thereby making \(T\) into a rooted tree. A rooted tree can always be decomposed into branches. The tree \(T\) shown in Figure 5.2 has three branches B1, B2, and B3. DEFINITION 5.1: Let \(T\) have root vertex \(u\). The branches of \(T\) are the maximal subtrees in which \(u\) has degree one.
Thus, the root is in every branch, but the branches have no other vertices in common. The number of branches equals the degree of the root vertex. If we know the branches of some tree \(T\), then we can easily recombine them to get \(T\).
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\section*{FI GURE 5.2}

\section*{Decomposition into branches}

Therefore, two rooted trees have the same structure; that is, they are isomorphic, if and only if they have the same number of branches, and their branches have the same structure.
Any vertex of a tree which has degree one is called a leaf. If the root is a leaf, then \(T\) is itself a branch. In this case, let \(u\) be the unique vertex adjacent to \(u\), the root of \(T\). Then \(T^{\prime}=T-u\) is a rooted tree, with root \(u\). This is illustrated in Figure 5.3. \(T^{\prime}\) can then be further broken down into branches, which can in turn be reduced to rooted trees, etc. This gives a recursive decomposition of rooted trees into branches, and branches into rooted trees.


FI GURE 5.3
Reducing a branch to a rooted tree
This technique can be developed into a method for determining when two rooted trees have the same structure.

\subsection*{5.2 Non-rooted trees}

All non-rooted trees on five and fewer vertices are displayed in Figure 5.4. Table 5.1 gives the number of trees up to 10 vertices.
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\begin{tabular}{|c|c|c|c|}
\hline\(n=1\) & \(n=2\) & \(n=3\) \\
\(n=5\) & \(n=4\) \\
\hline
\end{tabular}

\section*{FI GURE 5.4}

\section*{The trees on five and fewer vertices}

If a leaf be removed from a tree on \(n\) vertices, a tree on \(n-1\) vertices is obtained. Thus, one way to list all the trees on \(n\) vertices is to begin with a list of those on \(n-1\) vertices, and add a leaf in all possible ways, discarding duplicates. How can we recognize when two trees have the same structure? We shall see that nonrooted trees can always be considered as rooted trees, by choosing a special vertex as root, in the center of \(T\), denoted CTR \((T)\). The center is defined recursively.
DEFINITION 5.2: Let \(T\) be a tree on \(n\) vertices.
1. If \(n=1\), say \(V(T)=\{u\}\), then \(\operatorname{CTR}(T)=u\).
2. If \(n=2\), say \(V(T)=\{u, u\}\), then \(C T R(T)=u u\).
3. If \(n>2\), then \(T\) has at least two leaves. Delete all the leaves of \(T\) to get a tree \(T^{\prime}\). \(\operatorname{Then} \operatorname{CTR}(T)=\operatorname{CTR}\left(T^{\prime}\right)\). Thus the center of a tree is either a vertex or an edge, since eventually case (1) or (2) of the definition is used in determining the center of \(T\). Trees whose centre consists of a single vertex are called central trees. Trees with two vertices in the center (i.e., \(\operatorname{CTR}(T)\) is an edge) are called bicentral trees. Figure 5.5 shows two trees, one central and one bicentral.
A central tree can always be considered a rooted tree, by taking the centre as the root. A bicentral tree can also be considered a rooted tree, but we must have a means of deciding which of two vertices to take as the root. Thus we can say that every tree is a rooted tree.

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TABLE 5.1
The number of trees up to 10 vertices
\begin{tabular}{rr}
4 & 2 \\
5 & 3 \\
6 & 6 \\
7 & 6 \\
8 & 11 \\
9 & 23 \\
10 & 47 \\
\hline
\end{tabular}


FI GURE 5.5
A central tree and a bicentral tree

\section*{Exercises}
5.2.1 Find the centre of the trees shown in Figures 5.1 and 5.3.
5.2.2 Prove that any longest path in a tree \(T\) contains the center.
5.2.3 Prove that \(T\) is central if \(\operatorname{DI} \mathrm{AM}(T)\) is even, and bicentral if \(\mathrm{DI} \mathrm{AM}(T)\) is odd.
5.2.4 A binary tree is a rooted tree such that the root vertex has degree two, and all other vertices which are not leaves have degree three. Show that if \(T\) is a binary tree on \(n\) vertices, that \(n\) is odd, and that \(T\) has ( \(n+1\) )/2 leaves.

\subsection*{5.3 Read's tree encoding algorithm}

There are a number of interesting algorithms for encoding trees. Here we present one of Read's algorithms. It is basically an algorithm to find CTR(T), keeping certain information for each vertex as it progresses. When the centre is reached, a root node is uniquely chosen. Read's algorithm encodes a tree as an integer. Its page_92

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description reveals a number of interesting properties satisfied by trees.
Let \(T\) be a tree whose centre is to be found. Instead of actually deleting the leaves of \(T\), let us simply draw a circle around each one. Draw the circle in such a way that it encloses all the circles of any adjacent nodes which have been previously circled. The last vertices to be circled form the centre.


FI GURE 5.6
A circled tree
This system of nested circles can be redrawn in various ways.


\section*{FI GURE 5.7}

\section*{Nested circles}

Each circle corresponds to a vertex of \(T\). The largest circle which encloses the entire system corresponds to the centre of \(T\). Two circles correspond to adjacent vertices if and only if one circle is nested inside the other. The circles not containing a nested circle are the leaves of \(T\). If we cut off the top and bottom of each circle in Figure 5.7, we are left with a set of matched parentheses: (()(()())(()()())). By writing 0 for each left parenthesis and 1 for each right parenthesis, this can be considered a binary number, 001001011001010111 , which represents an integer.
The internal circles in Figure 5.6 have been sorted and arranged in order of increasing complexity. For example, the first inner circle can be denoted 01 . This is less than the second circle, which can be denoted 001011, which in turn is less that the third circle 00101011, considered as binary numbers. Thus, there is a
natural ordering associated with these systems of nested circles.
The binary number associated with each vertex \(u\) is called its tag, denoted \(t(U)\). Initially each leaf has a tag of 01 . The algorithm to find \(\operatorname{CTR}(T)\) constructs the vertex tags as it proceeds.
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\section*{Algorithm 5.3.1: TREEENCODE(T)}
comment: constructs an integer tree code to represent the tree \(T\) repeat construct \(L(T)\), the set of leaves of \(T\), stored on an array
for each leaf \(v \in L(T)\)
do \(\{\) sort the tagged vertices adjacent to \(v\) by tag,
say \(t\left(u_{1}\right) \leq t\left(u_{2}\right) \leq \cdots \leq t\left(u_{k}\right)\)
\(t(v) \leftarrow 0 t\left(u_{1}\right) t\left(u_{2}\right) \cdots t\left(u_{k}\right) 1 \quad\) "concatenate them"
\(T \leftarrow T-L(T)\) "just mark the vertices deleted"
until all of \(T\) has been tagged
if \(L(T)\) contains one vertex \(u\)
then return ( \(t(u)\) )
else \(\left\{\begin{array}{l}\text { comment: } L(T) \text { contains two vertices } u \text { and } v, \text { say } t(u) \leq t(v) \\ \text { return }\left(0 t(u) t^{\prime}(v)\right)\end{array}\right.\)
On the last iteration, when the centre was found, either one or two vertices will have been tagged. They form the centre of \(T\). If \(T\) is a central tree, with \(\operatorname{CTR}(T)=u\), we choose \(u\) as the root of \(T\). Then \(t(u)\), the tag of the centre, represents the entire system of nested circles. It is chosen as the encoding of \(T\).
If \(T\) is a bicentral tree, with centre \(u u\), we must decide which vertex to choose as the root of \(T\). We arbitrarily choose the one with the larger tag. Suppose that \(t(u) \leq t(u)\), so that \(u\) is chosen as the root. The code for the entire tree is formed by altering the enclosing circle of \(u\) so as to enclose the entire tree. This is illustrated in Fiqure 5.8.


\section*{FI GURE 5.8}

Choosing the root vertex
Thus, the tree code for \(T\) in the bicentral case is the concatenation \(0 t(u) t^{\prime}(u)\), where \(t^{\prime}(u)\) is formed from \(t(u)\) by dropping one initial 0-bit.
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If \(t(u)=t(u)\) for some bicentral tree \(T\), then we can obviously select either \(u\) or \(u\) as the root vertex. The easiest way to store the tags \(t(u)\) is as an integer array \(t[U]\). We also need to store the length \(\ell[u]\), of each tag, that is, its length in bits. Initially each leaf \(u\) has \(t[u]=1\) and \(\ell[u]=2\). To concatenate the tags \(0 t(u 1) t(u 2) \ldots t(u k) 1\) we use a loop.
\[
\begin{gathered}
t[U] \leftarrow 0 \\
\text { for } i \leftarrow 1 \text { to } k \\
\text { do }\left\{\begin{array}{l}
\text { shift } t[v] \text { left } \ell[v] \text { bits } \\
t[v] \leftarrow t[v]+t\left[u_{i}\right] \\
\ell[v] \leftarrow \ell[v]+\ell\left[u_{i}\right]
\end{array}\right. \\
t[U] \leftarrow 2 t[U]+1 \\
\ell[U] \leftarrow \ell[U]+2
\end{gathered}
\]

If a primitive left shift operation is not available, one can always store a table of powers of 2 , and use
multiplication by \(2^{\ell\left[u_{i}\right]}\) in order to shift \(t[u]\) left by \(\ell[u i]\) bits.

\subsection*{5.3.1 The decoding algorithm}

If the circles of Figure 5.6 are unnested, they can be redrawn so as to emphasize their relation to the structure of \(T\).


\section*{FI GURE 5.9 \\ Unnested circles}

The decoding algorithm scans across the system of nested circles. Each time a new circle is entered, a vertex is assigned to it. The first circle entered is that corresponding to the root vertex. The decoding algorithm uses a global vertex counter \(k\), which is initially zero, and constructs a global tree \(T\). It can be programmed to scan the tree code from right to left as follows:
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Algorithm 5.3.2: TREEDECODE(Tcode, ui)
comment: a new circle has just been entered, from circle ui.
\(k \leftarrow k+1\) "create a new vertex" join \(v_{k} \longrightarrow v_{i}\)
Tcode \(\leftarrow\) Tcode/2 "shift right 1 bit"
while Tcode is odd
do \(\left\{\begin{array}{l}\text { comment: the rightmost bit }=1, \text { a new circle is entered } \\ \text { TreeDecode }\left(\text { Tcode, }, v_{k}\right) \\ \text { Tcode } \leftarrow \text { Tcode } / 2\end{array}\right.\)
The easiest way to use Algorithm 5.3 .2 is to create a dummy vertex \(\omega 0\) which will only be joined to the root vertex, \(u 1\) and then delete \(u 0\) once the tree has been constructed.
\(k \leftarrow 0\)
TREEDECODE(Tcode,u0)
delete 40

\section*{Exercises}
5.3.1 Encode the trees of Figure 5.5 into nested circles by hand. Write down their tree codes.
5.3.2 Work through Algorithm 5.3.2 by hand, for the tree codes 001011 and 0001011011.
5.3.3 Write Algorithm 5.3.2 so as to scan the tree code from left to right, using multiplication by 2 to shift Tcode to the right, and using the sign bit of the code to test each bit. Assume a word length of 32 bits.
5.3.4 Program the encoding and decoding algorithms.
5.3.5 If \(T\) has \(n\) vertices, what is the total length of its tree code, in bits? How many 1's and 0's does the code contain? How many leading 0 's does the code begin with? What is the maximum value that \(n\) can be if \(T\) is to be encoded in 32 bits?
5.3.6 Let \(T\) be a tree. Prove that the tree obtained by decoding TREEENCODE( \(T\) ), using the decoding algorithm, is isomorphic to \(T\).
5.3.7 Let \(T 1\) and \(T 2\) be two trees. Prove that \(T_{1} \cong T_{2}\) if and only if \(\operatorname{TREEENCODE}(T 1)=\operatorname{TREEENCODE}(T 2)\).
5.3.8 Consider the expression \(x 1 \times 2 \ldots x n+1\), where \(x 1, x 2, \ldots, x n+1\) are variables. If parentheses are inserted so as to take exactly two terms at a time, we obtain a valid bracketing of the expression, with \(n\) pairs of matched parentheses (e.g., ( \(x 1(x 2 x 3)) x 4)\), where \(n=3\) ). Each pair of matched parentheses contains exactly two terms. Describe the type of rooted tree on \(n\) vertices that corresponds to such
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a valid bracketing. They are called binary plane trees. Each leaf of the tree corresponds to a variable xi and each internal node corresponds to a pair of matched parentheses, giving \(2 n+1\) vertices in total.
5.3.9 Let \(p\) denote the number of binary plane trees with \(n\) leaves (e.g., \(p 0=0, p 1=1, p 2=1, p 3=2\), etc.). We take \(p 1=1\) corresponding to the tree consisting of a single node. Let \(p(x)=p 0+p 1 x+p 2 x 2+\ldots\)..be the generating function for the numbers \(p n\), where \(x\) is a variable. If \(T 1\) and \(T 2\) are two binary plane trees with n 1 and n 2 leaves, respectively, then they can be combined by adding a new root vertex, adjacent to the roots of \(T 1\) and \(T 2\). This gives a binary plane tree with \(n 1+n 2\) leaves. There are \(p_{n_{1}} p_{n_{2}}\) ways of constructing a tree in this way. This term arises from \(p 2(x)\) as part of the coefficient of \(x^{n_{1}+n_{2}}\). This holds for all values of \(n 1\) and \(n 2\). Therefore, we can write \(p(x)=x+p 2(x)\). Solve this identity for \(p(x)\) in terms of \(x\), and then use the binomial theorem to write it as a power series in \(x\). Finally, obtain a binomial expression for \(p n\) in terms of \(n\). The numbers \(p n\) are called the Catalan numbers. (The answer should be \(p_{n}=\frac{1}{2 n-1}\binom{2 n-1}{n}\).

\subsection*{5.4 Generating rooted trees}

One way to generate a list of all the trees on \(n\) vertices would be to add a new leaf to the trees on \(n-1\) vertices in all possible ways, and to discard duplicates, using the tree codes to identify isomorphic trees. However, they can also be generated directly, one after the other, with no duplicates.
Let \(T\) be a central tree, rooted at its centre \(u\). Decompose \(T\) into its branches \(B 1, B 2, \ldots, B k\). Each branch \(B i\) is also rooted at \(u\). Write \(T=(B 1, B 2, \ldots, B k)\) to indicate the decomposition into branches. Since \(u\) is the centre of \(T\), it is the middle vertex of every longest path in \(T\). Therefore, the two "tallest" branches of \(T\) will have equal height, where we define the height of a branch \(B\) rooted at \(u\) as \(h(B)=\operatorname{Max}\{\operatorname{Dist}(v, w) \mid w \in B\}\). If the branches of the central tree \(T\) have been ordered by height, so that \(h(B 1) \geq h(B 2) \geq \ldots \geq h(B k)\), then we know that \(h(B 1)=h(B 2)\). Any rooted tree for which the two highest branches have equal height is necessarily rooted at its centre. Therefore, when generating central trees, the branches must be ordered by height.
Generating the rooted trees on \(n\) vertices in a sequence implies a linear ordering on the set of of all rooted trees on \(n\) vertices. In order to construct a data structure representing a rooted tree \(T\) as a list of branches, we also require a linear order on the set of all branches. Then we can order the branches of \(T\) so that \(B 1 \geq B 2 \geq \ldots \geq B k\). This will uniquely identify \(T\), as two trees with the same set branches will have the same ordered set of branches. The smallest possible branch will evidently be of height one, and have two verticesit is \(K 1\) rooted at a vertex. The tree shown in Figure 5.10 has five branches of height one; call them elementary branches. The next smallest branch is of height two, and has three vertices-it is
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the path \(P 2\) rooted at a leaf.
Associated with each branch \(B\) is a rooted tree, as shown in Figure 5.3, constructed by advancing the root to its unique adjacent vertex, and deleting the original root. We will use the ordering of rooted trees to define recursively an ordering of all branches, and the ordering of all branches to define recursively an ordering of all rooted trees. We know that all branches on at most three vertices have already been linearly ordered.


\section*{FI GURE 5.10}

\section*{A tree with all branches of height one}

DEFINITION 5.3: Suppose that all branches on at most \(m \geq 2\) vertices have been linearly ordered. Let \(T=(B 1\), \(B 2, \ldots, B k)\) and \(T^{\prime}=\left(B_{1}^{\prime}, B_{2}^{\prime}, \ldots, B_{\ell}^{\prime}\right)\), where \(k+\ell \geq 3\), be any two distinct rooted trees with given branches, such that each branch has at most \(m\) vertices, ordered so that \(B 1 \geq B 2 \geq \ldots \geq B k\) and \(B_{1}^{\prime} \geq B_{2}^{\prime} \geq \cdots \geq B_{\ell}^{\prime}\). Suppose that \(|T| \leq\left|T^{\prime}\right|\). Then \(T\) and \(T^{\prime}\) are compared as follows:
1. If \(|T|<\left|T^{\prime}\right|\), then \(T<T^{\prime}\).
2. Otherwise, compare ( \(B 1, B 2, \ldots, B \mathrm{~K}\) ) and \(\left(B_{1}^{\prime}, B_{2}^{\prime}, \ldots, B_{\ell}^{\prime}\right)\) lexicographically. That is, find \(i\), the first subscript such that \(B_{i} \neq B_{i}^{\prime}\); then \(T<T^{\prime}\) if \(B_{i}<B_{i}^{\prime}\).
The first condition is to ensure that all rooted trees on \(n\) vertices precede all trees on \(n+1\) vertices in the linear order. The second condition defines an ordering of trees based on the lexicographic ordering of branches. Notice that if \(k \neq \ell\) there must be an \(i\) such that \(B_{i} \neq B_{i}^{\prime}\); for if every \(B_{i}=B_{i}^{\prime}\), but \(k \neq \ell\), then \(T\) and \(T^{\prime}\) would have different numbers of vertices, so that condition (1) would apply. This defines a linear ordering on all rooted trees whose branches all have at most \(m\) vertices. This includes all trees on \(m+1\) vertices with at least two branches. In fact, it includes all rooted trees on \(m+1\) vertices, except for the path Pm, rooted at a leaf. As this is a branch on \(m+1\) vertices, it is handled by Definition 5.4.
We now have an ordering of rooted trees with at least two branches, based on the ordering of branches. We
can use it in turn to extend the ordering of branches. Let \(B\) and \(B^{\prime}\) be two distinct branches. In order to compare \(B\) and \(B^{\prime}\), we advance their roots, as in Figure 5.3 , to the unique adjacent vertex in each, and delete the

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original root. Let the rooted trees obtained in this way be \(T\) and \(T^{\prime}\). respectively. Then \(B<B^{\prime}\) if \(T<T^{\prime}\). In summary, branches are compared as follows:
DEFINITION 5.4: Suppose that all rooted trees on \(\leq m-1\) vertices have been linearly ordered, and that all rooted trees on \(m\) vertices with at least two branches have also been linearly ordered. Let \(B\) and \(B^{\prime}\) be
branches on \(m\) vertices, with corresponding rooted trees \(T\) and \(T^{\prime}\). Suppose that \(|B| \leq\left|B^{\prime}\right|\) and that if \(|B|=\mid B\) \(\downarrow\), then \(B\) is the branch of smaller height. Then \(B\) and \(B^{\prime}\) are compared as follows:
1. If \(|B|<\left|B^{\prime}\right|\), then \(B<B^{\prime}\).
2. Otherwise, if \(h(B)<h\left(B^{\prime}\right)\) then \(B<B^{\prime}\).
3. Otherwise, \(B<B^{\prime}\) if \(T<T^{\prime}\).

We have a recursive ordering which compares trees by the ordering of their branches, and branches by the ordering of their trees. We must prove that the definition is valid.
THEOREM 5.1 Definitions 5.3 and 5.4 determine a linear order on the set of all rooted trees.
PROOF Notice that rooted trees have a sub-ordering based on the number of vertices-all rooted trees on \(n\) vertices precede all rooted trees on \(n+1\) vertices. Branches have an additional sub-ordering based on heightall branches of height \(h\) on \(n\) vertices precede all branches of height \(h+1\) on \(n\) vertices. A branch is a special case of a rooted tree, in which the root vertex has degree one. If a branch \(B\) and tree \(T\) on \(n\) vertices are compared, where \(T\) has at least two branches, then that first branch of \(T\) has fewer than \(n\) vertices, so that \(T<B\), by Definition 5.3. Therefore all trees whose root vertex has degree two or more precede all branches on \(n\) vertices.
The definitions are clearly valid for all rooted trees on \(\leq 3\) vertices. Suppose that the set of all rooted trees on \(\leq n\) vertices is linearly ordered by these definitions, and consider two distinct rooted trees
\(T=(B 1, B 2, \ldots, B k)\)
and
\[
T^{\prime}=\left(B_{1}^{\prime}, B_{2}^{\prime}, \ldots, B_{\ell}^{\prime}\right)
\]
on \(n+1\) vertices. If \(k=\ell=1\), then \(T\) and \(T^{\prime}\) are both branches. The trees formed by advancing their root vertices have only \(n\) vertices, and so can be compared by Definition 5.3. Otherwise at least one of \(T\) and \(T^{\prime}\) has two or more branches.
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Therefore at least one of each pair \(B i\) and \(B_{i}^{\prime}\) of branches has \(\leq n\) vertices. Therefore the branches \(B i\) and \(B_{i}^{\prime}\) can be compared by Definition 5.4. The conclusion follows by induction.
In this ordering of branches and trees, the first rooted tree on \(n\) vertices is a star consisting of the tree \(K 1, n-1\) rooted at its centre. The last rooted tree on \(n\) vertices is a path Pn, rooted at a leaf. The first branch on \(n\) vertices is \(K 1, n-1\), rooted at a leaf, and the last branch on \(n\) vertices is also the path \(P n\), rooted at a leaf. The first few rooted trees are shown in Figure 5.11.


\section*{FI GURE 5.11}

The beginning of the linear order of rooted trees
Let \(T=(B 1, B 2, \ldots, B k)\) be the list of branches of a rooted tree \(T\), with root vertex \(u\). The recursive data structure we use to represent \(T\) is a linked list of branches. Each branch Bi also has root \(U\). It is in turn represented in terms of the rooted tree \(T^{\prime}\), whose root vertex is the unique vertex adjacent to \(u\). Thus, a record representing a tree \(T\) has four fields:
- NodeNum \(\langle T\rangle\), the node number of the root, which is used for printing.
- nNodes \(\langle T\rangle\), the number of vertices in the tree.
- FirstBranch \(\langle T\rangle\), a pointer to the first branch.
- LastBranch \(\langle T\rangle\), a pointer to the last branch.

Each branch \(B\) of \(T\) has a corresponding rooted tree \(T^{\prime} . B\) is represented as a record having four fields:
- Height \(\langle B\rangle\), the height of the branch.
- NextRoot \(\langle B\rangle\), a pointer to the rooted tree \(T^{\prime}\).
- NextBranch \(\langle B\rangle\), a pointer to the next branch of \(T\).
- PrevBranch \(\langle B\rangle\), a pointer to the previous branch of \(T\).
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It is not necessary to store the number of vertices of a branch \(B\), as it is given by \(\langle\operatorname{NextRoot}\langle B\rangle\rangle+1\). The functions which compare two trees and branches are given as follows. They return an integer, whose value is one of three constants, LessThan, EqualTo, or GreaterThan.

Algorithm 5.4.1: COMAPARETREES \(T 1, T 2\) )
comment: \(T 1\) and \(T 2\) both have at least one branch
if \(|T 1|<|T 2|\) then return (LessThan)
if \(|T 1|>|T 2|\) then return (GreaterThan)
comment: otherwise \(|T 1|=|T 2|\)
\(B_{1} \leftarrow\) FirstBranch \(\left\langle T_{1}\right\rangle\)
\(B_{2} \leftarrow\) FirstBranch \(\left\langle T_{2}\right\rangle\)
Result \(\leftarrow\) COMPAREBRANCHES (B1, B2)
while Result=EqualTo
do \(\left\{\begin{array}{l}\text { if } B_{1}=\text { LastBranch }\left\langle T_{1}\right\rangle \text { then return (EqualTo) } \\ B_{1} \leftarrow \text { NextBranch }\left\langle B_{1}\right\rangle \\ B_{2} \leftarrow \text { NextBranch }\left\langle B_{2}\right\rangle \\ \text { Result } \leftarrow \text { CompareBRANCHES }\left(B_{1}, B_{2}\right) \\ \text { return (Result) } \\ \text { Algorithm 5.4.2: COMPAREBRANCHES }(B 1, B 2)\end{array}\right.\)
comment: \(B 1\) and \(B 2\) both have a unique vertex adjacent to the root
if \(|B 1|<|B 2|\) then return (LessThan)
if \(|B 1|>|B 2|\) then return (GreaterThan)
comment: otherwise \(|B 1|=|B 2|\)
Height \(\left\langle B_{1}\right\rangle<\) Height \(\left\langle B_{2}\right\rangle\) then return (LessThan)
Height \(\left\langle B_{1}\right\rangle>\operatorname{Height}\left\langle B_{2}\right\rangle\) then return (GreaterThan)
comment: otherwise Height \(\left\langle B_{1}\right\rangle=\operatorname{Height}\left\langle B_{2}\right\rangle\)
if Height \(\left\langle B_{1}\right\rangle=1_{\text {then }}\) return (EqualTo)
\(T_{1} \leftarrow \operatorname{NextRoot}\left\langle B_{1}\right\rangle\)
\(T_{2} \leftarrow \operatorname{NextRoot}\left\langle B_{2}\right\rangle\)
return (COMPARETREES ( \(T 1, T 2\) ))
Using these functions we can generate all rooted trees on \(n\) vertices, one after the other, beginning with the first tree, which consists of a root vertex and \(n-1\) elementary branches of height one, until the last tree is reached, which has only one branch, of height \(n-1\). Written in pseudo-code, the technique is as follows, page_101

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where NEXTTREE \((T)\) is an procedure which replaces \(T\) with the next tree, and returns true unless \(T\) was the last tree (see Algorithm 5.4.3). FIRSTTREE( \(n\) ) is a procedure that constructs the first tree on \(n\) vertices.
\(T \leftarrow\) FIRSTTREE( \(n\) )
repeat
PRINTTREE(T)
until not NEXTTREE \((T)\)

Suppose that \(T\) has branch decomposition ( \(B 1, B 2, . ., B k\) ), where \(B 1 \geq B 2 \geq \ldots \geq B k\). The procedure NEXTTREE (T) works by finding the last branch \(B i\) such that \(B i \neq B i-1\). Then \(B i, B i+1, \ldots, B k\) is a sequence of isomorphic branches. So long as \(B i\) is not simply a path of length \(h(B i)\), there is a larger branch with the same number of vertices. \(B i\) is then replaced with the next larger branch and the subsequent branches \(\mathrm{Bi}+1, \mathrm{Bi}+2\), \(\ldots, B k\) are replaced with a number of elementary branches. This gives the lexicographically next largest tree. This is illustrated in Figure 5.12 . Here, \(B 2\) was replaced with a branch of height three, and \(B 3\) was replaced with three elementary branches.


\section*{FI GURE 5.12}

\section*{Constructing the next tree}

But if \(B i\) is simply a path, then it is the last branch with | Bi/ vertices. In order to get the next branch we must add another vertex. \(B i\) is then replaced with the first branch with one more vertex. This is the unique branch with \(|B i|+1\) vertices and height two. \(T\) is then filled in with as many elementary branches as needed. This is illustrated in Figure 5.13.
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Algorithm 5.4.3: NEXTTREE(T)
\(B_{1} \leftarrow\) LastBranch \(\langle T\rangle\)
\(B_{1}=\) FirstBranch \(\langle T\rangle\)
then \(\left\{\begin{array}{l}\text { comment: only one branch - advance the root } \\ \text { if } n N o d e s\langle T\rangle=\operatorname{Height}\left\langle B_{1}\right\rangle+1 \text { then return (false ) } \\ \left.\text { return (NEXTTREE }\left(\operatorname{NextRoot}\left\langle B_{1}\right\rangle\right)\right)\end{array}\right.\)
comment: otherwise at least two branches
\(B_{2} \leftarrow \operatorname{Prev}\) Branch \(\left\langle B_{1}\right\rangle\)
while COMPAREBRANCHES(B1, B2)=EqualTo
do \(\left\{\begin{array}{l}B_{1} \leftarrow B_{2} \\ \text { if } B_{1}=\text { FirstBranch }\langle T\rangle \text { then go to } 1 \\ B_{2} \leftarrow \operatorname{PrevBranch}\left\langle B_{2}\right\rangle\end{array}\right.\)
1: comment: delete the branches of \(T\) following \(B 1\)
\(N \leftarrow\) DELETEBRANCHES( \(T, B 1\) ) " \(N\) nodes are deleted"
comment: replace \(B 1\) with next branch, if possible if \(n\) Nodes \(\left\langle\right.\) NextRoot \(\left.\left\langle B_{1}\right\rangle\right\rangle>\operatorname{Height}\left\langle B_{1}\right\rangle\)
then \(\left\{\begin{array}{l}\text { NEXTTREE }\left(\operatorname{NextRoot}\left\langle B_{1}\right\rangle\right) \\ \text { fill in } T \text { with } N \text { elementary branches } \\ \text { return (true ) }\end{array}\right.\)
comment: otherwise construct the first branch with one more node
\(M \leftarrow \operatorname{DESTROYTREE}\left(\operatorname{NextRoot}\left\langle B_{1}\right\rangle\right)\) " \(M\) nodes are deleted"
NextRoot \(\left\langle B_{1}\right\rangle \leftarrow \operatorname{FirstTreE}(M+1)\)
fill in \(T\) with \(N-1\) elementary branches
if \(N>0\) then return (true)
comment: otherwise there's no branch following \(B 1\) to take a node from
```

if B
until COMPAREBRANCHES (B1, B2) = EqualTo
2: comment: delete the branches of T following B1
N\leftarrowDELETEBRANCHES(T, B1) " N nodes are deleted"
comment: replace B1 with next branch
if nNodes }\langle\mathrm{ NextRoot }\langle\mp@subsup{B}{1}{}\rangle\rangle>>\operatorname{Height}\langle\mp@subsup{B}{1}{}
then NextRoot }\langle\mp@subsup{B}{1}{}
else NextRoot }\langle\mp@subsup{B}{1}{}\rangle\leftarrow\operatorname{FIRSTTREE}(nNodes \NextRoot \langle\mp@subsup{B}{1}{}\rangle\rangle+1+1) fill in T with elementary branche
return (true)
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```
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\(B_{1}\)


\section*{FI GURE 5.13}

\section*{Constructing the next tree}

The procedure DELETEBRANCHES ( \(T, B 1\) ) destroys all branches of \(T\) following \(B 1\), and returns the number of nodes deleted. Similarly DESTROYTREE \((T)\) is a procedure that destroys all branches of \(T\), and returns the total number of nodes deleted.
THEOREM 5.2 Let \(T\) be a tree on \(n\) vertices. Algorithm NEXTTREE( \(T\) ) constructs the next tree on \(n\) vertices after \(T\) in the linear order of trees, if there is one.
PROOF The proof is by induction on \(n\). It is easy to check that it works for trees on \(n=2\) and \(n=3\) vertices. Suppose that it holds up to \(n-1\) vertices, and let \(T\) have \(n\) vertices. Let \(T=(B 1, B 2, \ldots, B k)\) be the branches of \(T\), where \(B 1 \geq B 2 \geq \ldots \geq B k\). The algorithm first checks whether there is only one branch. If so, and \(T\) is a branch of height \(n-1\), it returns false. Otherwise let \(T^{\prime}\) be the rooted tree corresponding to \(B 1\) by advancing the root. The algorithm calls NEXTTREE \(\left(T^{\prime}\right)\). Since \(T^{\prime}\) has \(n-1\) vertices, this gives the next branch following \(B 1\) in the linear order, as required.
Otherwise, \(T\) has at least two branches. It finds the branch \(B i\) such that \(B i=B i+1=\ldots=B k\), but \(B i-1 \neq B i\), if there is one (possibly \(i=1\) ). The first tree following \(T\) must differ from \(T\) in \(B i\), unless \(i=k\) and \(B i\) is a path. In the first case, the algorithm replaces Bi with the next branch in the linear order, and fills in the remaining branches of \(T\) with elementary branches. This is the smallest tree on \(n\) vertices following \(T\). Otherwise \(i=k\) and \(B i\) is a path, so that there is no tree following \(B i\). Since there are at least two branches, the algorithm finds the branch \(B j\) such that \(B j=B j+1=\ldots=B k-1>B k\) (possibly \(j=1\) ). It then replaces \(B j\) with the first branch following it, and fills in \(T\) with elementary branches. In each case the result is the next tree after \(T\).
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\section*{Exercises}
5.4.1 Work through the \(\operatorname{NEXTTREE}(T)\) algorithm by hand to construct all the rooted trees on 4, 5, 6, and 7 vertices.
5.4.2 Write the recursive procedure PRINTTREE \((T)\) to print out a tree as shown in Figure 5.14, according to the distance of each vertex from the root.


FI GURE 5.14

\section*{Printing a rooted tree}
5.4.3 Write the recursive functions \(\operatorname{DESTROYTREE}(T)\) and \(\operatorname{DELETEBRANCHES}(T, B 1)\), both of which return the number of vertices deleted.
5.4.4 Program the \(\operatorname{NEXTTREE}(T)\) algorithm, and use it to find the number of rooted trees up to 10 vertices.

\subsection*{5.5 Generating non-rooted trees}

The NEXTTREE(T) algorithm generates the rooted trees on \(n\) vertices in sequence, beginning with the first tree of height 1 and ending with the tree of height \(n-1\). In order to generate non-rooted trees, we must root them in the centre. Since every non-rooted tree can be viewed as a rooted tree, all non-rooted trees also occur in the linear order of trees. Central trees can be generated by modifying NEXTTREE(T) so that the two highest branches are always required to have equal height. This can be done with another procedure,
NEXTCENTRALTREE(T), which in turn calls NEXTTREE(T) when forming the next branch. Bicentral trees are slightly more difficult, since the highest branches \(B 1\) and \(B 2\) satisfy \(h(B 1)=h(B 2)+1\). If we generate trees in which the heights of the two highest branches differ by one, then most bicentral trees will be constructed twice, once for each vertex in the centre. For example, Figure 5.15 shows two different branch

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decompositions of the same bicentral tree. It would therefore be generated twice, since it has different branch decompositions with respect to the two possible roots.

\(T_{1}\)

\(T_{2}\)

\section*{FI GURE 5.15}

\section*{Two decompositions of a bicentral tree}

The easiest solution to this is to subdivide the central edge with a new vertex, taking it as the root. Then each bicentral tree on \(n\) vertices corresponds to a unique central tree on \(n+1\) vertices, with exactly two branches. We can construct these by generating rooted trees with only two branches, which have equal height, and then ignoring the extra root vertex.

\section*{Exercises}
5.5.1 Write and program the procedures NEXTCENTRALTREE(T) and NEXTBICENTRALTREE(T), and use them to construct all the non-rooted trees on \(n\) vertices, up to \(n=15\).

\subsection*{5.6 Prüfer sequences}

Read's algorithm encodes a tree according to its isomorphism type, so that isomorphic trees have the same code. This can be used to list all the isomorphism types of trees on \(n\) vertices. A related question is to make a list all the trees on the \(n\) vertices \(V n=\{1,2, \ldots, n\}\). These are sometimes referred to as labeled trees. For example, Figure 5.16 illustrates the three distinct, or labeled trees on three vertices, which are all isomorphic to each other.
Let \(T\) be a tree with \(V(T)=\{1,2, \ldots, n\}\). A Prüfer sequence for \(T\) is a special encoding of \(T\) as an integer sequence. For example, the tree of Figure 5.17 with \(n=9\) has Prüfer sequence \(t=(6,9,1,4,4,1,6)\).
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1


2


3

FI GURE 5.16
Three distinct trees on three vertices

3


\section*{FI GURE 5.17}

\section*{Finding the Prüfer sequence of a tree}

This is constructed as follows. The leaves of \(T\) are \(L(T)=\{2,3,5,7,8\}\). The numerically smallest leaf is 2 . Since \(2 \rightarrow 6\), we take \(t 1=6\) as the first member of \(t\). We now set \(T:=T-2\), and find \(L(T)=\{3,5,7,8\}\). We again choose the smallest leaf, 3 , and since \(3 \rightarrow 9\), we take \(t 2=9\) as the second member of the sequence, and set \(T:=T-3\). Notice that when 3 is deleted, 9 becomes a leaf. Therefore, on the next iteration we will have \(L(T)=\{5,7,8,9\}\). The general technique is the following:
for \(k \leftarrow 1\) to \(n-2\)
\[
\text { do }\left\{\begin{array}{l}
\text { find } L(T) \\
\text { select } v \in L(T), \text { the smallest leaf } \\
t_{k} \leftarrow \text { the unique vertex adjacent to } v \\
T \leftarrow T-v
\end{array}\right.
\]
comment: \(T\) now has 2 vertices left
This always gives a sequence of \(n-2\) integers \(t=(t 1, t 2, \ldots, t n-2)\), where each \(t_{i} \in V_{n}\). Notice that at each step, a leaf of \(T\) is deleted, so that \(T\) is always a tree throughout all the steps. Since \(T\) is a tree, we can always choose a leaf to delete. When \(T\) reduces to a single edge, the algorithm stops. Therefore no leaf of \(T\) is ever chosen as any \(t k\). In fact, if \(D E G(u) \geq 2\), then \(u\) will appear in \(t\) each time a leaf adjacent to \(u\) is deleted. When the degree drops to one, \(u\) itself
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becomes a leaf, and will appear no more in \(t\). Therefore, each vertex \(u\) appears in \(t\) exactly DEG(U)-1 times. THEOREM 5.3 Let \(t=(t 1, t 2, \ldots, t n-2)\) be any sequence where each \(t_{i} \in V_{n}=\{1,2, \ldots, n\}\). Then \(t\) is the Prüfer sequence of a tree \(T\) on \(n\) vertices.
PROOF The sequence \(t\) consists of \(n-2\) integers of \(V n\). Therefore at least two members of \(V n\) are not used in \(t\). Let \(L\) be those numbers not used in \(t\). If \(t\) were formed by encoding a graph using the above technique, then the smallest element \(v \in L\) must have been a leaf adjacent to \(t 1\). So we can join \(u \rightarrow t 1\), and discard \(t 1\) from \(t\). Again we find \(L\), the numbers not used in \(t\), and pick the smallest one, etc. This is summarized in the following pseudo-code:
\(N_{\leftarrow} \leftarrow\{1,2, \ldots, n\}\)
for \(k \leftarrow 1\) to \(n-2\)
\[
\left\{\begin{array}{l}
\text { construct } L \text {, those numbers of } N \text { not used in } t \\
\text { select the smallest } v \in L \\
\text { join } v \longrightarrow t_{k} \\
\text { discard } t_{k} \text { from } t \\
N \leftarrow N-v
\end{array}\right.
\]
comment: \(T\) now has 2 vertices left, \(u\) and \(u\) join \(u \longrightarrow v\)
This creates a graph \(T\) with \(n-1\) edges. It is the only graph which could have produced the Prüfer sequence \(t\), using the above encoding technique. Must \(T\) be a tree? If \(T\) were not a tree, then it could not be connected,
since \(T\) has only \(n-1\) edges. In that case, some component of \(T\) would have a cycle \(C\). Now the encoding technique only deletes leaves. No vertex on a cycle could ever be deleted by this method, for the degree of every \(u \in C\) is always at least two. This means that a graph containing a cycle would not produce a Prüfer sequence of length \(n-2\). Therefore \(T\) can have no cycle, which means that it must be a tree.
Thus we see that every tree with vertex set \(\{1,2, \ldots, n\}\) corresponds to a unique Prüfer sequence, and that every sequence \(t\) can only be obtained from one tree. The corollary is that the number of trees equals the number of sequences. Now it is clear that there are \(n n-2\) such sequences, since each of the \(n-2\) elements \(t k\) can be any of the numbers from 1 to \(n\). This result is called Cayley's theorem.
THEOREM 5.4 (Cayley's theorem) The number of distinct trees on \(n\) vertices is \(n n-2\).

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\subsection*{5.7 Spanning trees}

Consider the problem of making a list of all the spanning trees of a graph \(G\). If \(G \cong K_{n}\), then there are \(n n-2\) spanning trees, and each one corresponds to a Prüfer sequence. If \(G \neq K_{n}\), then we can find all the spanning trees of \(G\) as follows. Choose any edge \(u v\) of \(G\). First find all the spanning trees that use uu and then find all the trees that don't use \(u u\). This gives all spanning trees of \(G\). Write \(T(G)\) for the number of spanning trees of \(G\). The spanning trees \(T\) that don't use edge \(u v\) are also spanning trees of \(G-u u\), and their number is \(T(G-u u)\). If \(T\) is a spanning tree that does use \(u u\), then we can contract the edge \(u u\), identifying \(u\) and \(u\) so that they become a single vertex. Let \(T \cdot u u\) denote the reduced tree. It is a spanning tree of \(G \cdot u u\). Every spanning tree of \(G u v\) is a contraction \(T \cdot u u\) of some spanning tree \(T\) of \(G\); for just expand the contracted edge back into uu to get \(T\). This gives:
LEMMA 5.5 Let \(G\) be any graph. Then \(T(G)=T(G-u u)+\tau(G-u u)\).
This applies equally well to simple graphs and multigraphs. It is illustrated in Figures 5.18 and 5.19.


\section*{FI GURE 5.18}

Deletion and contraction of edge uu
Notice that even when \(G\) is a simple graph, \(G u u\) will often be a multigraph, or have loops. Now loops can be discarded, since they cannot be part of any spanning tree. However multiple edges must be kept, since they correspond to different spanning trees of \(G\).
In the example above, the 5-cycle obviously has five spanning trees. The other graph is then decomposed, giving "trees" which contain some multiple edges (i.e., replacing the multiple edges with single edges gives a tree). The two such "trees" shown clearly have two and four spanning trees each, respectively. Therefore the original graph has \(5+2+4=11\) spanning trees.


\section*{FI GURE 5.19}

\section*{Finding the number of spanning trees}

In general, if \(G\) has an edge of multiplicity \(k\) joining vertices \(u\) and \(u\), then deleting any one of the equivalent \(k\) edges will give the same number of spanning trees. Contracting any one of them forces the rest to collapse
into loops, which are then discarded. This gives the following lemma:
LEMMA 5.6 Let edge uu have multiplicity \(k\) in \(G\). Replace the multiple edges having endpoints \(u\) and \(u\) by a single edge uu to get a graph Guu. Then
\[
T(G)=T(G u v-u u)+k T(G u v \cdot u u)
\]

This gives a recursive technique for finding the number of spanning trees of a connected graph \(G . G\) is stored as a weighted simple graph, for which the weight of an edge represents its multiplicity.

\section*{Algorithm 5.7.1: SPTREES(G)}
find an edge uu on a cycle
if there is no such edge
then \(\left\{\begin{array}{l}\text { comment: } G \text { is a tree } \\ \text { return (product of edge weights) }\end{array}\right.\)
else return (SPTREES( G-uu)+WT(uu)*SPTREES( G uu))
This can be expanded to make a list of all spanning trees of \(G\). However, if only the number of spanning trees is needed, there is a much more efficient method.
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\subsection*{5.8 The matrix-tree theorem}

The number of spanning trees of \(G\) can be computed as the determinant of a matrix. Let \(A(G)\) denote the adjacency matrix of \(G\). The degree matrix of \(G\) is \(D(G)\), all of whose entries are 0 , except for the diagonal, which satisfies [D]uu=DEG(u), for vertex \(u\). The Kirchhoff matrix of \(G\) is \(K(G)=D-A\). This matrix is sometimes also called the Laplacian matrix of \(G\). The number of spanning trees is found from the Kirchhoff matrix. First, notice that the row and column sums of \(K\) are all 0 , since the row and column sums of \(A\) are the degrees of \(G\). Therefore, \(\operatorname{det}(K)=0\). Consider the expansion of \(\operatorname{det}(K)\) into cofactors along row \(u\). Write
\[
\operatorname{det}(K)=\sum_{v=1}^{n}(-1)^{u+v} k_{u v} \operatorname{det}\left(K_{u v}\right)
\]

Here kuu denotes the entry in row \(u\) and column \(u\) of \(K\), and \(K u u\) denotes the submatrix formed by crossing out row \(u\) and column \(u\). The cofactor of kuu is \((-1) u+u \operatorname{det}(K u u)\). There are \(n\) vertices.
THEOREM 5.7 (Matrix-Tree Theorem) Let \(K\) be the Kirchhoff matrix of \(G\). Then \(\tau(G)=(-1) u+u \operatorname{det}(K u u)\), for any row index \(u\) and any column index \(u\).
PROOF Notice that the theorem says that all cofactors of \(K\) have the same value, namely, the number of spanning trees of \(G\). The proof is by induction on the number of vertices and edges of \(G\). Suppose first that \(G\) is a disconnected graph; let one of the components be \(H\). Order the vertices so that vertices of \(H\) come before the rest of \(G\). Then \(K(G)\) is a block matrix, as shown in Figure 5.20.
If row \(u\) and column \(u\), where \(u, v \in V(H)\), are crossed off, then the row and column sums of \(G-H\) will be all 0 , so that the cofactor corresponding to Kuu will be zero. Similarly, if any other row and column are crossed off, the remaining cofactor will be zero. Therefore, if \(G\) is disconnected, the theorem is true, since \(T(G)=0\).
Suppose now that \(G\) is a tree. Choose a leaf \(u\) and let \(v \longrightarrow u\). Without loss of generality, we can order the vertices so that \(u\) is the last vertex. Write \(K \cdot u u=K(G \cdot u u)\). The two Kirchhoff matrices are shown in Figure 5.21, where \(a=k u u\).

If \(n=2\), there is only one tree, with Kirchhoff matrix \(\left[\begin{array}{cc}1 & -1 \\ -1 & 1\end{array}\right]\). All the cofactors have value \(\pm 1\), as desired. If \(n>2\), we assume that the matrix-tree theorem holds for all trees with at most \(n-1\) vertices, and form \(K-u u\). Now \(\operatorname{det}(K \cdot u u)=0\), since it is a Kirchhoff matrix. The submatrix \(K u u\) differs from \(K \cdot u u\) only in the single term a instead of a-1 in entry \(u u\). When we expand \(\operatorname{det}(K \cup u)\) along row
\(K(G)=\)\begin{tabular}{|c|c|}
\hline\(K(H)\) & 0 \\
\hline & \\
\hline 0 & \(K(G-H)\) \\
\hline
\end{tabular}

\section*{FI GURE 5.20}

\section*{Kirchhoff matrix of a disconnected graph}
\(u\), all the terms are identical to expanding \(\operatorname{det}(K \cdot u u)\) along row \(u\), except for this one. Therefore
\(\operatorname{det}(K u u)-\operatorname{det}(K \cdot u u)\) equals the uu-cofactor in \(K \cdot u u\). By the induction hypothesis, this is \(T(G \cdot u u)=1\). Therefore \(\operatorname{det}(K u u)=1\). Striking off row and column \(u\) from \(K\), and expanding along row \(u\), shows that det (Kuu) again equals the uu-cofactor in \(K \cdot u u\), which is 1 . Therefore, the uu-cofactor in \(K\) equals \(T(G)\). Consider next the cofactor \(\operatorname{det}(K x y)\), where neither \(x\) nor \(y\) equals \(u\) or \(u\). Strike off row \(x\) and column \(y\) of \(K\). In order to evaluate \((-1) x+y \operatorname{det}(K x y)\), first add row \(u\) to row \(u\), and then expand the determinant along column \(u\). The value clearly equals the \(x y\)-cofactor of \(K \cdot u u\), which is \(T(G)=1\). If \(x=u\) but \(y \neq u\) or \(u\), a similar argument shows that the cofactor equals 1. If \(x=u\) but \(y \neq u\), then expand along column \(u\) to evaluate \((-1) x+y \operatorname{det}(K x y)\). The result is \((-1) x+y(-1)(-1) u+x-1\) times thedeterminantof \((K \cdot u u) u y\). This reduces to \((-1) u+y\)
\(\operatorname{det}\left(\left(K^{\prime} \cdot u u\right) u y\right)=T(G)\). Thus, in all cases, the cofactors equal \(T(G)=1\). By induction, the matrix-tree theorem is true for all trees.
If \(G\) is a multigraph with \(n=2\) vertices, then it consists of \(m\) parallel edges, for some \(m \geq 1\), so that \(T(G)=m\). It is easy to see that the theorem is true in this case, as the Kirchhoff matrix is \(\left[\begin{array}{cc}m-m \\ -m & m\end{array}\right]\). Suppose now that it holds for all multigraphs with fewer than \(n\) vertices, where \(n>2\), and for all multigraphs on \(n\) vertices with less than \(\varepsilon\) edges, where \(\varepsilon>n-1\), since we know that it holds for trees. Choose an edge \(u \cup\) of \(G\), and write \(\tau(G)=\tau(G-u u)+\tau(G \cdot u u)\). The corresponding Kirchhoff matrices are illustrated in Figure 5.22, where we write \(K-u u\) for \(K(G-u u)\).
The diagram is drawn as though the edge uu had multiplicity 1, but the proof is general, and holds for any multiplicity \(m \geq 1\). Let a denote the entry kuu and \(b\) the entry kuu.
Consider the \(u v\)-cofactor of \(K\). It is nearly identical to the \(u v\)-cofactor of
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\section*{FI GURE 5.21}

\section*{Kirchhoff matrices for a tree}
\(K-u u\), differing only in the uu-entry, which is a in \(K\) but \(a-1\) in \(K-u u\). Expanding \(\operatorname{det}(K u u)\) along row \(u\) shows that \(\operatorname{det}(K u u)-\operatorname{det}((K-u u) u u)\) equals the \(u u\)-cofactor of \(K-u u\), with row and column \(u\) removed. Comparison with Figure 5.23 shows that this is identical to the uu-cofactor of \(K \cdot u u\). Therefore
\(\operatorname{det}(K u u)-\operatorname{det}((K-u u) u u)=\operatorname{det}((K \cdot u u) u u)\). By the induction hypothesis, this gives \(\operatorname{det}(K u u)=\tau(G \cdot u u)+\tau(G-u u)=T(G)\), as desired.
Consider now Kuu, formed by striking off row \(u\) and column \(u\) of \(K\). This matrix is almost identical to that formed by striking off row \(u\) and column \(u\) from \(K-u u\). The only difference is in the \(u u\)-entry. Expanding along row \(u\) shows that the difference of the determinants, \(\operatorname{det}(K u u)-\operatorname{det}((K-u u) u u)\) is \((-1) u+u-1(-1)\) \(\operatorname{det}((K \cdot u u) u u)\). Therefore \((-1) u+u \operatorname{det}(K u u)=(-1) u+u \operatorname{det}((K-u u) u u)+\operatorname{det}((K \cdot u u) u u)=T(K-u u\) Thus, the \(u u-\) cofactor and the \(u v\)-cofactor both equal \(T(G)\).
Finally, we show that the remaining entries in row \(u\) also have cofactors equal to \(T(G)\). Consider any entry Kuw, where \(w \neq u, u\). Strike off row \(u\) and column \(w\) of \(K\) and of \(K-u u\). In order to evaluate det (Kuw) and \(\operatorname{det}((K-u u) u w)\), first add the remaining part of column \(u\) to column \(u\) in both matrices. This is illustrated in Figure 5.24.
Kuw and \((K-u u)\) uw are now identical, except for the uu-entry. Expand det(Kuw) along row \(u\). All terms are equal to the corresponding terms in the expansion of \(\operatorname{det}((K-u u) u w)\) along row \(u\), except for the last term. The difference is \((-1) u+u-1(-1) \operatorname{det}((K \cdot u u) u w)\). Therefore \((-1) u+w \operatorname{det}(K u w)=(-1) u+w\) \(\operatorname{det}((K-u u) u w)+(-1) u+w \operatorname{det}((K \cdot u u) u w)\). As before, we get \((-1) u+w \operatorname{det}(K u w)=T(G)\). Thus, all the cofactors of \(K\) from row \(u\) have equal
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\section*{FI GURE 5.22}

\section*{Kirchhoff matrices \(K\) and \(K-u u\)}
value, namely, \(\tau(G)\). Since \(u\) could be any vertex, all the cofactors of \(K\) have this value. This completes the proof of the matrix-tree theorem.
A nice illustration of the use of the matrix-tree theorem is to compute \(T(K n)\). The Kirchhoff matrix is
\[
K\left(K_{n}\right)=\left[\begin{array}{ccccc}
n-1 & -1 & -1 & \cdots & -1 \\
-1 & n-1 & -1 & \cdots & -1 \\
-1 & -1 & n-1 & & \vdots \\
\vdots & \vdots & & \ddots & -1 \\
-1 & -1 & \cdots & -1 & n-1
\end{array}\right]
\]

Strike off the last row and column. In order to evaluate the determinant, add all the rows to the first row, to get
\[
\left[\begin{array}{cccc}
1 & 1 & \cdots & \\
-1 & n-1 & & \\
\vdots & & \ddots & \\
& & & n-1
\end{array}\right]_{(n-1) \times(n-1)}
\]

Now add the first row to each row in turn, in order to get n's on the diagonal and 0's off the diagonal. Thus, the determinant is \(n n-2\), as expected.
The Kirchhoff matrix was first used to solve electrical circuits. Consider a simple electrical network consisting of resistors and a battery producing voltage \(V\). Let the nodes in the network be \(u 1, u 2, \ldots\), un, and suppose that the resistance

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(v)

\(K \cdot u v\)
FI GURE 5.23

\section*{Kirchhoff matrix \(K \cdot u \cup\)}
connecting \(u i\) to \(u j\) is rij. The battery causes current to flow in the network, and so sets up a voltage Vi at each node ui. The current from \(u i\) to \(u j\) is ( \(V i-V j\) )/rij. This is illustrated below.
The law of conservation of charge say sthat the total current flowing out of node ui must equal the total current flowing in, that is, not counting the current flowing through the battery,
\[
\sum_{u_{j} \rightarrow u_{i}} \frac{\left(V_{i}-V_{j}\right)}{r_{i j}}=0,
\]
for all nodes ui. The battery maintains a constant voltage difference of \(V\) across nodes \(u \mathrm{l}\) and \(u\) n, say. Let \(I\) denote the current through the battery. Then the \(u 1\)-equation must be modified by setting the right-hand side to \(I\) instead of 0 ; and the un-equation requires the right-hand side to be- \(I\). This gives a system of linear equations in the variables \(V i\) and \(I\). If we consider the network as a multigraph in which \(u i\) is joined to \(u j\) by \(1 /\) rij parallel edges, then the diagonal entries of the matrix corresponding to the equations are the degrees of the nodes. The offdiagonal entries form the negated adjacency matrix of the network. Thus, this is the Kirchhoff matrix of the network. Since the Kirchhoff matrix has determinant zero, there is no unique solution to the system. However, it is voltage differences that are important, and we know that the battery maintains a constant voltage difference of \(V\). Therefore, we can arbitrarily set \(V n=0\) and \(V I=V\), so that we can cross off the nth column from the matrix. The rows are linearly dependent, so that we can also discard any row. The system then has a unique solution, since striking off a row and column from the Kirchhoff matrix leaves the spanning tree matrix. Notice that once the current in each edge is known, each spanning tree
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FIGURE 5.24

\section*{Evaluating det(Kuw)}
of the network will determine the voltage distribution uniquely, since a spanning tree has a unique path connecting any two vertices.

\section*{Exercises}
5.8.1 Find \(T(K 3,3)\) using the recursive method.
5.8.2 Find \(\tau(K n-u u)\), where \(u v\) is any edge of \(K n\), using the matrix-tree theorem.
5.8.3 Find \(T(C n)\), where \(C n\) is the cycle of length \(n\), using the matrix-tree theorem.
5.8.4 What is the complexity of finding the determinant of an \(n \times n\) matrix, using Gaussian elimination?

Accurately estimate an upper bound on the number of steps needed.
5.8.5 Let \(G\) be a graph with \(n\) vertices. Replace each edge of \(G\) with \(m\) multiple edges to get a graph \(G m\). Prove that \(T(G m)=m n-1 T(G)\).
5.8.6 Program the recursive algorithm to find the number of spanning trees. Use a breadth-first search to find an edge on a cycle.
5.8.7 Solve the electrical circuit of Figure 5.25, taking all resistances equal to one. Solve for the voltage Vi at each node, the current in each edge, and the total current \(I\), in terms of the battery voltage \(V\).

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\section*{FI GURE 5.25}

\section*{A simple network of resistors}

\subsection*{5.9 Notes}

Read's tree encoding algorithm is from READ [98]. Prüfer sequences date back to 1918-PRÜFER [95]. They are described in several books, including BONDY and MURTY [19]. The matrix-tree theorem is one of the most fundamental theorems in graph theory.
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6
Connectivity

\subsection*{6.1 I ntroduction}

Trees are the smallest connected graphs. For deleting any edge will disconnect a tree. The following figure shows three graphs in order of increasing connectivity.

\[
\begin{gathered}
\kappa=1 \\
\kappa^{\prime}=1
\end{gathered}
\]

\[
\kappa=2
\]
\[
\kappa^{\prime}=3
\]

\[
\begin{gathered}
\kappa=4 \\
\kappa^{\prime}=4
\end{gathered}
\]

FI GURE 6.1
Three graphs with increasing connectivity

The second graph can be disconnected by deleting the two shaded vertices, but three edges must be deleted in order to disconnect it. The third graph is complete and cannot be disconnected by deleting any number of vertices. However, the deletion of four edges will do so. Thus, connectivity is measured by what must be deleted from a graph \(G\) in order to disconnect it. Because one can delete vertices or edges, there will be two measures of connectivity.
The vertex-connectivity of \(G\) is \(\kappa(G)\), the minimum number of vertices whose deletion disconnects \(G\). If \(G\) cannot be disconnected by deleting vertices, then \(\kappa(G)=|G|-1\). A disconnected graph requires the deletion of 0 vertices, so it has \(\kappa=0\). The complete graph has \(\kappa(K n)=n-1\). Hence, \(\kappa(K 1)=0\),

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but all other connected graphs have \(\kappa \geq 1\). Any set of vertices whose deletion disconnects \(G\) is called a separating set or vertex cut of \(G\).
The edge-connectivity of \(G\) is \(\kappa^{\prime}(G)\), the minimum number of edges whose deletion disconnects \(G\). If \(G\) has no edges, then \(\kappa^{\prime}(G)=0\). A disconnected graph does not need any edges to be deleted, and so it has \(\kappa^{\prime}=0 . K 1\) also has \(\kappa^{\prime}=0\) because it has no edges, but all other connected graphs have \(\kappa^{\prime} \geq 1\).
The edge-connectivity is always at most \(\delta(G)\), since deleting the \(\delta\) edges incident on a vertex of minimum degree will disconnect \(G\). The following inequality always holds.

\section*{THEOREM \(6.1 \kappa \leq \kappa^{\prime} \leq \delta\).}

PROOF We know that \(\kappa^{\prime} \leq \delta\). We prove that \(\kappa \leq \kappa^{\prime}\) by induction on \(\kappa^{\prime}\). If \(\kappa^{\prime}=0\), then either \(G\) has no edges, or else it is disconnected. In either case, \(\kappa=0\). Suppose that it is true whenever \(\kappa^{\prime} \leq m\), and consider \(\kappa^{\prime}=m+1\). If \(\kappa^{\prime}=|G|-1\), then \(\delta=\kappa^{\prime}\) and thus \(\kappa \leq \kappa^{\prime}\); so suppose that \(\kappa^{\prime}<|G|-1\). Let \([S, \bar{S}]\) be an edge-cut containing \(m+1\) edges. Pick any edge \(u v \in[S, \bar{S}]\) and form \(H=G-u u\). Then \([S, \bar{S}]-u v\) is an edge-cut of \(H\) containing \(m\) edges, so \(\kappa^{\prime}(H) \leq m\). By the induction hypothesis, \(\kappa(H) \leq m\). Let \(U \subseteq V(H)\) be a minimum separating set of \(H\). Then \(|U| \leq m\), and \(H-U\) consists of two or more components. We now want to put the edge uu back. Where does it go?


\section*{FI GURE 6.2}

\section*{A minimum separating set of \(H\)}

If \(H-U\) had three or more components, then \(U\) would also be a separating set of \(G\), in which case \(\kappa(G) \leq|U|=m\). If \(H-U\) has exactly two components, \(C u\) and \(C U\), containing \(u\) and \(U\), respectively, then \(U\) will not be a separating set of \(G\), for the edge \(u u\) will keep it connected. However, \(\kappa^{\prime}(G)<|G|-1\), so that \(m=\kappa\) \('-1<|G|-2\). Therefore, one of \(C u\) and \(C u\) contains two or more vertices, say \(C u\) does. Then \(U^{\prime}=U \cup\{u\}\) is a separating set of \(G\) with \(m+1\) vertices, so that \(\kappa(G) \leq \kappa^{\prime}(G)\). By induction, the theorem is true for all values of \(\kappa^{\prime}\).
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Except for this inequality, the parameters \(\kappa, \kappa^{\prime}\), and \(\delta\) are free to vary considerably, as shown by CHARTRAND and HARARY [23]. For example, the graph of Fiqure 6.3 has \(\kappa=2, \kappa^{\prime}=3\), and \(\delta=4\).


\section*{FI GURE 6.3}

\section*{A graph with \(\kappa=2, \kappa^{\prime}=3\), and \(\delta=4\)}

Given any three non-negative integers \(a, b\), and \(c\) satisfying \(a \leq b \leq c\), we can easily make a graph with \(\kappa=a, \kappa\)
\({ }^{\prime}=b\), and \(\delta=c\), as illustrated in Figure 6.3. Take two complete graphs \(G^{\prime}\) and \(G^{\prime \prime}\), isomorphic to \(K c+1\). They have minimum degree \(\delta=c\). Choose any a vertices \(U^{\prime} \subseteq V\left(G^{\prime}\right)\), and a corresponding vertices \(U^{\prime \prime} \subseteq V\left(G^{\prime \prime}\right)\). Join them up in pairs, using a edges. Then \(U^{\prime}\) is a separating set of the graph, containing a vertices. Now add \(b-\) a edges connecting \(G^{\prime}\) to \(G^{\prime \prime}\), such that every edge added has one endpoint in \(U^{\prime}\). Clearly the graph constructed has \(\kappa=a, \kappa^{\prime}=b\), and \(\delta=c\).

\section*{Exercises}
6.1.1 Let \(G\) be connected and let \(u v \in E(G)\). Prove that \(u v\) is in every spanning tree of \(G\) if and only if \(u u\) is a cut-edge of \(G\).
6.1.2 Show that a connected graph with exactly two vertices that are not cut-vertices is a tree. Hint: Consider a spanning tree of \(G\).
6.1.3 Prove that if \(G\) is a \(k\)-regular bipartite graph with \(k>1\) then \(G\) has no cut-edge.
6.1.4 Prove that if \(G\) is connected, with all even degrees, then \(\omega(G-v) \leq \frac{1}{2} \operatorname{DEG}(v)\), for any \(v \in V(G)\), where \(w(G)\) is the number of connected components of \(G\).
6.1.5 Let \(G\) be a 3 -regular graph.
(a) If \(\kappa=1\), show that \(\kappa^{\prime}=1\).
(b) If \(\kappa=2\), show that \(\kappa^{\prime}=2\).

Conclude that \(\kappa=\kappa^{\prime}\) for 3 -regular graphs.
6.1.6 Let \(G\) be a 4 -regular graph with \(\kappa=1\). Prove that \(\kappa^{\prime}=2\).
6.1.7 Let ( \(d 1, d 2, \ldots, d n\) ), where \(d 1 \leq d 2 \leq \ldots \leq d n\), be the degree sequence of a graph \(G\). Prove that if \(d j \geq j\), for \(j=1,2, \ldots, n-1-d n\), then \(G\) is connected.
6.1.8 Give another proof that \(\kappa \leq \kappa^{\prime}\), as follows. Let \([S, \bar{S}]\) be a minimum edge-cut of \(G\), containing \(\kappa^{\prime}\) edges. Construct a set \(U \subseteq S\) consisting of all vertices \(u \in S\), such

\section*{Page 122}
that there is an edge \(u v \in[S, \bar{S}]\). Then \(|U| \leq \kappa^{\prime}\). If \(U \neq S\), then \(U\) is a separating set of \(G\) with \(\leq \kappa^{\prime}\) vertices.
Therefore \(\kappa \leq \kappa^{\prime}\). Show how to complete the proof when \(U=S\).

\subsection*{6.2 Blocks}

Any graph \(G\) with \(k \geq 1\) is connected. Consequently \(G\) is said to be 1 -connected. Similarly, if \(k \geq 2\), then at least two vertices must be deleted in order to disconnect \(G\), so \(G\) is said to be 2 -connected. It is usually easier to determine a lower bound, such as \(\kappa \geq 2\) or \(k \geq 3\), than to compute the exact value of \(\kappa\). In general, \(G\) is said to be \(k\)-connected if \(k \geq k\), for some integer \(k\).
If \(G\) is a disconnected graph, then its structure is determined by its components, that is, its maximal connected subgraphs. A component which is an isolated vertex will have \(\kappa=0\), but all other components will be 1-connected.
If a connected graph \(G\) has a cut-vertex \(u\), then it is said to be separable, since deleting \(u\) separates \(G\) into two or more components. A separable graph has \(\kappa=1\), but it may have subgraphs which are 2 -connected, just as a disconnected graph has connected subgraphs. We can then find the maximal non-separable subgraphs of \(G\), just as we found the components of a disconnected graph. This is illustrated in Figure 6.4.


FIGURE 6.4

\section*{A graph (a) and its blocks (b)}

The maximal non-separable subgraphs of \(G\) are called the blocks of \(G\). The

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graph illustrated above has eight blocks, held together by cut-vertices. Every separable graph will have two or more blocks. Any 2-connected graph is non-separable. However, \(K 2\), a graph which consists of a single edge, is also non-separable, since it has no cut-vertex. Therefore every edge of \(G\) is a non-separable subgraph, and so will be contained in some maximal non-separable subgraph. Can an edge be contained in two distinct blocks? We first describe some properties of 2-connected graphs.
Notice that cycles are the smallest 2-connected graphs, since a connected graph with no cycle is a tree, which is not 2 -connected. Any two vertices \(u\) and \(u\) on a cycle \(C\) divide \(C\) into two distinct paths with only the endpoints \(u\) and \(u\) in common. Paths which have only their endpoints in common are said to be internally disjoint; see Figure 6.5.


\section*{FI GURE 6.5}

\section*{Three internally disjoint paths}

THEOREM 6.2 A graph \(G\) with three or more vertices is 2-connected if and only if every pair of vertices is connected by at least two internally disjoint paths.
PROOF \(\Leftarrow\) : Suppose that every pair of vertices of \(G\) is connected by at least two internally disjoint paths. If a vertex \(w\) is deleted, then every remaining pair of vertices is still connected by at least one path, so that \(w\) is not a cut-vertex. Therefore \(\kappa>2\).
\(\Rightarrow\) : Let \(G\) be 2-connected, and let \(v \in V(G)\). We prove by induction on \(\operatorname{DIST}(u, u)\) that \(u\) and \(u\) are connected by two internally disjoint paths. If DIST \((u, u)=1\), then \(G-u v\) is still connected, since \(\kappa^{\prime} \geq \kappa \geq 2\). Therefore \(G\) - \(u v\) contains a \(u u\)-path \(P\), so that \(G\) has two \(u u\)-paths, \(P\) and \(u u\). Suppose that the result holds when \(\operatorname{DIST}(u, u) \leq m\) and consider \(\operatorname{DIST}(u, u)=m+1\). Let \(P\) be a \(u u\)-path of length \(m+1\) and let \(w\) be the last vertex before \(u\) on this path. Then \(\operatorname{DIST}(u, w)=m\), since \(P\) is a shortest path. By the induction hypothesis, \(G\) contains internally disjoint \(u w\)-paths \(P w\) and \(Q w\).
Since \(G\) is 2 -connected, \(G-w\) is still connected, and so has a uu-path \(R\). \(R\) has the endpoint \(u\) in common with both \(P w\) and \(Q w\). Let \(x\) be the last vertex common to \(R\) and either of \(P w\) or \(Q w\), say \(x \in P_{w}\). Then \(P w[u, x] R[x, u]\) and
page_123
Page 124
\(u\)


\section*{FI GURE 6.6}

Internally disjoint paths Pwand \(Q u\)
Qwwu are two internally disjoint uu-paths. By induction, the result holds for all pairs \(u, u\) of vertices.
So in a 2-connected graph, every pair \(u\), \(u\), of vertices are connected by at least two internally disjoint paths \(P\) and \(Q\). Since \(P\) and \(Q\) together form a cycle, we know that every pair of vertices lies on a cycle. Another consequence of this theorem is that every pair of edges also lies on a cycle.
COROLLARY 6.3 A graph \(G\) with three or more vertices is 2-connected if and only if every pair of edges lies on a cycle.

PROOF Let \(G\) be 2 -connected and pick edges \(u u, x y \in E(G)\). Subdivide \(u u\) with a new vertex \(w\), and \(x y\) with a new vertex \(z\) to get a graph \(G^{\prime}\). Now \(G\) has no cut-vertex, so neither does \(G^{\prime}\). By the previous theorem, \(w\) and \(z\) lie on a cycle in \(G^{\prime}\), so that \(u u\) and \(x y\) lie on a cycle in \(G\).
\(\Leftarrow\) : Suppose now that every pair of edges lies on a cycle. Then every vertex has degree at least two, since no cycle could pass through an edge incident on a vertex of degree one. Choose any two vertices \(u\) and \(x\). Choose any vertex \(v \longrightarrow u\) and a vertex \(y \longrightarrow x\), such that \(y \neq u\). Then the edges \(u u\) and \(x y\) must lie on a cycle \(C\). Clearly \(C\) contains \(u\) and \(x\), so that every pair \(u, x\), of vertices lies on a cycle. It follows that \(G\) is 2connected.
LEMMA 6.4 Each edge uu of \(G\) is contained in a unique block.
PROOF Let \(u u\) be an edge in a graph \(G\), and let \(B\) be a maximal 2-connected subgraph of \(G\) containing \(u u\). If \(B^{\prime}\) is another maximal 2-connected subgraph

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FI GURE 6.7
Two edges on a cycle
containing \(u v\), where \(B \neq B^{\prime}\), then choose any edge \(x y \in B^{\prime}-B \cdot B^{\prime}\) contains a cycle \(C\) containing both \(u u\) and \(x y\), since \(B^{\prime}\) is 2-connected. The subgraph \(B \cup C\) is 2-connected, and is larger than \(B\), a contradiction. Therefore, each edge \(u u\) is contained in exactly one block of \(G\).

\subsection*{6.3 Finding the blocks of a graph}

The first algorithm to find the blocks of a graph was discovered by READ [99]. It uses the fundamental cycles with respect to a spanning tree \(T\). Because each edge of \(G\) is contained in a unique block Buu, the algorithm begins by initializing Buu to contain only \(u v\) and uses the merge-find data structure to construct the full blocks Buu. For each edge \(u v \notin T\), the fundamental cycle Cuu is found. Because Cuu is 2-connected, all its edges are in one block. So upon finding Cuu, we merge all the blocks \(B x y\), where \(x y \in C_{u v}\), into one. Any spanning tree \(T\) can be used. If we choose a breadth-first tree, we have Algorithm 6.3.1.
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Algorithm 6.3.1: BLOCKS(G)
comment: \(G\) is a connected graph
for each \(u v \in E(G)\)
do initialize Buu to contain uu
pick any vertex \(x \in V(G)\)
place \(x\) on ScanQ
repeat
select \(u\) from head of ScanQ
for each \(v \longrightarrow u\)
do if \(v \notin S c a n Q\)
then \(\left\{\begin{array}{l}\text { comment: } u v \text { forms part of the spanning tree } T \\ \text { add edge } u v \text { to } T \\ \text { add } v \text { to } S c a n Q\end{array}\right.\)
else \(\left\{\begin{array}{l}\text { comment: } u v \text { creates a fundamental cycle } \\ \text { construct } C_{u v} \\ \text { for each edge } x y \in C_{u v} \\ \text { do } B_{u v} \leftarrow B_{u v} \cup B_{x y}\end{array}\right.\) until all vertices on ScanQ are processed
comment: each Buu now consists of the unique block containing uu
LEMMA 6.5 At the beginning of each iteration of the repeat loop, each Buvis either a single edge, or else is 2connected.
PROOF The proof is by induction on the number of iterations of the repeat loop. At the beginning of the first iteration it is certainly true. Suppose that it is true at the beginning of the kth iteration. If the edge uu chosen forms part of the spanning tree \(T\), it will also be true for the \((k+1)\) st iteration, so suppose that \(u u\) creates a fundamental cycle Cuu. Each Bxy for which \(x y \in C_{u v}\) is either a single edge, or else 2-connected. The new Buu is formed by merging all the \(B x y\) into one, say \(B_{u v}=B_{x_{1} y_{1}} \cup B_{x_{2} y_{2}} \cup \cdots \cup B_{x_{m} y_{m}}\). Pick any two edges \(a b, c d \in B_{u v}\), say \(a b \in B_{x_{i} y_{i}}\) and \(c d \in B_{x_{j} y_{j}}\). Weshow that Buu contains a cycle containing both ab and \(c d\). If \(a b, c d \in C_{u v}\), it is certainly true. Otherwise, notice that each \(B_{x_{i} y_{i}}\) contains some edge of Cuu. Without loss of generality, we can suppose that the edges \(x_{l} y_{l} \in C_{u v}\), for \(l=1,2, \ldots, m\). Since \(B_{x_{i} y_{i}}\) is 2-connected, it contains a cycle \(C i\) containing both xiyi and ab, and \(B_{x_{j} y_{j}}\) contains a cycle \(C j\) containing both xiyj and cd. This is illustrated in Figure 6.8. Then \(C_{u v} \oplus C_{i} \oplus C_{j}\) is a cycle contained in Buu and containing ab and cd. By Corollary 6.3, the new
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Buu is 2-connected. Therefore the result is true at the beginning of the \((k+1)\) st iteration. By induction it is true for all iterations.
COROLLARY 6.6 Upon completion of Algorithm 6.3.1, each Buucontains the unique block containing uu.
PROOF By the previous lemma, each Buu will either be a single edge, or else 2-connected. If Buu is not the unique block \(B\) containing uu, then pick some edge \(x y \in B-B_{u v} \cdot B\) contains a cycle \(C\) containing both uu and \(x y\). By Theorem 4.3, \(C\) can be written in terms of fundamental cycles with respect to the spanning tree \(T\) constructed by Algorithm 6.3.1, \(C=C_{u_{1} v_{1}} \oplus C_{u_{2} v_{2}} \oplus \cdots \oplus C_{u_{m} v_{m}}\). But each of the fundamental cycles \(C_{u_{i} v_{i}}\) will have been processed by the algorithm, so that all edges of \(C\) are contained in one Buu, a contradiction. Therefore, each Buu consists of the unique block containing uu.


FI GURE 6.8
Merging the \(B x y\)

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\section*{Exercises}
6.3.1 Given an edge uu which creates a fundamental cycle Cuu, describe how to find Cuu using the Parent[•] array created by the BFS.
6.3.2 Let ( \(d 1, d 2, \ldots, d n\) ), where \(d 1 \leq d 2 \leq \ldots \leq d n\), be the degree sequence of a graph \(G\). Prove that if \(d j \geq j+1\), for \(j=1,2, \ldots, n-1-d n-1\), then \(G\) is 2 -connected.
6.3.3 Program the BLOCKS() algorithm. One way to store the merge-find sets Buu is as an \(n\) by \(n\) matrix BlockRep \([\cdot \cdot \cdot]\). Then the two values \(x \leftarrow\) BlockRep \([u, u]\) and \(y \leftarrow B l o c k R e p[u, u]\) together define the edge \(x y\) representing uu. Another way is to assign a numbering to the edges, and use a linear array.
6.3.4 Try to estimate the complexity of the algorithm BLOCKS(). It is difficult to obtain a close estimate because it depends on the sum of the lengths of all \(\varepsilon-n+1\) fundamental cycles of \(G\), where \(n=|G|\).
6.3.5 The Block-Cut-Vertex Tree. (See HARARY [59].) Let \(G\) be a connected separable graph. Let \(B\) denote the set of blocks of \(G\) and \(C\) denote the set of cut-vertices. Each cut-vertex is contained in two or more blocks, and each block contains one or more cut-vertices. We can form a bipartite graph \(B C(G)\) with vertex-set \(\mathcal{B} \cup \mathcal{C}\) by joining each \(B \in \mathcal{B}\) to the cut-vertices \(v \in \mathcal{C}\) that it contains.
(a) Show that \(B C(G)\) has no cycles, and consequently is a tree.
(b) In the block-cut-vertex tree \(B C(G)\), the degree of each \(v \in \mathcal{C}\) is the number of blocks of \(G\) containing \(u\).

Denote this value by \(b(u)\), for any vertex \(v \in V(G)\). Show that
\[
\sum_{v \in V(G)} b(v)-1=\sum_{v \in \mathcal{C}} b(v)-1=|\mathcal{B}|-1,
\]
so that the number of blocks of \(G\) is given by
\[
|\mathcal{B}|=1+\sum_{v \in V(G)} b(v)-1
\]
(c) Prove that every separable graph has at least two blocks which contain only one cut-vertex each.

\subsection*{6.4 The depth-first search}

There is an easier, more efficient way of finding the blocks of a graph than using fundamental cycles. It was discovered by Hopcroft and Tarjan. It uses a depth-first search (DFS) to construct a spanning tree. With a depth-first search the fundamental cycles take a very simple form-essentially we find them for free, as they require no extra work. The basic form of the depth-first search follows. It is a recursive procedure, usually organized with several global variables initialized by the calling procedure. The example following uses a global counter DFCount, and two arrays DFNum[u] and Parent[u]. Each vertex \(u\) is assigned

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a number DFNum[U], being the order in which the DFS visits the vertices of \(G\), and a value Parent[U], being the vertex \(u\) from which \(\operatorname{DFS}(u)\) was called. It is the parent of \(u\) in the rooted spanning tree constructed.
Initially all DFNum[•] values are set to 0 .
Algorithm 6.4.1: DFS(u)
comment: extend a depth-first search from vertex \(u\)
DFCount \(\leftarrow\) DFCount+1
DFNum[u] \(\leftarrow\) DFCount
for each \(v \longrightarrow u\)
do if \(D F N u m\lceil u 7=0\)
then \(\left\{\begin{array}{l}\text { comment: } v \text { is not visited yet } \\ \text { add edge } u v \text { to the spanning tree } \\ \text { Parent }[v] \leftarrow u \\ \operatorname{DFS}(v)\end{array}\right.\)
else \{comment: uu creates a fundamental cycle The calling procedure can be written as follows:
DFCount \(\leftarrow 0\)
for \(u \leftarrow 1\) to \(n\) do \(\operatorname{DFNum}[u] \leftarrow 0\)
select a staring vertex \(u\)
DFS(u)
Figure 6.9 shows a depth-first search in a graph. The numbering of the vertices shown is that of DFNum[•], the order in which the vertices are visited.
Notice that while visiting vertex \(u\), \(\operatorname{DFS}(u)\) is called immediately, for each \(v \longrightarrow u\) discovered. This means that before returning to node \(u\), all vertices that can be reached from \(u\) on paths that do not contain \(u\) will be visited; that is, all nodes of \(G-u\) that are reachable from \(u\) will be visited. We state this important property as a lemma. For any vertex \(u\), let \(A(u)\) denote \(u\), plus all ancestors of \(u\) in the depth-first tree, where an ancestor of \(u\) is either its parent, or any vertex on the unique spanning tree path from \(u\) to the root vertex.
LEMMA 6.7 Suppose that DFS (U) is called while visiting node \(u\). Then DFS (U) visits every vertex in \(V(G)-A(u)\) reachable from \(u\) before returning to node \(u\).
PROOF The statement if \(\operatorname{DFNum}[u]=0\) then...
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\section*{FI GURE 6.9}

\section*{A depth-first search}
ensures that no vertex of \(A(u)\) will be visited before returning to node \(u\). To show that every vertex of \(G-A(u)\) connected to \(u\) is visited, let \(w\) be a vertex of \(V(G)-A(u)\), with \(\operatorname{DIST}(u, w)=k\). The proof is by induction on \(k\). It is clear that all \(w \longrightarrow v\) will be visited before returning to \(u\), so that the statement is true when \(k=1\). If \(k>1\), let \(P\) be a \(u w\)-path of length \(k\), and let \(x \longrightarrow v\) be the first vertex of \(P\). Now \(x\) will certainly be visited before returning to node \(u\). When \(x\) is visited, either \(w\) will already have been visited, or else some DFS(y) called from node \(x\) will visit \(w\) before returning to \(u\), since \(\operatorname{DIST}(x, w)=k-1\). Therefore all vertices of \(G-u\) connected to \(u\) will be visited before returning to \(u\).

This makes it possible to detect when \(u\) is a cut-vertex. It also means that a spanning tree constructed by a depth-first search will tend to have few, but long, branches. The following diagram shows the DF-tree constructed by the DFS above.
Suppose that while visiting node \(u\), a vertex \(v \longrightarrow u\) with DFNum [u] \(\neq 0\) is encountered. This means that \(u\) has already been visited, either previously to \(u\), or as a descendant of \(u\). While visiting \(u\), the edge \(u u\) will have been encountered. Therefore, if \(u\) was visited previously, either DFS( \(u\) ) was called from node \(u\), so that Parent[u]=u, or else DFS( \(u\) ) was called from some descendant \(w\) of \(u\). So we can state the following fundamental property of depth-first searches:
LEMMA 6.8 Suppose that while visiting vertex \(u\) in a depth-first search, edge uu creating a fundamental cycle is encountered. Then either \(u\) is an ancestor of \(u\), or else \(u\) is an ancestor of \(u\).
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FI GURE 6.10

\section*{A depth-first tree with fronds}

Edges which create fundamental cycles with respect to a depth-first spanning tree \(T\) are called fronds of \(T\). Some of the fronds of the graph of Figure 6.9 are shown in Figure 6.10.
Now suppose that \(G\) is a separable graph with a cut-vertex \(u\). \(u\) will occur somewhere in \(T\), say that DFS( \(u\) ) was called from node \(u\), and that node \(u\) in turn calls DFS \((w)\), where \(u\) and \(w\) are in different blocks. Thus, edges \(u u\) and \(u w\) do not lie on any cycle of \(G\). This is illustrated in Figure 6.11. Consider any descendant \(x\) of \(w\) and a frond \(x y\) discovered while visiting node \(x\), where \(y\) is an ancestor of \(x\). Now \(y\) cannot be an ancestor of \(u\), for then the fundamental cycle Cxy would contain both edges \(u u\) and \(u w\), which is impossible. Therefore either \(y=u\), or else \(y\) is a descendant of \(u\). So, for every frond \(x y\), where \(x\) is a descendant of \(w\), either \(y=u\), or else \(y\) is a descendant of \(u\). We can recognize cut-vertices during a DFS in this way. Ancestors of \(u\) will have smaller DFNum[•] values than \(u\), and descendants will have larger values. For each vertex \(u\), we need to consider the endpoints \(y\) of all fronds \(x y\) such that \(x\) is a descendant of \(u\).
DEFINITION 6.1: Given a depth-first search in a graph G. The low-point of a vertex \(u\) is LowPt[u], the minimum value of DFNum[ \(y\) ], for all edges \(u y\) and all fronds \(x y\), where \(x\) is a descendant of \(u\).
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FI GURE 6.11
A cut-vertex \(u\) in a DF-tree
We can now easily modify DFS ( \(u\) ) to compute low-points and find cut-vertices. In addition to the global variables DFCount, DFNum[ \(\cdot]\), and Parent [•], the algorithm keeps a stack of edges. Every edge encountered by the algorithm is placed on the stack. When a cut-vertex is discovered, edges on the top of the stack will be the edges of a block of \(G\).
The procedure DFSEARCH \((u)\) considers all \(v \longrightarrow u\). If \(u\) has been previously visited, there are two possibilities, either \(u=\) Parent[ \(u\) ], or else \(u v\) is a frond. When \(u u\) is a frond, there are also two possibilities, either \(u\) is an ancestor of \(u\), or \(u\) is an ancestor of \(u\). These two cases are shown in Figure 6.12. The algorithm only needs those fronds \(u u\) for which \(u\) is an ancestor of \(u\) in order to compute LowPt[ \(u\) ].
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Algorithm 6.4.2: DFBLOCKS(G)
comment: \(\left\{\begin{array}{l}\text { DFS to find the blocks of a connected graph } G, \\ \text { on } n \text { vertices. }\end{array}\right.\)
procedure DFSEARCH(u)
comment: extend a depth-first search from \(u\)
DFCount \(\leftarrow\) DFCount+1
DFNum[u] \(\leftarrow D F C o u n t\)
for each \(v \longrightarrow u\) do
(if DFNum \([v]=0\)
then \(\left\{\begin{array}{l}\text { comment: }\left\{\begin{array}{l}v \text { is not visited yet, } \\
\text { add } u v \text { to the spanning tree }\end{array}\right. \\
\text { Parent }[v] \leftarrow u \\
\text { stack } u v \\
\operatorname{LowPt}[v] \leftarrow \text { DFNum }[u] \quad \text { "initial value" } \\
\text { DFSEARCH }(v) \\
\text { comment: LowPt }[v] \text { is now known } \\
\text { if LowPt }[v]=\text { DFNum }[u]\end{array}\right\}\)\begin{tabular}{l} 
then \(\left\{\begin{array}{l}\text { comment: } u \text { is a cut-vertex } \\
\text { unstack all edges up to, and including, } u v\end{array}\right.\) \\
else \(\left\{\begin{array}{l}\text { comment: otherwise } \begin{array}{l}\text { cowPt }[v]<\text { DFNum }[u] \\
\text { if } \text { LowPt }[v]<L o w P t[u] \\
\text { then LowPt }[u] \leftarrow \text { LowPt }[v]\end{array}\end{array}\right.\)
\end{tabular}

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\section*{FI GURE 6.12}

\section*{Two kinds of fronds}

The next thing to notice is that the LowPt[u] is correctly computed. For if \(u\) is a leaf-node of the search tree, then all edges \(u u\) are fronds, and \(u\) is an ancestor of \(u\). The algorithm computes LowPt[ \(u\) ] as the minimum DFNum[u], for all such \(u\). Therefore, if \(u\) is a leaf-node, the low-point is correctly computed. We can now use induction on the depth of the recursion. If \(u\) is not a leaf-node, then some DFSEARCH( \(u\) ) will be called from node \(u\). The depth of the recursion from \(u\) will be less than that from \(u\), so that we can assume that DFSEARCH(U) will correctly compute LowPt[U]. Upon returning from this recursive call, LowPt[U] is compared with the current value of LowPt[u], and the minimum is taken, for every unsearched \(v \longrightarrow u\). Therefore, after visiting node \(u\), LowPt[u] will always have the correct value.
So far the algorithm computes low-points and uses them to find the cut-vertices of G. We still need to find the edges in each block. While visiting node \(u\), all new edges \(u u\) are stacked. If it is discovered that LowPt[u]=DFNum[u]; so that \(u\) is a cut-vertex, then the edges in the block containing uu are all those edges on the stack up to, and including, uu.
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THEOREM 6.9 Each time that
\[
\operatorname{LowPt}[\mathrm{U}]=\text { DFNum }[u]
\]
occurs in DFSEARCH(u), the block containing uu consists of those edges on the stack up to and including uu. PROOF Let Buu denote the block containing uu. The proof is by induction on the number of times that LowPt[u]=DFNum[u] occurs. Consider the first time it occurs. DFSEARCH(u) has just been called, while visiting node \(u\). Edge \(u u\) has been placed on the stack. DFSEARCH(u) constructs the branch of the search tree at \(u\) containing \(u\). By Lemma 6.7, this contains all vertices of \(G-u\) connected to \(u\). Call this set of vertices \(B\). By the definition of the low-point, there are no fronds joining \(u\) or any de scendant of \(u\) to any ancestor of \(u\). So \(u\) separates \(B\) from the rest of the graph. Therefore \(B_{u v} \subseteq G[B \cup\{u\}]\).
Suppose now that \(B\) contained a cut-vertex w. No leaf-node of the DF-tree can be a cut-vertex, so some \(\operatorname{DFSEARCH}(x)\) is called while visiting node \(w\). It will visit all nodes of \(G-w\) connected to \(x\). Upon returning to node \(w\), it would find that LowPt \([x]=\operatorname{DFNum}[w]\), which is impossible, since this occurs for the first time at node \(u\). Therefore Buu consists of exactly those edges encountered while performing DFSEARCH(u); that is, those on the stack.
Upon returning from \(\operatorname{DFS}(U)\) and detecting that \(\operatorname{LowPt}[u]=\operatorname{DFNum}[u]\), all edges on the stack will be unstacked up to, and including, uu. This is equivalent to removing all edges of Buu from the graph. The remainder of the DFS now continues to work on \(G-B\). Now Buu is an end-block of \(G\) (i.e., it has at most one cut-vertex) for \(u\) is the first vertex for which a block is detected. If \(G\) is 2 -connected, then \(G=B u u\), and the algorithm is finished. Otherwise, \(G-B\) is connected, and consists of the remaining blocks of \(G\). It has one less block than \(G\), so that each time LowPt[ \(u]=\operatorname{DFNum}[u]\) occurs in \(G-B\), the edges on the stack will contain another block. By induction, the algorithm finds all blocks of \(G\).

Each time the condition LowPt[u]=DFNum[u] occurs, the algorithm has found the edges of a block of \(G\). In this case, \(u\) will usually be a cut-vertex of \(G\). The exception is when \(u\) is the root of the DF-tree, since \(u\) has no ancestors in the tree. Exercise 6.4.3 shows how to deal with this situation.

\subsection*{6.4.1 Complexity}

The complexity of \(\operatorname{DFBLOCKS}()\) is very easy to work out. For every \(u \in V(G)\), all \(v \longrightarrow u\) are considered. This takes
\[
\sum_{u} \operatorname{DEG}(u)=2 \varepsilon
\]
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steps. Each edge is stacked and later unstacked, and a number of comparisons are performed in order to compute the low-points. So the complexity is \(O(\varepsilon)\).

\section*{Exercises}
6.4.1 Can the spanning tree shown in the graph illustrated in Figure 6.13 be a DF-tree, with the given rootnode? If so, assign a DF-numbering to the vertices.


\section*{FI GURE 6.13}

\section*{I s this a DF spanning tree?}
6.4.2 Program the DFBLOCKS() algorithm to find all the blocks of a connected graph G. Print a list of the edges in each block. Choose the starting vertex to be sometimes a cut-vertex, sometimes not.
6.4.3 Modify your program to print also a list of the cut-vertices of \(G\), by storing them on an array. A vertex \(u\) is a cut-vertex if LowPt[u]=DFNum[u] occurs while visiting edge \(u u\) at node \(u\). However, if \(u\) is the root-node of the DF-tree, then it will also satisfy this condition, even when \(G\) is 2 -connected. Find a way to modify the algorithm so that it correctly determines when the root-node is a cut-vertex.
6.4.4 A separable graph has \(\kappa=1\), but can be decomposed into blocks, its maximal non-separable subgraphs.

A tree is the only separable graph which does not have a 2 -connected subgraph, so that every block of a tree is an edge. Suppose that \(G\) has \(\kappa=2\). In general, \(G\) may have 3 -connected subgraphs. Characterize the class of 2-connected graphs which do not have any 3-connected subgraph.
6.4.5 Let \(G\) be 3-connected. Prove that every pair of vertices is connected by at least three internally disjoint paths.
6.4.6 Let \(G\) have \(\kappa=2\), and consider the problem of finding all separating pairs \(\{u, u\}\) of \(G u s i n g\) a DFS. Prove that for every separating pair \(\{u, u\}\), one of \(u\) and \(u\) is an ancestor of the other in any DF-tree. Refer to Figure 6.14 .
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\(v\)

\(v\)

\section*{FI GURE 6.14}

DFS with separating set \(\{u, u\}\)
6.4.7 Suppose that deleting \(\{u, u\}\) separates \(G\) into two or more components. Let \(G 1\) denote one component and \(G 2\) the rest of \(G\). Show that there are two possible ways in which a DFS may visit \(u\) and \(u\), as illustrated in Figure 6.14. Devise a DFS which will find all pairs \(\{u, u\}\) which are of the first type. (Hint: You will need LowPt2 [U], the second low-point of \(u\). Define it and prove that it works.)

\subsection*{6.5 Notes}

The example of Figure 6.3 is based on HARARY [59]. Read's algorithm to find the blocks of a graph is from READ [99]. The depth-first search algorithm is from HOPCROFT and TARJ AN [67]. See also TARJ AN [111]. Hopcroft and Tarjan's application of the depth-first search to find the blocks of a graph was a great breakthrough in algorithmic graph theory. The depth-first search has since been applied to solve a number of difficult problems, such as determining whether a graph is planar in linear time, and finding the 3-connected components of a graph.
Algorithms for finding the connectivity and edge-connectivity of a graph are described in Chapter 8. An excellent reference for connectivity is TUTTE [120], which includes a detailed description of the 3-connected components of a graph. A depth-first seach algorithm to find the 3-connected components of a graph can be found in HOPCROFT and TARJ AN [68].
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7
Alternating Paths and Matchings

\subsection*{7.1 I ntroduction}

Matchings arise in a variety of situations as assignment problems, in which pairs of items are to be matched together, for example, if people are to be assigned jobs, if sports teams are to matched in a tournament, if tasks are to be assigned to processors in a computer, whenever objects or people are to be matched on a one-to-one basis.
In a graph \(G\), a matching \(M\) is a set of edges such that no two edges of \(M\) have a vertex in common. Figure 7.1 illustrates two matchings \(M 1\) and \(M 2\) in a graph \(G\).


\(M_{2}\)

FI GURE 7.1
Matchings
Let \(M\) have \(m\) edges. Then 2 m vertices of \(G\) are matched by \(M\). We also page_139

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say that a vertex \(u\) is saturated by \(M\) if it is matched, and unsaturated if it is not matched. In general, we want \(M\) to have as many edges as possible.
DEFINITION 7.1: \(M\) is a maximum matching in \(G\) if no matching of \(G\) has more edges.
For example, in Figure 7.1, \(|M 1|=3\) and \(|M 2|=4\). Since \(|G|=8, M 2\) is a maximum matching. A matching which saturates every vertex is called a perfect matching. Obviously a perfect matching is always a maximum matching. \(M 1\) is not a maximum matching, but it is a maximal matching; namely \(M 1\) cannot be extended by the addition of any edge \(u \cup\) of \(G\). However, there is a way to build a bigger matching out of \(M 1\). Let \(P\) denote the path ( \(u 1, u 2, \ldots, u 6)\) in Figure 7.1.

DEFINITION 7.2: Let \(G\) have a matching \(M\). An alternating path \(P\) with respect to \(M\) is any path whose edges are alternately in \(M\) and not in \(M\). If the endpoints of \(P\) are unsaturated, then \(P\) is an augmenting path. So \(P=(u 1, u 2, \ldots, u 6)\) is an augmenting path with respect to \(M 1\). Consider the subgraph formed by the exclusive or operation \(M=M_{1} \oplus E(P)\) (also called the symmetric difference, \(\left.\left(M_{1}-E(P)\right) \cup\left(E(P)-M_{1}\right)\right)\). \(M\) contains those edges of \(P\) which are not in \(M 1\), namely, \(u 1 u 2, u 3 u 4\), and \(u 5 u 6 . ~ M\) is a bigger matching than M1. Notice that \(M=M 2\).
LEMMA 7.1 Let \(G\) have a matching \(M\). Let \(P\) be an augmenting path with respect to \(M\). Then \(M^{\prime}=M \oplus E(P)\) is a matching with one more edge than \(M\).
PROOF Let the endpoints of \(P\) be \(u\) and \(u . M^{\prime}\) has one more edge than \(M\), since \(u\) and \(u\) are unsaturated in \(M\), but saturated in \(M^{\prime}\). All other vertices that were saturated in \(M\) are still saturated in \(M^{\prime}\). So \(M^{\prime}\) is a matching with one more edge.
The key result in the theory of matchings is the following:
THEOREM 7.2 (Berge's theorem) A matching \(M\) in \(G\) is maximum if and only if \(G\) contains no augmenting path with respect to \(M\).
PROOF \(\Rightarrow\) : If \(M\) were a maximum matching and \(P\) an augmenting path, then \(M \oplus E(P)\) would be a larger matching. So there can be no augmenting path if \(M\) is maximum.
\(\Leftarrow\) : Suppose that \(G\) has no augmenting path with respect to \(M\). If \(M\) is not maximum, then pick a maximum matching \(M^{\prime}\). Clearly \(\left|M^{\prime}\right|>|M|\). Let \(H=M \oplus M^{\prime}\). Consider the subgraph of \(G\) that \(H\) defines. Each vertex \(u\) is incident on at most one \(M\)-edge and one \(M^{\prime}\)-edge, so that in \(H, \operatorname{DEG}(U) \leq 2\). Every path in

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\(H\) alternates between \(M\)-edges and \(M^{\prime}\)-edges. So \(H\) consists of alternating paths and cycles, as illustrated in Figure 7.2.


FI GURE 7.2

\section*{Alternating paths and cycles}

Each cycle must clearly have even length, with an equal number of edges of \(M\) and \(M^{\prime}\). Since \(\left|M^{\prime}\right|>|M|\), some path \(P\) must have more \(M^{\prime}\)-edges than \(M\)-edges. It can only begin and end with an \(M^{\prime}\)-edge, so that \(P\) is augmenting with respect to \(M\). But we began by assuming that \(G\) has no augmenting path for \(M\). Consequently, \(M\) was initially a maximum matching.
This theorem tells us how to find a maximum matching in a graph. We begin with some matching \(M\). If \(M\) is not maximum, there will be an unsaturated vertex \(u\). We then follow alternating paths from \(u\). If some unsaturated vertex \(u\) is reached on an alternating path \(P\), then \(P\) is an augmenting uu-path. Set
\(M \leftarrow M \oplus E(P)\), and repeat. If the method that we have chosen to follow alternating paths is sure to find all such paths, then this technique is guaranteed to find a maximum matching in \(G\).
In bipartite graphs it is slightly easier to follow alternating paths and therefore to find maximum matchings, because of their special properties. Let \(G\) have bipartition ( \(X, Y\) ). If \(S \subseteq X\), then the neighbor set of \(S\) is \(N(S)\), the set of \(Y\)-vertices adjacent to \(S\). Sometimes \(N(S)\) is called the shadow set of \(S\). If \(G\) has a perfect matching \(M\), then every \(x \in S\) will be matched to some \(y \in Y\) so that \(|N(S)| \geq|S|\), for every \(S \subseteq X\). HALL [58] proved that this necessary condition is also sufficient.
THEOREM 7.3 (Hall's theorem) Let \(G\) have bipartition \((X, Y) . G\) has a matching saturating every \(x \in X\) if and only if \(|N(S)| \geq|S|\), for all \(S \subseteq X\).
PROOF We have all ready discussed the necessity of the conditions. For the converse suppose that |N(S)| \(\geq|S|\), for all \(S \subseteq X\). If \(M\) does not saturate all of


\section*{FI GURE 7.3}

\section*{The neighbor set}
\(X\), pick an unsaturated \(u \in X\), and follow all the alternating paths beginning at \(u\). (See Figure 7.4.)


\section*{FI GURE 7.4}

\section*{Follow alternating paths}

Let \(S \subseteq X\) be the set of \(X\)-vertices reachable from \(u\) on alternating paths, and let \(T\) be the set of \(Y\)-vertices reachable. With the exception of \(u\), each vertex \(x \in S\) is matched to some \(y \in T\), for \(S\) was constructed by extending alternating paths from \(y \in T\) to \(x \in S\) whenever \(x y\) is a matching edge. Therefore \(|S|=|T|+1\). Now there may be other vertices \(X-S\) and \(Y-T\). However, there can be no edges [ \(S, Y-T\) ], for such an edge would extend an alternating path to a vertex of \(Y-T\), which is not reachable from \(u\) on an alternating path. So every \(x \in S\) can
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only be joined to vertices of \(T\); that is, \(T=N(S)\). It follows that \(|S|>|N(S)|\), a contradiction. Therefore every vertex of \(X\) must be saturated by \(M\).
COROLLARY 7.4 Every \(k\)-regular bipartite graph has a perfect matching, if \(k>0\).
PROOF Let \(G\) have bipartition \((X, Y)\). Since \(G\) is \(k\)-regular, \(\varepsilon=k \cdot|X|=k \cdot|Y|\), so that \(|X|=|Y|\). Pick any \(S \subseteq X\). How many edges have one end in \(S\) ? Exactly \(k \cdot|S|\). They all have their other end in \(N(S)\). The number of edges with one endpoint in \(N(S)\) is \(k \cdot|N(S)|\). So \(k \cdot|S| \leq k-|N(S)|\), or \(|S| \leq|N(S)|\), for all \(S \subseteq X\). Therefore \(G\) has a perfect matching.

\section*{Exercises}
7.1.1 Find a formula for the number of perfect matchings of \(K 2 n\) and \(K n, n\).
7.1.2 (Hall's theorem.) Let \(A 1, A 2, \ldots, A n\) be subsets of a set \(S\). A system of distinct representatives for the family \(\{A 1, A 2, \ldots, A n\}\) is a subset \(\{a 1, a 2, \ldots, a n\}\) of \(S\) such that \(a_{1} \in A_{1}, a_{2} \in A_{2}, \ldots, a_{m} \in A_{m}\), and \(a i \neq a j\), for \(i \neq j\). Example:
A1=students taking computer science 421
\(A 2=\) students taking physics 374
\(A 3=s t u d e n t s\) taking botany 464
A4 =students taking philosophy 221
The sets \(A 1, A 2, A 3, A 4\) may have many students in common. Find four distinct students a1, a2, a3, a4, such
that \(a_{1} \in A_{1}, a_{2} \in A_{2}, a_{3} \in A_{3}\), and \(a_{4} \in A_{4}\) to represent each of the four classes.
Show that \(\{A 1, A 2, \ldots, A n\}\) has a system of distinct representatives if and only if the union of every combination of \(k\) of the subsets \(A i\) contains at least \(k\) elements, for all \(k=1,2, \ldots, n\). (Hint: Make a bipartite graph \(A 1, A 2, \ldots, A n\) versus all \(a_{j} \in S\), and use Hall's theorem.)

\subsection*{7.2 The Hungarian algorithm}

We are now in a position to construct an algorithm which finds a maximum matching in bipartite graphs, by following alternating paths from each unsaturated \(u \in X\). How can we best follow alternating paths? Let \(n=|G|\). Suppose that we store the matching as an integer array Match \([x], x=1,2, \ldots, n\), where Match [ \(x\) ] is the vertex matched to \(x\) (so Match [Match [x]) \(=x\), if \(x\) is saturated).

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We use Match \([x]=0\) to indicate that \(x\) is unsaturated. We could use either a DFS or BFS to construct the alternating paths. A DFS is slightly easier to program, but a BF-tree tends to be shallower than a DF-tree, so that a BFS will likely find augmenting paths more quickly, and find shorter augmenting paths, too. Therefore the BFS is used for matching algorithms.
The array used to represent parents in the BF-tree can be used in combination with the Match[•] array to store the alternating paths. We write PrevPt[u] for the parent of \(u\) in a BF-tree. It is the previous point to \(u\) on an alternating path to the root. This is illustrated in Figure 7.5.


FI GURE 7.5

\section*{Storing the alternating paths}

We also need to build the sets \(S\) and \(N(S)\) as queues, which we store as the arrays ScanQ and \(N S\), respectively. The algorithm for finding a maximum matching in bipartite graphs is Algorithm 7.2.1. It is also called the Hungarian algorithm for maximum matchings in bipartite graphs.
comment: \(u\) is currently unsaturated
delete \(S\) and \(N(S)\) from the graph
1:
comment: Match[•] now contains a maximum matching
Notice that the algorithm needs to be able to determine whether \(y \in N S\). This can be done by storing a
boolean array. Another possibility is to use PrevPt[U]=0 to indicate that \(v \notin N(S)\). We can test if \(y\) is
unsaturated by checking whether Match[y]=0. AUGMENT(y) is a procedure that computes \(M \leftarrow M \oplus E(P)\),
where \(P\) is the augmenting path found. Beginning at vertex \(y\), it alternately follows PrevPt[•] and Match[•]
back to the initial unsaturated vertex, which is the root-node of the BF-tree being constructed. This is
illustrated in Figure 7.6.
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Algorithm 7.2.2: AUGMENT (y)
comment: follow the augmenting path, setting \(M \leftarrow M \oplus E(P)\) repeat
\(w \leftarrow\) PreuPt[y]
Match \([y] \leftarrow w\)
\(u \leftarrow\) Match [w]
Match \([w] \leftarrow y\)
\(y \leftarrow u\)
until \(y=0\)


\section*{FI GURE 7.6}

\section*{Augmenting the matching}

The BFS constructs an alternating search tree. It contains all vertices reachable from the root-node \(u\) on alternating paths. Vertices at even distance from \(u\) in the tree form the set \(S\), and those at odd distance form \(N(S)\). The vertices of \(S\) are sometimes called outer vertices, and those of \(N(S)\) inner vertices. All the actual searching is done from the outer vertices.
THEOREM 7.5 The Hungarian algorithm constructs a maximum matching in a bipartite graph.
PROOF Let \(G\) have bipartition \((X, Y)\). If the algorithm saturates every vertex of \(X\), then it is certainly a maximum matching. Otherwise some vertex \(u\) is not matched. If there is an augmenting path \(P\) from \(u\), it must alternate between \(X\) and \(Y\), since \(G\) is bipartite. The algorithm constructs the sets \(S\) and \(N(S)\), consisting of all vertices of \(X\) and \(Y\), respectively, that can be reached on alternating paths. So \(P\) will be found if it exists. If \(u\) cannot be saturated, then we know that \(|S|=|N(S)|+1\). Every vertex of \(S\) but \(u\) is matched. \(S\) and \(N(S)\) are then deleted from the graph. Does the deletion of these vertices affect the rest of the page_146

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FI GURE 7.7

\section*{The alternating search tree}
algorithm? As in Hall's theorem, there are no edges [ \(S, Y-N(S)\) ]. Suppose that alternating paths from a vertex \(v \in X\) were being constructed. If such a path were to reach a vertex \(y\) in the deleted \(N(S)\), it could only extend to other vertices of \(S\) and \(N(S)\). It could not extend to an augmenting path. Therefore these vertices can be deleted. Upon completion, the algorithm will have produced a matching \(M\) for which there are no augmenting paths in the graph. By Theorem 7.2, \(M\) is a maximum matching.

\subsection*{7.2.1 Complexity}

Suppose that at the beginning of the for-loop, \(M\) has \(m\) edges. The largest possible size of \(S\) and \(N(S)\) is then \(m+1\), and \(m\), respectively. The number of edges \([S, N(S)]\) is at most \(m(m+1)\). In the worst possible case, \(S\) and \(N(S)\) will be built up to this size, and \(m(m+1)\) edges between them will be encountered. If an augmenting path is now found, then \(m\) will increase by one to give a worst case again for the next iteration. The length of the augmenting path will be at most \(2 m+1\), in case all \(m\) matching edges are in the path. The number of steps performed in this iteration of the for-loop will then be at most \(m(m+1)+(2 m+1)\). Since \(|X|+|Y|=n\), the number of vertices, one of \(|X|\) and \(|Y|\) is \(\leq n / 2\). We
page_147
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can take \(X\) as the smaller side. Summing over all iterations then gives
\[
\begin{gathered}
\sum_{m=0}^{\frac{n}{2}-1} m(m+1)+(2 m+1)=\sum 2\binom{m+1}{2}+(2 m+1) \\
=2\binom{n / 2+1}{3}+2\binom{n / 2}{2}+\frac{n}{2}
\end{gathered}
\]

The leading term in the expansion is \(n 3 / 24\), so that the algorithm is of order \(O(n 3)\), with a small constant coefficient. It can be improved with a more careful choice of augmenting paths. HOPCROFT and KARP [65] maintain several augmenting paths, and augment simultaneously on all of them to give \(O(n 2.5)\). This can also be accomplished with network flow techniques.

\section*{Exercises}
7.2.1 Program the Hungarian matching algorithm. The output should consist of a list of the edges in a maximum matching. If there is no matching saturating the set \(X\), this should be indicated by printing out the sets \(S \subseteq X\) found whose neighbor set \(N(S)\) is smaller than \(S\). Use the graphs in Figure 7.8 for input. The set \(X\) is marked by shaded dots, and \(Y\) by open dots.

\subsection*{7.3 Perfect matchings and 1-factorizations}

Given any graph \(G\) and positive integer \(k\), a \(k\)-factor of \(G\) is a spanning subgraph that is \(k\)-regular. Thus a perfect matching is a 1-factor. A 2-factor is a union of cycles that covers \(V(G)\), as illustrated in Figure 7.9. The reason for this terminology is as follows. Associate indeterminates \(x 1, x 2, \ldots, x n\) with the \(n\) vertices of a graph. An edge connecting vertex \(i\) to \(j\) can be represented by the expression \(x i-x j\). Then the entire graph can be represented (up to sign) by the product \(P(G)=\prod_{i j \in E(G)}\left(x_{i}-x_{j}\right)\). For example, if \(G\) is the 4-cycle, this product becomes \((x 1-x 2)(x 2-x 3)(x 3-x 4)(x 4-x 1)\). Since the number of terms in the product is \(\varepsilon(G)\), when it is multiplied out, there will be \(\varepsilon x^{\prime}\) s in each term. A 1 -factor of \(P(G)\), for example, \((x 1-x 2)(x 3-x 4)\), is a factor that contains each xi exactly once. This will always correspond to a perfect matching in \(G\), and so on. With some graphs it is possible to decompose the edge set into perfect matchings. For example, if \(G\) is the cube, we can write \(E(G)=M_{1} \cup M_{2} \cup M_{3}\), where
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\section*{FI GURE 7.8}

\section*{Sample graphs}
\(M 1=\{12,34,67,85\}, M 2=\{23,14,56,78\}\), and \(M 3=\{15,26,37,48\}\), as shown in Figure 7.10 . Each edge of \(G\) is in exactly one of \(M 1, M 2\), or \(M 3\).
In general, a \(k\)-factorization of a graph \(G\) is a decomposition of \(E(G)\) into \(H_{1} \cup H_{2} \cup \ldots \cup H_{m}\), where each Hi is a \(k\)-factor, and each Hi and Hj have no edges in common. The above decomposition of the cube is a 1factorization. Therefore we say the cube is 1-factorable.
LEMMA 7.6 Kn, nis 1-factorable.
PROOF Let \((X, Y)\) be the bipartition of \(K n, n\), where \(X=\{x 0, x 1, \ldots, x n-1\}\) and \(Y=\{y 0, y 1, \ldots, y n-1\}\). Define \(M 0=\{x i y i \mid i=0,1, \ldots, n-1\}, M 1=\{x i y i+1 \mid i=0,1, \ldots, n-1\}\), etc., where the addition is modulo \(n\). In general \(M k=\{x i y i+k \mid i=0,1, \ldots, n-1\}\). Clearly \(M j\) and \(M k\) have no edges in common, for any \(j\) and \(k\), and together \(M 0\), \(M 1, \ldots, M n-1\) contain all of \(E(G)\).
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FI GURE 7.9 2-factors of the cube


4
3
FI GURE 7.10

\section*{A 1-factorization of the cube}

Thus we have a 1-factorization of \(K n, n\)
LEMMA 7.7 K2nis 1-factorable.
PROOF Let \(V\left(K_{2 n}\right)=\{0,1,2, \ldots, 2 n-2\} \cup\{\infty\}\). Draw \(K 2 n\) with the vertices \(0,1, \ldots, 2 n-2\) in a circle, placing \(\infty\) in the center of the circle. This is illustrated for \(n=4\) in Fiqure 7.11. Take \(M_{0}=\{(0, \infty),(1,2 n-2),(2,2 n-3), \ldots,(n-1, n)\}=\{(0, \infty)\} \cup\{(i,-i) \mid i=1,2, \ldots, n-1\}\), where the addition is modulo \(2 n-1\). MO is illustrated by the thicker lines in Figure 7.11.
We can then "rotate" M0 by adding one to each vertex, \(M_{1}=M_{0}+1=\left\{(i+1, j+1) \mid(i, j) \in M_{0}\right\}\), where \(\infty+1=\infty\), and addition is modulo \(2 n-1\). It is easy to see from the diagram that \(M 0\) and \(M 1\) have no edges in common. Continuing like this, we have
\[
\begin{gathered}
M 0, M 1, M 2, \ldots, M 2 n-2, \\
\text { page_150 }
\end{gathered}
\]

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where \(M K=M 0+k\). They form a 1-factorization of \(K 2 n\).


FI GURE 7.11
1-factorizing \(K 2 n\), where \(n=4\)
We can use a similar technique to find a 2-factorization of \(K 2 n+1\).
LEMMA 7.8 K2n+1 is 2-factorable.
PROOF Let \(V\left(K_{2 n+1}\right)=\{0,1,2, \ldots, 2 n-1\} \cup\{\infty\}\). As in the previous lemma, draw the graph with the vertices in a circle, placing \(\infty\) in the center. The first 2 -factor is the cycle \(H 0=(0,1,-1,2,-2, \ldots, n-1, n+1, n\), \(\infty\) ), where the arithmetic is modulo \(2 n\). This is illustrated in Figure 7.12 , with \(n=3\). We then rotate the cycle to get \(H 1, H 2, \ldots, H n-1\), giving a 2 -factorization of \(K 2 n+1\).

\section*{Exercises}
7.3.1 Find all perfect matchings of the cube. Find all of its 1-factorizations.
7.3.2 Find all perfect matchings and 1-factorizations of \(K 4\) and \(K 6\).
7.3.3 Prove that the Petersen graph has no 1-factorization.
7.3.4 Prove that for \(k>0\) every \(k\)-regular bipartite graph is 1-factorable.
7.3.5 Describe another 1-factorization of \(K 2 n\), when \(n\) is even, using the fact that \(K n, n\) is a subgraph of \(K 2 n\). page_151

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3
FI GURE 7.12
2-factorizing \(K 2 n+1\), where \(n=3\)
7.3.6 Let \(M 1, M 2, . ., M k\) and \(M_{1}^{\prime}, M_{2}^{\prime}, \ldots, M_{k}^{\prime}\) be two 1-factorizations of a \(k\)-regular graph \(G\). The two factorizations are isomorphic if there is an automorphism \(\theta\) of \(G\) such that for each \(i, \theta\left(M_{i}\right)=M_{j}^{\prime}\), for some \(j\); that is, \(\theta\) induces a mapping of \(M 1, M 2, \ldots, M k\) onto \(M_{1}^{\prime}, M_{2}^{\prime}, \ldots, M_{k}^{\prime}\). How many non-isomorphic 1factorizations are there of \(K 4\) and \(K 6\) ?
7.3.7 How many non-isomorphic 1 -factorizations are there of the cube?

\subsection*{7.4 The subgraph problem}

Let \(G\) be a graph and let \(f: V(G) \rightarrow\{0,1,2, \ldots\}\) be a function assigning a non-negative integer to each vertex
of \(G\). An \(f\)-factor of \(G\) is a subgraph \(H\) of \(G\) such that \(\operatorname{deg}(u, H)=f(u)\), for each \(u \in V(G)\). So a 1-factor is an \(f\) factor in which each \(f(u)=1\).

Problem 7.1: Subgraph Problem
I nstance: a graph \(G\) and a function \(f: V(G) \rightarrow\{0,1,2, \ldots\}\).
Find: an \(f\)-factor in \(G\), if one exists.
There is an ingenious construction by TUTE [117], that transforms the subgraph problem into the problem of finding a perfect matching in a larger graph \(G^{\prime}\).
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Construct \(G^{\prime}\) as follows. For each edge \(e=u u\) of \(G, G^{\prime}\) has two vertices eu and \(e u\), such that \(e_{u} e_{v} \in E\left(G^{\prime}\right)\). For each vertex it of \(G\), let \(m(u)=\operatorname{deg}(u)-f(u)\). Corresponding to \(u \in V(G), G^{\prime}\) has \(m(u)\) vertices \(u 1, u 2, \ldots\), \(u m(u)\). For each edge \(e=u v \in E(G), u 1, u 2, \ldots, u m(u)\) are all adjacent to \(e_{u} \in V\left(G^{\prime}\right)\). This is illustrated in Fiqure 7.13, where \(\operatorname{deq}(u)=5\) and \(f(u)=3\).
\(G\)


\section*{FI GURE 7.13}

\section*{Tutte's transformation}

THEOREM 7.9 G has an \(f\)-factor if and only if \(G^{\prime}\) has a perfect matching.
PROOF Suppose that \(G\) has an \(f\)-factor \(H\). Form a perfect matching \(M\) in \(G^{\prime}\) as follows. For each edge \(u v \in H, e_{u} e_{v} \in M\). There are \(m(u)=\operatorname{deg}(u)-f(u)\) remaining edges at vertex \(u \in V(G)\). In \(G^{\prime}\), these can be matched to the vertices \(u 1, u 2, \ldots, u m(u)\) in any order.
Conversely, given a perfect matching \(M \subseteq G^{\prime}\), the vertices \(u 1, u 2, \ldots, u m(u)\) will be matched to \(m(u)\) vertices, leaving \(f(u)\) adjacent vertices of the form eu not matched to any ui. They can therefore only be matched to vertices of the form eu for some \(u\). Thus \(f(u)\) edges eueu are selected corresponding to each vertex \(u\). This gives an \(f\)-factor of \(G\).
So finding an \(f\)-factor in \(G\) is equivalent to finding a perfect matching in \(G^{\prime}\). If \(G\) has \(n\) vertices and \(\varepsilon\) edges, then \(G^{\prime}\) has
\[
4 \varepsilon-\sum f(u)
\]
vertices and
\[
\varepsilon+\sum\left(\operatorname{deg}^{2}(u)-\operatorname{deg}(u) f(u)\right)
\]
edges. Finding perfect matchings in non-bipartite graphs is considerably more complicated than in bipartite graphs, but is still very efficient. Edmonds' algorithm [38] will find a maximum matching in time \(O\) (n3). Thus, the subgraph
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problem can be solved using perfect matchings. However, it can be solved more efficiently by a direct algorithm than by constructing \(G^{\prime}\) and then finding a maximum matching.
7.5 Coverings in bipartite graphs

A covering or vertex cover of a graph \(G\) is a subset \(U \subseteq V(G)\) that covers every edge of \(G\); that is, every edge has at least one endpoint in \(U\).


\section*{FI GURE 7.14}

\section*{Coverings in a graph}

In general, we want the smallest covering possible. This is called a minimum covering. Figure 7.14 shows two coverings, indicated by shaded vertices. The covering with six vertices is minimal; namely, it has no subset that is a smaller covering. The other is a minimum covering; namely, \(G\) has no smaller covering. This is because any covering must use at least three vertices of the outer 5-cycle, and at least two vertices of the inner triangle, giving a minimum of five vertices.
In bipartite graphs, there is a very close relation between minimum coverings and maximum matchings. In general, let \(M\) be a matching in a graph \(G\), and let \(U\) be a covering. Then since \(U\) covers every edge of \(M,|U|\) \(\geq|M|\). This is true even if \(U\) is minimum or if \(M\) is maximum. Therefore, we conclude that if \(|U|=|M|\) for some \(M\) and \(U\), then \(U\) is minimum and \(M\) is maximum. In bipartite graphs, equality can always be achieved.
THEOREM 7.10 (König's theorem) If \(G\) is bipartite, then the number of edges in a maximum matching equals the number of vertices in a minimum covering.
PROOF Let \(M\) be a maximum matching, and let \((X, Y)\) be a bipartition of \(G\), where \(|X| \leq|Y|\). Let \(W \subseteq X\) be the set of all \(X\)-vertices not saturated by \(M\). If \(W=\emptyset\), then \(U=X\) is a covering with \(|U|=|M|\). Otherwise construct the
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set of all vertices reachable from \(W\) on alternating paths. Let \(S\) be the \(X\)-vertices reachable, and \(T\) the \(Y\) vertices reachable. Take \(U=T \cup(X-S)\). Then \(U\) is a covering with \(|U|=|M|\), as illustrated in Figure 7.15.


FI GURE 7.15

\section*{Minimum covering and maximum matching in a bipartite graph}

\subsection*{7.6 Tutte's theorem}

Tutte's theorem gives a necessary and sufficient condition for any graph to have a perfect matching. Let \(S \subseteq V(G)\). In general, \(G-S\) may have several connected components. Write odd( \(G-S\) ) for the number of components with an odd number of vertices. The following proof of Tutte's theorem is due to LOVÁSZ [84]. THEOREM 7.11 (Tutte's theorem) A graph \(G\) has a perfect matching if and only if odd(G-S) \(\leq|S|\), for every subset \(S \subseteq V(G)\).
\(\mathbf{P R O O F} \Rightarrow\) : Suppose that \(G\) has a perfect matching \(M\) and pick any \(S \subseteq V(G)\). Let \(G 1, G 2, \ldots, G m\) be the odd components of \(G-S\). Each \(G i\) contains at least one vertex matched by \(M\) to a vertex of \(S\). Therefore \(\operatorname{odd}(G-S)=m \leq|S|\). See Figure 7.16.
\(\Leftarrow\) : Suppose that \(\operatorname{odd}(G-S)=m \leq|S|\), for every \(S \subseteq V(G)\). Taking \(S=\emptyset\) gives odd \((G)=0\), so \(n=|G|\) is even. The proof is by reverse induction on \(\varepsilon(G)\), for any given \(n\). If \(G\) is the complete graph, it is clear that \(G\) has a perfect


FI GURE 7.16

\section*{Odd and even components of \(G-S\)}


FI GURE 7.17
\(H=M_{1} \oplus M_{2}\), case 1
matching, so the result holds when \(\varepsilon=\binom{n}{2}\). Let \(G\) be a graph with the largest \(\varepsilon\) such that \(G\) has no perfect matching. If \(u v \notin E(G)\), then because \(G+u u\) has more edges than \(G\), it must be that \(G+u u\) has a perfect matching. Let \(S\) be the set of all vertices of \(G\) of degree \(n-1\), and let \(G^{\prime}\) be any connected component of \(G-S\). If \(G^{\prime}\) is not a complete graph, then it contains three vertices \(x, y, z\) such that \(x \longrightarrow y \longrightarrow z\), but \(x \nmid z\). Since \(y \notin S\), \(\operatorname{deg}(y)<n-1\), so there is a vertex \(w \nrightarrow y\). Let \(M 1\) be a perfect matching of \(G+x z\) and let \(M 2\) be a perfect matching of \(G+y w\), as shown in Figures 7.17 and 7.18. Then \(x z \in M_{1}\) and \(y w \in M_{2}\). Let \(H=M_{1} \oplus M_{2} \cdot H\) consists of one or more alternating cycles in \(G\). Let \(C x z\) be the cycle of \(H\) containing \(x z\) and let Cyw be the cycle containing yw.
Case 1. \(C x z \neq C y w\).

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Form a new matching \(M\) by taking \(M 2\)-edges of \(C x z, M 1\)-edges of \(C y w\), and \(M 1\) edges elsewhere. Then \(M\) is a perfect matching of \(\mathcal{G}\), a contradiction.
Case 2. \(C x z=C y w=C\).
\(C\) can be traversed in two possible directions. Beginning with the vertices \(y, w\), we either come to \(x\) first or \(z\) first. Suppose it is \(z\). Form a new matching \(M\) by taking \(M 1\)-edges between \(w\) and \(z, M 2\)-edges between \(x\) and \(y\), and the edge \(y z\). Then take \(M 1\) edges elsewhere. Again \(M\) is a perfect matching of \(G\), a contradiction.


FI GURE 7.18
\(H=M_{1} \oplus M_{2}\), case 2
We conclude that every component \(G^{\prime}\) of \(G-S\) must be a complete graph. But then we can easily construct a
perfect matching of \(G\) as follows. Each even component of \(G-S\) is a complete graph, so it has a perfect matching. Every odd component is also a complete graph, so is has a near perfect matching, namely, one vertex is not matched. This vertex can be matched to a vertex of \(S\), since odd \((G-S) \leq|S|\). The remaining vertices of \(S\) form a complete subgraph, since they have degree \(n-1\), so they also have a perfect matching. It follows that every \(G\) satisfying the condition of the theorem has a perfect matching.
Tutte's theorem is a powerful criterion for the existence of a perfect matching. For example, the following graph has no perfect matching, since \(G-u\) has three odd components.
We can use Tutte's theorem to prove that every 3-regular graph \(G\) without cut-edges has a perfect matching. Let \(S \subseteq V(G)\) be any subset of the vertices. Let \(G 1, G 2, \ldots, G k\) be the odd components of \(G-S\). Let mi be the number of edges connecting \(G i\) to \(S\). Then mi>1, since \(G\) has no cut-edge. Since \(\sum_{v \in G_{i}} \operatorname{DEG}(v)=2 \varepsilon\left(G_{i}\right)+m_{i}=3\left|G_{i}\right|_{\text {_an }}\) odd number, we conclude that \(m i\) is odd. Therefore mi \(\geq 3\), for each \(i\). But \(\sum_{v \in S} \operatorname{DEG}(v)=3|S| \geq \sum_{i} m_{i}\), since all of the mi edges have one endpoint in \(S\). It follows that \(3|S| \geq 3 k\), or \(|S| \geq \operatorname{odd}(G-S)\), for all \(S \subseteq V(G)\). Therefore \(G\) has a perfect matching \(M\). \(G\) also has a 2-factor, since \(G-M\) has degree two.

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\section*{FI GURE 7.19}

\section*{A 3-regular graph with no perfect matching \\ Exercises}
7.6.1 For each integer \(k>1\), find a \(k\)-regular graph with no perfect matching.
7.6.2 A near perfect matching in a graph \(G\) is a matching which saturates all vertices of \(G\) but one. A near 1factorization is a decomposition of \(E(G)\) into near perfect matchings. Prove that \(K 2 n+1\) has a near 1factorization.
7.6.3 Find a condition similar to Tutte's theorem for a graph to have a near perfect matching.

\subsection*{7.7 Notes}

Algorithms for maximum matchings in non-bipartite graphs are based on blossoms, discovered by EDMONDS [38]. A blossom is a generalization of an odd cycle in which all but one vertex is matched. An excellent description of Edmonds' algorithm appears in PAPADI MITRIOU and STEIGLITZ [94]. A good source book for the theory of matchings in graphs is LOVÁSZ and PLUMMER [85]. Exercise 7.1.2 is from BONDY and MURTY [19].
The proof of Tutte's theorem presented here is based on a proof by LOVÁSZ [84]. Tutte's transformation to reduce the subgraph problem to a perfect matching problem is from TUTTE [117]. His Factor theorem, TUTTE [118],

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is a solution to the subgraph problem. It is one of the great theorems of graph theory. The theory of 1factorizations has important applications to the theory of combinatorial designs.

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\section*{Network Flows}

\subsection*{8.1 Introduction}

A network is a directed graph used to model the distribution of goods, data, or commodities, etc., from their centers of production to their destinations. For example, Figure 8.1 shows a network in which goods are produced at node \(s\), and shipped to node \(t\). Each directed edge has a limited capacity, being the maximum number of goods that can be shipped through that channel per time period (e.g., 3 kilobytes per second or 3 truckloads per day). The diagram indicates the capacity as a positive integer associated with each edge. The actual number of goods shipped on each edge is shown in square brackets beside the capacity. This is called the flow on that edge. It is a non-negative integer less that or equal to the capacity. Goods cannot accumulate at any node; therefore the total in-flow at each node must equal the out-flow at that node. The problem is to find the distribution of goods that maximizes the net flow from \(s\) to \(t\).
This can be modeled mathematically as follows. When the edges of a graph have a direction, the graph is called a directed graph or digraph. A network \(N\) is a directed graph with two special nodes \(s\) and \(t\); \(s\) is called the source and \(t\) is called the target. All other vertices are called intermediate vertices. The edges of a directed graph are ordered pairs \((u, u)\) of vertices, which we denote by \(\overrightarrow{u v}\). We shall find it convenient to say that \(u\) is adjacent to \(u\) even when we do not know the direction of the edge. So the phrase \(u\) is adjacent to \(u\) means either \(\overrightarrow{u v}\) or \(\overrightarrow{v u}\) is an edge. Each edge \(\overrightarrow{u v} \in E(N)\) has a capacity ( \(\overrightarrow{u v}\) ), being a positive integer, and a flow \(f(\overrightarrow{u v})\), a non-negative integer, such that \(f(\overrightarrow{u v}) \leq \operatorname{CAP}(\overrightarrow{u v})\). If \(u\) is any vertex of \(N\), the out-flow at \(u\) is
\[
f^{+}(v)=\sum_{u, v \longrightarrow u} f(\overrightarrow{v u})
\]
where the sum is over all vertices \(u\) to which \(u\) is joined. The in-flow is the sum

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\(s\)


FI GURE 8.1 A network
over all incoming edges at \(u\)
\[
f^{-}(v)=\sum_{u, u \Longrightarrow v} f(\overrightarrow{u v})
\]

A valid flow \(f\) must satisfy two conditions.
1. Capacity constraint: \(0 \leq f(\overrightarrow{u v}) \leq \operatorname{CAP}(\overrightarrow{u v})\), for all \(\overrightarrow{u v} \in E(N)\).
2. Conservation condition: \(f+(u)=f-(u)\), for all \(u \neq s, t\).

Notice that in Figure 8.1 both these conditions are satisfied. The value of the flow is the net out-flow at \(s\); in this case, \(\operatorname{VAL}(f)=20\).
In general, there may be in-edges as well as out-edges at \(s\). The net flow from \(s\) to \(t\) will then be the out-flow at the source minus the in-flow. This is called the value of the flow, \(\operatorname{VAL}(f)=f+(s)-f-(s)\). The max-flow problem is:

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\section*{Problem 8.1: Max-Flow}

\section*{Instance: a network \(N\).}

Find: \(\quad\) a flow \(f\) for \(N\) of maximum value.
Any flow \(f\) that has maximum value for the network \(N\) is called a max-flow of \(N\). This problem was first formulated and solved by Ford and Fulkerson. In this chapter we shall present the Ford-Fulkerson algorithm, and study several applications of the max-flow-min-cut theorem.
It is possible that a network encountered in practice will have more than one source or target. If \(s 1, s 2, \ldots, s k\) are all sources in a network \(N\), and \(t 1, t 2, \ldots, t m\) are all targets, we can replace \(N\) with a network \(N^{\prime}\) with only one source and one target as follows. Add a vertex \(s\) to \(N\), and join it to \(s 1, s 2, \ldots, s k\). Add a vertex \(t\) and join \(t 1, t 2, \ldots, t m\) to \(t\). Assign a capacity \(\operatorname{CAP}\left(\overrightarrow{s_{i}}\right)\) being the sum of the capacities of the out-edges at si, and a capacity \(\operatorname{CAP}\left(\overrightarrow{t_{i} t}\right)\), being the sum of the capacities of all incoming edges to \(t i\). Call the resulting network \(N^{\prime}\). For every flow in \(N\) there is a corresponding flow in \(N^{\prime}\) with equal value, and vice-versa. Henceforth we shall always assume that all networks have just one source and target. The model we are using assumes that edges are one-way channels and that goods can only be shipped in the direction of the arrow. If a two-way channel from \(u\) to \(u\) is desired, this can easily be accommodated by two directed edges \(\overrightarrow{u v}\) and \(\overrightarrow{v u}\). Let \(S \subseteq V(N)\) be a subset of the vertices such that \(s \in S, t \notin S\). Write \(\bar{S}=V(N)-S\). Then \([S, \bar{S}]\) denotes the set of all edges of \(N\) directed from \(S\) to \(\bar{S}\). See Figure 8.2. Consider the sum
\[
\begin{equation*}
\sum_{v \in S}\left(f^{+}(v)-f^{-}(v)\right) . \tag{8.1}
\end{equation*}
\]

Since \(f+(u)=f-(u)\), if \(u \neq s\), this sum equals VAL(f). On the other hand, \(f+(u)\) is the total out-flow at \(v \in S\). Consider an out-edge \(\overrightarrow{v u}\) at \(u\). Its flow \(f(\overrightarrow{v u})\) contributes to \(f+(u)\). It also contributes to \(f-(u) . u \in S\), then \(f(\overrightarrow{v u})\) will appear twice in the sum 8.1, once for \(f+(U)\) and once for \(f-(u)\), and will therefore cancel. See Figure 8.2, where \(S\) is the set of shaded vertices. If \(u \notin S\), then \(f(\overrightarrow{v u})\) will appear in the summation as part of \(f+(u)\), but will not be canceled by \(f-(u)\). A similar argument holds if \(v \in \bar{S}\) and \(u \in S\). Therefore
\[
\begin{aligned}
\operatorname{VAL}(f) & =\sum_{v \in S}\left(f^{+}(v)-f^{-}(v)\right) \\
& =\sum_{\substack{v u}[S, \bar{S}]}^{\substack{\text { page_ }\\
}} f(\overrightarrow{v u})-\sum_{\overrightarrow{v u} \in[\bar{S}, S]} f(\overrightarrow{v u})
\end{aligned}
\]

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This says that the value of the flow can be measured across any edge-cut \([S, \bar{S}]\), such that \(s \in S\) and \(t \in \bar{S}\). If we write
\[
f^{+}(S)=\sum_{v \in S} f^{+}(v)
\]
and
\[
f^{-}(S)=\sum_{v \in S} f^{-}(v)
\]
then
If we write
\[
\operatorname{VAL}(f)=f+(S)-f-(S)
\]
\[
f([S, \bar{S}])=\sum_{\bar{v} \vec{u} \in[S, \bar{S}]} f(\overrightarrow{v u})
\]
and
\[
f([\bar{S}, S])=\sum_{\overrightarrow{v u} \in[\bar{S}, S]} f(\overrightarrow{v u})
\]
then we can also express this as
\[
\operatorname{VAL}(f)=f([S, \bar{S}])-f([\bar{S}, S])
\]

Let \(K=[S, \bar{S}]\) be any edge-cut with \(s \in S\) and \(t \in \bar{S}\). The capacity of \(K\) is
\[
\operatorname{CAP}(K)=\sum_{\overrightarrow{u v} \in K} \operatorname{CAP}(\overrightarrow{u v})
\]

This is the sum of the capacities of all edges out of \(S\). The value of any flow in \(N\) is limited by the capacity of any edge-cut \(K\). An edge-cut \(K\) is a min-cut if it has the minimum possible capacity of all edge-cuts in \(N\).
LEMMA 8.1 Let \(K=[S, \bar{S}]_{\text {be }}\) an edge-cut in a network \(N\) with \(s \in S, t \in \bar{S}\) and flow \(f\). Then \(\operatorname{VAL}(f) \leq \operatorname{CAP}(K)\). If \(\operatorname{VAL}(f)=\operatorname{CAP}(K)\), then \(f\) is a max-flow and \(K\) is a min-cut.
PROOF Clearly the maximum possible flow out of \(S\) is bounded by \(\operatorname{CAP}(K)\); that is, \(f+(S) \leq \operatorname{CAP}(K)\). This holds even if \(K\) is a min-cut or \(f\) a max-flow. The flow into \(S\) is non-negative; that is, \(f-(S) \geq 0\). Therefore \(\operatorname{VAL}(f)=f+(S)-f-(S) \leq \operatorname{CAP}(K)\). If \(\operatorname{VAL}(f)=\operatorname{CAP}(K)\), then it must be that \(f\) is maximum, for the value of no flow can exceed the capacity of any cut. Similarly \(K\) must be a min-cut. Note that in this situation \(f+(S)=\operatorname{CAP}(K)\) and \(f-(S)=0\). That is, every edge \(\overrightarrow{u v}\) directed out of \(S\) satisfies \(f(\overrightarrow{u v})=\operatorname{CAP}(\overrightarrow{u v})\). Every edge \(\overrightarrow{u v}\) into \(S\) carries no flow, \(f(\overrightarrow{u v})=0\).
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\section*{FI GURE 8.2}

\section*{A set \(S\) where}

In the next section we shall prove the max-flow-min-cut theorem. This states that the value of a max-flow and the capacity of a min-cut are always equal, for any network \(N\).

\subsection*{8.2 The Ford-Fulkerson algorithm}

If we assign \(f(\overrightarrow{u v})=0\), for all \(\overrightarrow{u v} \in E(N)\), this defines a valid flow in \(N\), the zero flow. The Ford-Fulkerson algorithm begins with the zero flow, and increments it through a number of iterations until a max-flow is obtained. The method uses augmenting paths. Consider the st-path \(P=s u 1 u 5 u 2 u 6 t\) in Figure 8.1. (We ignore the direction of the edges when considering these paths.) Each edge of \(P\) carries a certain amount of flow. The traversal of \(P\) from s to \(t\) associates a direction with \(P\). We can then distinguish two kinds of edges of \(P\), forward edges, those like sul whose direction is the same as that of \(P\), and backward edges, those like u5u2 whose direction is opposite to that of \(P\). Consider a forward

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edge \(\overrightarrow{u v}\) in an st-path \(P\). If \(f(\overrightarrow{u v})<\operatorname{CAP}(\overrightarrow{u v})\), then \(\overrightarrow{u v}\) can carry more flow. Define the residual capacity of \(u v \in E(P)\) to be
\[
\operatorname{RESCAP}(u v)= \begin{cases}\operatorname{CAP}(\overrightarrow{u v})-f(\overrightarrow{u v}), & \text { if } u v \text { is a forward edge } \\ f(\overrightarrow{v u}), & \text { if } u v \text { is a backward edge }\end{cases}
\]

The residual capacity of a forward edge \(u v \in E(P)\) is the maximum amount by which the flow on \(\overrightarrow{u v}\) can be increased. The residual capacity of a backward edge \(u v \in E(P)\) is the maximum amount by which the flow on \(v u\) can be decreased. For example, in the network of Figure 8.1, we increase the flow on all forward edges of \(P\) by 2 , and decrease the flow on all backward edges of \(P\) also by 2 . The result is shown in Figure 8.2. We have a new flow, with a larger value than in Figure 8.1.

In general, let \(P\) be an st-path in a network \(N\) with flow \(f\). Define the residual capacity of \(P\) to be
\[
\delta(P)=\operatorname{Min}\{\operatorname{RESCAP}(\overrightarrow{u v}): u v \in E(P)\}
\]

Define a new flow \(f^{*}\) in \(N\) as follows:
\[
f^{*}(\overrightarrow{u v})= \begin{cases}f(\overrightarrow{u v}), & \text { if } u v \text { is not an edge of } P \\ f(\overrightarrow{u v})+\delta(P), & \text { if } u v \text { is a forward edge of } P \text { and } \\ f(\overrightarrow{u v})-\delta(P), & \text { if } u v \text { is a backward edge of } P\end{cases}
\]

LEMMA \(8.2 f^{*}\) is a valid flow in \(N\) and VAL \(\left(f^{*}\right)=\operatorname{VAL}(f)+\delta(P)\).
PROOF We must check that the capacity constraint and conservation conditions are both satisfied by \(f^{*}\). It is clear that the capacity constraint is satisfied, because of the definition of the residual capacity of \(P\) as the minimum residual capacity of all edges in \(P\). To verify the conservation condition, consider any intermediate vertex \(u\) of \(P\). Let its adjacent vertices on \(P\) be \(u\) and \(w\), so that \(u v\) and \(u w\) are consecutive edges of \(P\). There are four cases, shown in Figure 8.3.
Case 1. \(u u\) and \(u w\) are both forward edges of \(P\).
Because \(f(\overrightarrow{u v})\) and \(f(\overrightarrow{v w})\) both increase by \(\delta(P)\) in \(f^{*}\), it follows that \(f+(U)\) and \(f-(u)\) both increase \(\delta(P)\). The net result on \(f+(u)-f-(u)\) is zero.
Case 2. \(u v\) is a forward edge and \(u w\) is a backward edge.
In this case \(f(\overrightarrow{u v})\) increases and \(f(\overrightarrow{w v})\) decreases by \(\delta(P)\) in \(f^{*}\). It follows that \(f+(U)\) and \(f-(U)\) are both unchanged.
Case 3. \(u v\) is a backward edge and \(u w\) is a forward edge.
In this case \(f(\overrightarrow{v u})\) decreases and \(f(\overrightarrow{v w})\) increases by \(\delta(P)\) in \(f^{*}\). It follows that \(f+(U)\) and \(f-(U)\) are both unchanged.
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Case 4. \(u u\) and \(u w\) are both backward edges of \(P\).
Because \(f(\overrightarrow{v u})\) and \(f(\overrightarrow{w v})\) both decrease by \(\delta(P)\) in \(f^{*}\), it follows that \(f+(U)\) and \(f-(U)\) both decrease by \(\delta(P)\). The net result on \(f+(u)-f-(u)\) is zero.
The value of \(f^{*}\) is \(f^{*}+(s)-f^{*}-(s)\). If the first edge of \(P\) is \(s u\), a forward edge, then it is clear that the value
increases by \(\delta(P)\), since \(f(\overrightarrow{s u})\) increases. If su is a backward edge, then \(f(\overrightarrow{s u})\) decreases, so that \(f\)-(s) also decreases, thereby increasing the value of the flow. Therefore \(\operatorname{VAL}(f *)=\operatorname{VAL}(f)+\delta(P)\).


Case 1


Case 3


Case 2


Case 4

\section*{FI GURE 8.3}

The four cases for edges \(u U\) and \(U W\) on path \(P\)
DEFINITION 8.1: An st-path \(P\) for which \(\delta(P)>0\) is called an augmenting path.
This method of altering the flow on the edges of \(P\) is called augmenting the flow. If \(\delta(P)>0\) it always results in a flow of larger value. We give an outline of the Ford-Fulkerson algorithm in Algorithm 8.2.1.

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Algorithm 8.2.1: \(\mathrm{FF}(N, \mathrm{~s}, t)\)
comment: \(\left\{\begin{array}{l}N \text { is a network with source } s \text { and target } t . \\ f \text { is the flow. } \\ P \text { is a path. }\end{array}\right.\)
\(f \leftarrow\) the zero flow
search for an augmenting path \(P\)
while a path \(P\) was found
do \(\left\{\begin{array}{l}\text { augment the flow on } P \\ \operatorname{VAL}(f) \leftarrow \operatorname{VAL}(f)+\delta(P) \\ \text { search for an augmenting path } P\end{array}\right.\)
comment: the flow is now maximum
The algorithm stops when \(N\) does not contain an augmenting path. We show that in this situation the flow must be maximum. The outline given in Algorithm 8.2.1 does not specify how the augmenting paths are to be found. Among the possibilities are the breadth-first and depth-first searches. We shall see later that the breadth-first search is the better choice. As the algorithm searches for an augmenting path, it will construct paths from \(s\) to various intermediate vertices \(u\). The paths must have positive residual capacity. An su-path with positive residual capacity is said to be unsaturated. A vertex \(u\) is \(s\)-reachable if \(N\) contains an unsaturated su-path. This means that \(u\) can be reached from \(s\) on an unsaturated path.
THEOREM 8.3 Let \(N\) be a network with a flow \(f\). Then \(f\) is maximum if and only if \(N\) contains no augmenting path.
PROOF Suppose that \(f\) is a max-flow. There can be no augmenting path in \(N\), for this would imply a flow of larger value. Conversely, suppose that \(f\) is a flow for which there is no augmenting path. We show that \(f\) is maximum. Let \(S\) denote the set of all \(s\)-reachable vertices of \(N\). Clearly \(s \in S\). Since there is no augmenting path, the target is not s-reachable. Therefore \(t \in \bar{S}\). Consider the edge-cut \(K=[S, \bar{S}]\). If \(\overrightarrow{u v} \in K\) is an edge out of \(S\), then \(\operatorname{RESCAP}(\overrightarrow{u v})=0\); for otherwise \(u\) would be \(s\)-reachable on the forward edge \(\overrightarrow{u v}\) from \(u \in S\). Therefore \(f(\overrightarrow{u v})=\operatorname{CAP}(\overrightarrow{u v})\) for all \(\overrightarrow{u v} \in K\) that are out edges of \(S\). Thus \(f+(S)=\operatorname{CAP}(K)\). If \(\overrightarrow{u v} \in[\bar{S}, S]\) is any edge into \(S\), then \(f(\overrightarrow{u v})=0\); for otherwise \(u\) would be \(s\)-reachable on the backward edge \(\overrightarrow{v u}\) from \(v \in S\). Consequently \(f-(S)=0\). It follows that \(\operatorname{VAL}(f)=\operatorname{CAP}(K)\), so that \(f\) is a max-flow and \(K\) a min-cut, by

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This is illustrated in Figure 8.2, in which all edges out of \(S\) are saturated. In this example there are no edges into S. If there were, they would carry no flow. So the flow in Figure 8.2 is maximum. Notice that a consequence of this theorem is that when \(f\) is a max-flow, the set \(S\) of \(s\)-reachable vertices defines a min-cut \(K=[S, \bar{S}]\). This is summarized as follows:
THEOREM 8.4 (Max-flow-min-cut theorem) In any network the value of a max-flow equals the capacity of a min-cut.
We are now ready to present the Ford-Fulkerson algorithm as a breadth-first search for an augmenting path. The vertices will be stored on a queue, the ScanQ, an array of s-reachable vertices. QSize is the current number of vertices on the ScanQ. The unsaturated su-paths will be stored by an array PrevPt[•], where \(\operatorname{PrevPt}[u]\) is the point previous to \(u\) on an su-path Pu. The residual capacity of the paths will be stored by an array ResCap[•], where ResCap[u] is the residual capacity \(\delta(P u)\) of \(P u\) from \(s\) up to \(u\). The algorithm is presented as a single procedure, but could be divided into smaller procedures for modularity and readability. page_169

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Algorithm 8.2.2: MAXFLOW( \(N, s, t\) )
\(f \leftarrow\) the zero flow
for all vertices \(u\) do \(\operatorname{PrevPt}[u] \leftarrow 0\)
while true "search for an augmenting path"
    for all \(v\) adjacent to \(u\)
    do if \(v \notin\) Scan \(Q\)



until \(k>\) QSize
comment: \(\left\{\begin{array}{l}\text { Flow is now maximum. } \\ \text { ScanQ contains the } s \text {-reachable vertices. }\end{array}\right.\)
output (ScanQ, and the flow on each edge)
exit
1 : comment: augmenting path found, re-initialize \(S c a n Q\)
AUGMENTFLow \((t)\)
for \(k \leftarrow 1\) to QSize do \(\operatorname{PrevPt}[\operatorname{Scan} Q[k]] \leftarrow 0\)

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The procedure which augments the flow starts at \(t\) and follows PrevPt[u] up to \(s\). Given an edge \(\overrightarrow{u v}\) on the augmenting path, where \(u=\operatorname{PrevPt}[u]\), a means is needed of determining whether \(\overrightarrow{u v}\) is a forward or backward edge. One way is to store PrevPt[u]=u for forward edges and PrevPt[u]=-u for backward edges. This is not indicated in Algorithm 8.2.2, but can easily be implemented.

Algorithm 8.2.3: AUGMENTFLOW ( \(t\) )
\(u \leftarrow t\)
\(u_{\leftarrow}\) PrevPt[u]
\(\delta \leftarrow \operatorname{ResCap}[t]\)
while \(u \neq 0\)
do \(\left\{\begin{array}{l}\text { if } u \longrightarrow v \\ \text { then } f(\overrightarrow{u v}) \leftarrow f(\overrightarrow{u v})+\delta \quad \text { "a forward edge" } \\ \text { else } f(\overrightarrow{u v}) \leftarrow f(\overrightarrow{u v})-\delta \quad \text { "a backward edge" } \\ v \leftarrow u \\ u \leftarrow \operatorname{PrevPt}[v] \\ \operatorname{VAL}(f) \leftarrow \operatorname{VAL}(f)+\delta\end{array}\right.\)

In programming the max-flow algorithm, the network \(N\) should be stored in adjacency lists. This allows the loop
for all \(u\) adjacent to \(u\) do
to be programmed efficiently. The out-edges and in-edges at \(u\) should all be stored in the same list. We need to be able to distinguish whether \(u \longrightarrow v\) or \(v \longrightarrow u\). This can be flagged in the record representing edge \(\overrightarrow{u v}\). If \(\overrightarrow{u v}\) appears as an out-edge in the list for node \(u\), it will appear as an in-edge in the list for vertex \(u\). When the flow on edge \(\overrightarrow{u v}\) is augmented, it must be augmented from both endpoints. One way to augment from both endpoints simultaneously is to store not the flow \(f(\overrightarrow{u v})\) itself, but a pointer to it. Then it is not necessary to find the other endpoint of the edge. Thus a node \(x\) in the adjacency list for vertex \(u\) contains following four fields:
- \(\operatorname{AdjPt}\langle x\rangle\), a vertex \(u\) that is adjacent to or from \(u\).
. OutEdge \(\langle x\rangle\), a boolean variable set to true if \(u \longrightarrow v\), and false if \(v \longrightarrow u\).
- Flow \(\langle x\rangle\), a pointer to the flow on \(\overrightarrow{u v}\).
- Next \(\langle x\rangle\), the next node in the adjacency list for \(u\).

This breadth-first search version of the Ford-Fulkerson algorithm is sometimes referred to as the "labeling" algorithm in some books. The values ResCap[v] and PrevPt[ \(u\) ] are considered the labels of vertex \(u\). page_171

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The algorithm works by constructing all shortest unsaturated paths from s. If an augmenting path exists, it is sure to be found. This can easily be proved by induction on the length of the shortest augmenting path. The flow is then augmented and the algorithm exits from the inner repeat loop by branching to statement 1. If no augmenting path exists, then the inner repeat loop will terminate. The vertices on the ScanQ will contain the set \(S\) of all \(s\)-reachable vertices, such that \([S, \bar{S}]\) is a min-cut.
It is difficult to form an accurate estimate of the complexity of the BF-FF algorithm. We shall prove that it is polynomial. This depends on the fact that only shortest augmenting paths are used. If non-shortest paths are used, the FF algorithm is not always polynomial. Consider the network of Figure 8.4. We augment first on path \(P=(s, a, b, t)\), which has residual capacity one. We then augment on path \(Q=(s, b, a, t)\), also of residual capacity one, since ba is a backward edge, and \(f(\overrightarrow{a b})=1\). Augmenting on \(Q\) makes \(\delta(P)=1\), so we again augment on \(P\), and then augment again on \(Q\), etc. After 2000 iterations a max-flow is achieved-the number of iterations can depend on the value of the max-flow. This is not polynomial in the parameters of the network. However, if shortest augmenting paths are used, this problem does not occur.
\(S\)

b

\section*{FI GURE 8.4}

\section*{A max-flow in \(\mathbf{2 0 0 0}\) iterations}

Consider an augmenting path \(P\) in a network \(N . \delta(P)\) is the minimum residual capacity of all edges of \(P\). Any edge \(\overrightarrow{u v} \in P\) such that \(\operatorname{RESCAP}(\overrightarrow{u v})=\delta(P)\) is called a bottleneck. Every augmenting path has at least one bottleneck, and may have several. Suppose that a max-flow in \(N\) is reached in \(m\) iterations, and let \(P j\) be the
augmenting path on iteration \(j\). Let \(d j(s, u)\) denote the length of a shortest unsaturated su-path in iteration \(j\), for all vertices \(u\).
LEMMA \(8.5 d j+1(s, u) \geq d j(s, u)\), for all \(u \in V(N)\).
PROOF Let \(Q j=Q j(s, u)\) be a shortest unsaturated su-path at the beginning of page_172

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iteration \(j\), and let \(Q j+1=Q j+1(s, u)\) be a shortest unsaturated su-path at the beginning of iteration \(j+1\). Then \(\ell(Q j)=d j(s, u)\) and \(\ell(Q j+1)=d j+1(s, u)\). If \(\ell(Q j) \leq \ell(Q j+1)\), the lemma holds for vertex \(u\), so suppose that \(\ell(Q j)>\ell(Q j+1)\), for some \(u\). Now \(Q j+1\) is not unsaturated at the beginning of iteration \(j\), so it must become unsaturated during iteration \(j\). Therefore \(P j\) and \(Q j+1\) have at least one edge in common that becomes unsaturated during iteration \(j\). The proof is by induction on the number of such edges. Suppose first that \(x y\) is the only such edge in common. Since \(x y\) becomes unsaturated during iteration \(j\), it has opposite direction on \(P j\) and \(Q j+1\). See Figure 8.5. Without loss of generality, \(Q j+1[s, x]\) and \(Q j+1[y, u]\) are unsaturated on iteration \(j\). Since \(P j\) is a shortest path, \(\ell(P j[s, x]) \leq \ell(Q j+1[s, x])\). But then \(P j[s, y] Q j+1[y, u]\) is an unsaturated su-path on iteration \(j\), and has length less than \(Q j+1(s, u)\), which in turn has length less than \(Q j(s, u)\), a contradiction. If \(P j[s, y]\) intersects \(Q j+1[y, u]\) at a vertex \(z\), then \(P j[s, z] Q j+1[z, u]\) is an even shorter unsaturated path.


\section*{FI GURE 8.5}

\section*{Paths Pj and \(Q j+1\)}

Suppose now that \(P j\) and \(Q j+1\) have more than one edge in common that becomes unsaturated during iteration \(j\). Let \(x y\) be the first such edge on \(Q j+1\) traveling from \(s\) to \(u\). Let \(z\) be the point on \(Q j+1\) nearest to u that \(P j[s, x]\) contacts before reaching \(x\) (maybe \(z=y\) ). Then \(Q j+1[s, x]\) and \(P j[s, z]\) are unsaturated at the beginning of iteration \(j\). Since \(P j\) is a shortest unsaturated path, \(\ell(P j[s, z])<\ell(Q j+1[s, x])<\ell(Q j+1[s, z])\). Now either \(Q j+1[z, u]\) is unsaturated on iteration \(j\), or else it has another edge in common with \(P j\). If it is unsaturated, then \(P j[s, z] Q j+1[z, u]\) is an unsaturated su-path that contradicts the assumption that \(d j(s\), \(u)>d j+1(s, u)\). If there is another edge in common with \(P j\), then we can repeat this argument. Let \(x^{\prime} y^{\prime}\) be the first edge of \(Q j+1[z, u]\) in
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common with \(P j\). Let \(z^{\prime}\) be the point on \(Q j+1[z, u]\) nearest to \(u\) that \(P j\left[s, x^{\prime}\right]\) contacts before reaching \(x^{\prime}\), etc. Proceeding in this way we eventually obtain an su-path that is shorter than \(Q j\) and unsaturated on iteration \(j\), a contradiction.
It follows that \(d j+1(s, t) \geq d j(s, t)\), for every iteration \(j\). By constructing unsaturated paths from \(t\) in a backward direction we can similarly prove that \(d j+1(u, t) \geq d j(u, t)\), for all vertices \(u\). If we can now prove that \(d j+1(s, t)>d j(s, t)\), then we can bound the number of iterations, since the maximum possible distance from \(s\) to \(t\) is \(n-1\), where \(n\) is the number of vertices of \(N\).
THEOREM 8.6 The breadth-first Ford-Fulkerson algorithm requires at most \(\frac{1}{2} n \varepsilon+1\) iterations.
PROOF On each iteration some edge is a bottleneck. After \(\varepsilon+1\) iterations, some edge has been a bottleneck twice, since there are only \(\varepsilon\) edges. Consider an edge \(\overrightarrow{u v}\) which is a bottleneck on iteration \(i\) and then later a bottleneck on iteration j. Refer to Figure 8.6.


FI GURE 8.6
Paths Piand \(P j\)
Then \(d i(s, t)=d i(s, u)+d i(u, t)+1\) and \(d j(s, t)=d j(s, u)+d j(u, t)+1\). But \(d i(s, u) \leq d j(s, u)=d j(s, u)+1\) and \(d i(u, t) \leq d j(u, t)=d j(u, t)+1\). Therefore \(d i(s, u)+d i(u, t) \leq d j(s, u)+d j(u, t)+2\). It follows that \(d i(s, t) \leq d j(s\), \(t)+2\). Each time an edge is repeated as a bottleneck, the distance from s to \(t\) increases by at least two. Originally \(d 1(s, t) \geq 1\). After \(\varepsilon+1\) iterations, some edge has been a bottleneck twice. Therefore \(d \varepsilon+1(s, t) \geq 3\). Similarly \(d 2 \varepsilon+1(s, t) \geq 5\), and so on. In general \(d k \varepsilon+1(s, t) \geq 2 k+1\). Since the maximum distance from \(s\) to \(t\) is \(n-1\), we have \(2 k+1 \leq n-1\), so that \(k \leq n / 2\). The maximum number of iterations is then \(k \varepsilon+1 \leq \frac{1}{2} n \varepsilon+1\).
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Each iteration of the BF-FF algorithm is a breadth-first search for an augmenting path. A breadth-first search takes at most \(O(\varepsilon)\) steps. Since the number of it-erations is at most \(\frac{1}{2} n \varepsilon+1\), this gives a complexity of \(O\) ( \(n \varepsilon 2\) ) for the breadth-first Ford-Fulkerson algorithm. This was first proved by EDMONDS and KARP [39].

\section*{Exercises}
8.2.1 Find a max-flow in the network shown in Figure 8.7. Prove your flow is maximum by illustrating a mincut \(K\) such that VAL \((f)=\operatorname{CAP}(K)\).


\section*{FI GURE 8.7}

\section*{A network}
8.2.2 Show that if there is no directed st-path in a network \(N\), then the maximum flow in \(N\) has value zero. Can there be a flow whose value is negative? Explain.
8.2.3 Explain why \(\sum_{v \in S} f^{+}(v)\) and \(\sum_{\overrightarrow{v u} \in[S, \bar{S}]} f(\overrightarrow{v u})\) are in general, not equal.
8.2.4 Consider the network \(N\) of Figure 8.7 with flow \(f\) defined as follows: \(f(s u 1)=6, f(s u 2)=0, f(s u 3)=2\), \(f(u 1 u 4)=2, f(u 1 u 5)=4, f(u 2 u 4)=0, f(u 2 u 6)=0, f(u 3 u 5)=2, f(u 3 u 6)=0, f(u 5 u 2)=0, f(u 4 t)=2, f(u 5 t)-6\), \(f(u 6 t)=0\). A breadth-first search of \(N\) will construct the subnetwork of all shortest, unsaturated paths in \(N\). This subnetwork is called the auxiliary network, Aux ( \(N, f\) ). A forward edge \(\overrightarrow{u v}\) of \(N\) is replaced by a forward edge \(\overrightarrow{u v}\) with capacity \(\operatorname{CAP}(\overrightarrow{u v})-f(\overrightarrow{u v})\) in the auxiliary network. A backward edge \(\overrightarrow{v u}\) of \(N\) is replaced by a forward edge \(\overrightarrow{u v}\) with capacity \(f(\overrightarrow{u v})\) in \(\operatorname{Aux}(N, f)\). Initially the flow in \(\operatorname{AUX}(N, f)\) is the zero flow. Construct the auxiliary network for the graph shown. Find a max-flow in

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Aux ( \(N, f\) ), and modify \(f\) in \(N\) accordingly. Finally, construct the new auxiliary network for \(N\).
8.2.5 Program the breadth-first Ford-Fulkerson algorithm. Test it on the networks of this chapter.
8.2.6 If \([S, \bar{S}]\) and \([T, \bar{T}]\) are min-cuts in a network \(N\), show that \([S \cup T, \overline{S \cup T}]\) and \([S \cap T, \overline{S \cap T}]\) are also mincuts. (Hint: Write \(S=S_{1} \cup(S \cap T)\) and \(T=T_{1} \cup(S \cap T)\) and use the fact that \([S, \bar{S}]\) and \([T, \bar{T}]\) are both mincuts.)
8.2.7 Describe a maximum flow algorithm similar to the Ford-Fulkerson algorithm which begins at \(t\) and constructs unsaturated paths \(P\) until \(s\) is reached. Given that \(P\) is a ts-path, how should the residual capacity of an edge be defined in this case?
8.2.8 Describe an alqorithm for finding an edge \(\overrightarrow{u v}\) in a network \(N\) such that the value of a max-flow \(f\) in \(N\) can be increased if \((\overrightarrow{u v})\) is increased. Prove that your algorithm is correct and find its complexity. Does there always exist such an edge \(\overrightarrow{u v}\) ? Explain.

\subsection*{8.3 Matchings and flows}

There is a marked similarity between matching theory and flow theory:
Matchings: A matching \(M\) in a graph \(G\) is maximum if and only if \(G\) contains no augmenting path.
Flows: A flow \(f\) in a network \(N\) is maximum if and only if \(N\) contains no augmenting path.
Hungarian algorithm: Construct alternating paths until an augmenting path is found.
Ford-Fulkerson algorithm: Construct unsaturated paths until an augmenting path is found.
The reason for this is that matching problems can be transformed into flow problems. Consider a bipartite graph \(G\), with bipartition \((X, Y)\) for which a max-matching is desired. Direct all the edges of \(G\) from \(X\) to \(Y\), and assign them a capacity of one. Add a source \(s\) and an edge \(s x\) for all \(x \in X\), with \(\operatorname{CAP}(\overrightarrow{s x})=1\). Add a target \(t\) and an edge \(y t\) for all \(y \in Y\), with \(\operatorname{CAP}(y t)=1\). Call the resulting network \(N\). This is illustrated in Figure 8.8. Now find a max-flow in \(N\). The flow-carrying edges of \([X, Y\) ] will determine a max-matching in \(G\). Because \((\overrightarrow{s x})=1\), there will be at most one flow-carrying edge out of each \(x \in X\). Since \(\operatorname{CAP}(\overrightarrow{y t})=1\), there will be at most one flow-carrying edge into \(y\), for each \(y \in Y\). The flow-carrying edges of \(N\) are called the support of the flow. An alternating path in \(G\) and an unsaturated path in \(N\) can be seen to be the same thing. If \(G\) is not bipartite there is no longer a direct correspondence between matchings and
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flows. However, it is possible to construct a special kind of balanced network such that a maximum balanced flow corresponds to a max-matching (see KOCAY and STONE [80] or FREMUTH-PAEGER and JUNGNI CKEL [45]).


\section*{FI GURE 8.8}

\section*{Matchings and flows}

The basic BF-FF algorithm can be improved substantially. As it is presented here, it constructs a breadth-first network of all shortest unsaturated paths until \(t\) is reached. At this point, \(f\) is augmented, and the process is repeated. There may be many augmenting paths available at the point when \(t\) is reached, but only one augmenting path is used. The remaining unsaturated paths which have been built are discarded, and a new BFS is executed. In order to improve the BF-FF algorithm, one possibility is to construct the set of all shortest unsaturated paths. This is the auxiliary network of Exercise 8.2.4. We then augment on as many paths as
possible in the auxiliary network before executing a new BFS. This has the effect of making \(d j+1(s, t)>d j(s, t)\) so that the number of iterations is at most \(n\). Several algorithms are based on this strategy. They improve the complexity of the algorithm markedly. See the book by PAPADIMITRIOU and STEIGLITZ [94] for further information.

\section*{Exercises}
8.3.1 Let \(G\) be a bipartite graph with bipartition \((X, Y)\). We want to find a subgraph \(H\) of \(G\) such that in \(H\), \(\operatorname{DEG}(x)=b(x)\) and \(\operatorname{DEG}(y)=b(y)\), where \(b(u)\) is a given non-negative integer, for all \(v \in V(G)\), if there exists such an \(H\). For example, if \(b(u)=1\) for all \(u\), then \(H\) would be a perfect matching. If \(b(u)=2\) for all \(u\), then \(H\) would be a 2 -factor. Show how to construct a network \(N\) such that a max-flow

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in \(N\) solves this problem.
8.3.2 Let \(N\) be a network such that every vertex \(v \in N\) has a maximum throughput \(t(u)\) defined. This is the maximum amount of flow that is allowed to pass through \(u\), that is, \(f-(u) \leq t(U)\) must hold at all times. Show how to solve this problem by constructing a network \(N^{\prime}\) such that a max-flow in \(N^{\prime}\) defines a max-flow in \(N\) with maximum throughput as given.

\subsection*{8.4 Menger's theorems}

Given any digraph, we can view it as a network \(N\) by assigning unit capacities to all edges. Given any two vertices \(s\), \(t \in V(N)\), we can compute a max-flow \(f\) from \(s\) to \(t\). If \(\operatorname{VAL}(f)=0\), then there are no directed paths from \(s\) to \(t\), since a directed st-path would be an augmenting path. If VAL \((f)=1\), then \(N\) contains a directed st-path \(P\); however, there are no directed st-paths which are edgedisjoint from \(P\), for such a path would be an augmenting path. In general, the value of a max-flow \(f\) in \(N\) is the maximum number of edge-disjoint directed \(s t\)-paths in \(N\). Suppose that \(\operatorname{VAL}(f)=k \geq 1\). The support of \(f\) defines a subgraph of \(N\) that contains at least one directed st-path \(P\). Delete the edges of \(P\) to get \(N^{\prime}\) and let \(f^{\prime}\) be obtained from \(f\) by ignoring the edges of \(P\). Then \(\operatorname{VAL}\left(f^{\prime}\right)=k-1\), and this must be a max-flow in \(N^{\prime}\), since \(f\) is a max-flow in \(N\). By induction, the number of edge-disjoint directed st-paths in \(N^{\prime}\) is \(k-1\), from which it follows that the number in \(N\) is \(k\).
A min-cut in \(N\) can also be interpreted as a special subgraph of \(N\). Let \(K=[S, \bar{S}]\) be a min-cut in \(N\), where \(s \in S\) and \(t \in \bar{S}\). If \(\operatorname{CAP}(K)=0\), there are no edges out of \(S\), so there are no directed \(s t\)-paths in \(N\). If \(\operatorname{CAP}(K)=1\), there is only one edge out of \(S\). The deletion of this edge will destroy all directed st-paths in \(N\). We say that \(s\) is disconnected from \(t\). In general, CAP \((K)\) equals the minimum number of edges whose deletion destroys all directed st-paths in \(N\). Suppose that \(\operatorname{CAP}(K)=k \geq 1\). Delete the edges of \(K\) to get a network \(N^{\prime}\). Then in \(N^{\prime}, \operatorname{CAP}([S, \bar{S}])=0\), so that \(N^{\prime}\) contains no directed st-paths. Thus the deletion of the edges of \(K\) from \(N\) destroys all directed st-paths. Since \(N\) contains \(k\) edge-disjoint such paths, it is not possible to delete fewer than \(k\) edges in order to disconnect \(s\) from \(t\). The max-flow-min-cut theorem now gives the first of Menger's theorems.
THEOREM 8.7 Let \(s\) and \(t\) be vertices of a directed graph \(N\). Then the maximum number of edge-disjoint directed st-paths equals the minimum number of edges whose deletion disconnects \(s\) from \(t\).
Recall that an undirected graph \(G\) is \(k\)-edge-connected if the deletion of fewer
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than \(k\) edges will not disconnect \(G\). In Chapter 6 we showed that a graph is 2 -edge-connected if and only if every pair of vertices is connected by at least two edge-disjoint paths. We will use Theorem 8.7 to prove a similar result for \(k\)-edge-connected graphs. In order to convert Theorem 8.7 to undirected graphs, we can replace each edge \(\overrightarrow{u v}\) of \(G\) by a "qadget", as shown in Figure 8.9, to get a directed graph \(N\).

FI GURE 8.9

\section*{A gadget for edge-disjoint paths}

The gadget contains a directed \(\overrightarrow{u v}\)-path and a directed \(\overrightarrow{v u}\)-path, but they both use the central edge of the gadget. Let \(s, t \in V(G)\). Then edge-disjoint st-paths of \(G\) will define edge-disjoint directed st-paths in \(N\). Conversely, edge-disjoint directed st-paths in \(N\) will define edge-disjoint st-paths in \(G\). This gives another of

Menger's theorems.
THEOREM 8.8 Let \(s\) and \(t\) be vertices of an undirected graph \(G\). Then the maximum number of edge-disjoint st-paths equals the minimum number of edges whose deletion disconnects \(s\) from \(t\).
It follows that a graph \(G\) is \(k\)-edge-connected if and only if every pair \(s, t\) of vertices are connected by at least \(k\) edge-disjoint paths. This immediately gives an algorithm to compute \(\kappa^{\prime}(G)\), the edge-connectivity of \(G\). Number the vertices of \(G\) from 1 to \(n\). Let the corresponding vertices of \(N\) also be numbered from 1 to \(n\). The algorithm computes the minimum max-flow over all pairs \(s, t\) of vertices. This is the minimum number of edges whose deletion will disconnect G. Exactly \(\binom{n}{2}\) max-flows are computed, so the algorithm has polynomial complexity.
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Algorithm 8.4.1: EDGE-CONNECTIVITY(G)
convert \(G\) to a directed graph \(N\)
\(\kappa^{\prime} \leftarrow n\)
for \(s \leftarrow 1\) to \(n-1\)
do \(\left\{\begin{array}{l}\text { for } t \leftarrow s+1 \text { to } n \\ \text { do }\left\{\begin{array}{l}M \leftarrow \operatorname{MAXFLOW}(N, s, t) \\ \text { if } M<\kappa^{\prime} \\ \text { then } \kappa^{\prime} \leftarrow M\end{array}\right.\end{array}\right.\)
return ( \(\kappa^{\prime}\) )

\section*{Exercises}
8.4.1 Let \(G\) be an undirected graph. Replace each edge \(u u\) of \(G\) with a pair of directed edges \(\overrightarrow{u v}\) and \(\overrightarrow{v u}\) to get a directed graph \(N\). Let \(s, t \in V(G)\). Show that the maximum number of edge-disjoint st-paths in \(G\) equals the maximum number of edge-disjoint directed \(s t\)-paths in \(N\).
8.4.2 Program the edge-connectivity algorithm, using the transformation of Exercise 8.4.1.

\subsection*{8.5 Disjoint paths and separating sets}

Recall that paths in a graph \(G\) are internally disjoint if they can intersect only at their endpoints. A graph is \(k\) connected if the deletion of fewer than \(k\) vertices will not disconnect it. We proved in Chapter 6 that \(G\) is 2connected if and only if every pair of vertices is connected by at least two internally disjoint paths. We prove a similar result for \(k\)-connected graphs by utilizing a relation between internally disjoint paths in \(G\) and directed paths in a network \(N\). We first make two copies \(u 1\), \(u 2\) of each vertex \(u\) of \(G . V(N)=\left\{u_{1}, u_{2} \mid u \in V(G)\right\}\). Let \(\overrightarrow{u v}\) be an edge of \(G\). \(N\) will contain the edges ( \(u 1, u 2\) ), (u1, u2), (u2,u1), and ( \(u 2, u 1\) ). This is illustrated in Figure 8.10. Let \(u \in V(G)\). Notice the following observations:
1. The only out-edge at \(u 1\) is \(u 1 u 2\).
2. The only in-edge at \(u 2\) is \(u 1 u 2\).
3. The edge \(\overrightarrow{u v} \in E(G)\) corresponds to \(u 2 u 1\) and \(u 2 u 1\) in \(N\).

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\(u_{1}\)


\section*{FI GURE 8.10}

A gadget for internally disjoint paths
Consequently any st-path suuw...t in G corresponds to an s2t1-path s2u1u2u1u2w1w2...t1 in N. Internally disjoint st-paths in \(G\) give rise to internally disjoint \(s 2 t 1\)-paths in \(N\). On the other hand, edge-disjoint paths in \(N\) are in fact internally disjoint because of items 1 and 2 . Therefore the maximum number of internally disjoint st-paths in \(G\) equals the maximum number of edge-disjoint directed \(s 2 t 1\)-paths in \(N\). This in turn equals the
minimum number of edges whose deletion will disconnect \(s 2\) from \(t 1\). If \(s_{2} \nrightarrow t_{1}\), then every s2t1-path in \(G\) will contain another vertex, say \(u 1\) or \(u 2\). By observations 1 and 2, it must contain both \(u 1\) and \(u 2\). Deleting ulu2 will destroy this path. If \(K=[S, \bar{S}]\) is a min-cut in \(N\), then the edges out of \(S\) must be of the form \(u 1 u 2\), since \(u 2\) can only be s2-reachable if \(u 1\) is, by observation 3 and Figure 8.10. Let \(U=\left\{u \mid u_{1} u_{2} \in K\right\}\). Then \(U\) is a set of vertices of \(G\) which separate \(s\) from \(t\). This gives another of Menger's theorems.
THEOREM 8.9 Let \(G\) be a graph and let \(s, t \in V(G)\), where \(s \nrightarrow t\). Then the maximum number of internally disjoint st-paths in \(G\) equals the minimum number of vertices whose deletion separates \(s\) from \(t\).
THEOREM 8.10 A graph \(G\) is \(k\)-connected if and only if every pair of vertices is connected by at least \(k\) internally disjoint paths.
PROOF Let \(s, t \in V(G)\). If \(s\) and \(t\) are connected by at least \(k\) internally disjoint paths, then clearly \(G\) is \(k\) connected; for at least \(k\) vertices must be deleted to disconnect \(s\) from \(t\). Conversely suppose that \(G\) is \(k-\) connected. Then deleting fewer than \(k\) vertices will not disconnect \(G\). If \(s \nrightarrow t\), then by the Theorem 8.10, \(G\) must contain at least \(k\) internally disjoint st-paths. If \(s \longrightarrow t\), then consider \(G-s t\). It is easy to see that \(G-s t\) is \((k-1)\)-connected. Therefore in \(G-s t\), there are at least \(k-1\) internally disjoint st-paths. The edge st is another st-path, giving \(k\) paths in total.
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We can also use this theorem to devise an algorithm which computes \(\kappa(G)\), the connectivity of \(G\). We suppose that the vertices of \(G\) are numbered 1 to \(n\), and that if \(s \in V(G)\), then \(s 1\) and \(s 2\) are the corresponding vertices of \(N\).

Algorithm 8.5.1: CONNECTIVITY(G)
convert \(G\) to a directed graph \(N\)
\(\kappa \leftarrow n-1\) "maximum possible connectivity"
\(s \leftarrow 0\)
while \(s<\kappa\)
\(s \leftarrow s+1 \quad\) "vertex \(s\) "
for \(t \leftarrow s+1\) to \(n\)
do if \(s \nrightarrow t\)
do
\(\left\{\begin{array}{l}M \leftarrow \operatorname{MaxFLOW}\left(N, s_{2}, t_{1}\right) \\ \text { if } M<\kappa \\ \text { then } \kappa \leftarrow M \\ \text { if } s>\kappa \\ \text { then return }(\kappa)\end{array}\right.\)
return ( \(\kappa\) )
THEOREM 8.11 Algorithm 8.5.1 computes \(\kappa(G)\)
PROOF Suppose first that \(G=K n\). Then \(\kappa(G)=n-1\). The algorithm will not call MAXFLOW() at all, since every \(s\) is adjacent to every \(t\). The algorithm will terminate with \(\kappa=n-1\). Otherwise \(G\) is not complete, so there exists a subset \(U \subseteq V(G)\) such that \(G-U\) has at least two components, where \(|U|=\kappa(G)\). The first \(\kappa(G)\) choices of vertex \(s\) may all be in \(U\). However, by the ( \(\kappa(G)+1\) )st choice of \(s\) we know that some \(s \notin U\) has been selected. So \(s\) is in some component of \(G-U\). The inner loop runs over all choices of \(t\). One of these choices will be in a different component of \(U\). For that particular \(t\), the value of \(\operatorname{MAXFLO} W(N, s 2, t 1)\) will equal |U|. After this, the value of \(\kappa\) in the algorithm will not decrease any more. Therefore we can conclude that some \(s \notin U\) will be selected; that the value of \(\kappa\) after that point will equal \(\kappa(G)\); and that after this point the algorithm can stop. This is exactly what the algorithm executes.
The algorithm makes at most
\[
\sum_{s=1}^{\kappa+1}(n-s)
\]
calls to MAXFLOW(). Thus, it is a polynomial algorithm.

\section*{Exercises}
8.5.1 Let \(G\) be \(k\)-connected. If \(s t \in E(G)\), prove that \(G\)-st is ( \(k-1\) )-connected.

\subsection*{8.5.2 Program Algorithm 8.5.1, the CONNECTIVITY() algorithm.}
8.5.3 Consider a network \(N\) where instead of specifying a capacity for each edge \(\overrightarrow{u v}\), we specify a lower bound \(b(\overrightarrow{u v}) \geq 0\) for the flow on edge \(\overrightarrow{u v}\). Instead of the capacity constraint \(f(\overrightarrow{u v}) \leq \operatorname{CAP}(\overrightarrow{u v})\), we now have a lower bound constraint \(f(\overrightarrow{u v}) \geq b(\overrightarrow{u v})\). The zero-flow is not a valid flow anymore. Show that there exists a valid flow in such a network if and only if for every edge \(\overrightarrow{u v}\) such that \(b(\overrightarrow{u v})>0, \overrightarrow{u v}\) is either: (i) on a directed st-path; or (ii) on a directed ts-path; or (iii) on a directed cycle. (Hint: If \(\overrightarrow{u v}\) is not on such a path or cycle, follow directed paths forward from \(u\) and backward from \(u\) to get a contradiction.)
8.5.4 Consider the problem of finding a minimum flow in a network with lower bounds instead of capacities.
(a) How should an unsaturated path be defined?
(b) How should the capacity of an edge-cut be defined?
(c) Find a min-flow in the network of Figure 8.11, where the numbers are the lower bounds. Prove that your flow is minimum by illustrating an appropriate edge-cut.
(d) Is there a max-flow in the network given in Figure 8.11?


\section*{FI GURE 8.11}

\section*{A network with lower bounds}
8.5.5 Suppose that a network \(N\) has both lower bounds \(b(\overrightarrow{u v})\) and capacities \(\operatorname{CAP}(\overrightarrow{u v})\) on its edges. We wish to find a max-flow \(f\) of \(N\), where \(b(\overrightarrow{u v}) \leq f(\overrightarrow{u v}) \leq \operatorname{CAP}(\overrightarrow{u v})\). Notice that zero-flow may be no longer a valid flow. Before applying an augmenting path algorithm like the FF algorithm, we must first find a valid flow.

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(a) Determine whether the networks of Figure 8.12 have a valid flow.
(b) How should residual capacity be defined?
(c) How should the capacity of an edge-cut be defined?


\section*{FI GURE 8.12}

\section*{Networks with lower bounds and capacities}
8.5.6 Let \(N\) be a network with lower bounds \(b(\overrightarrow{u v})\) and capacities \(\operatorname{CAP}(\overrightarrow{u v})\) specified on its edges. Before finding a max-flow in \(N\) we need to find a valid flow. Construct a network \(N^{\prime}\) as follows: Add a new source \(s^{\prime}\) and target \(t^{\prime}\). Join \(s^{\prime}\) to all vertices of \(N\). Join every vertex of \(N\) to \(t^{\prime}\). Add edges \(\overrightarrow{s t}\) and \(\overrightarrow{t s}\) to \(N^{\prime}\). The capacities in \(N^{\prime}\) are defined as follows:
\[
\begin{gathered}
\operatorname{CAP}^{\prime}\left(\overrightarrow{s^{\prime} u}\right)=\sum_{v} b(\overrightarrow{v u}), \text { (sum over in-edges at } u \in V(N) \text { ). } \\
\operatorname{CAP}^{\prime}\left(\overrightarrow{u t^{\prime}}\right)=\sum_{v} b(\overrightarrow{u v}),(\text { sum over out-edges at } u \in V(N)) . \\
\operatorname{CAP}^{\prime}(\overrightarrow{u v})=\operatorname{CAP}(\overrightarrow{u v})-b(\overrightarrow{u v}),(u, v \in V(N)) . \\
\operatorname{CAP}^{\prime}(\overrightarrow{s t})=\operatorname{CAP}^{\prime}(\overrightarrow{t s})=\infty .
\end{gathered}
\]

Prove that there exists a valid flow in \(N\) if and only if there is a flow in \(N^{\prime}\) that saturates all edges incident on \(s^{\prime}\).
8.5.7 Let \(N\) be a network such that there is a cost \(c(\overrightarrow{u v})\) of using edge \(\overrightarrow{u v}\), per unit of flow. Thus the cost of flow \(f(\overrightarrow{u v})\) edge \(\overrightarrow{u v}\) is \(f(\overrightarrow{u v}) c(\overrightarrow{u v})\). Devise an algorithm to find a max-flow of min-cost in \(N\).
8.5.8 The circulation of money in the economy closely resembles a flow in a network. Each node in the economy represents a person or organization that takes part in economic activity. The main differences are that there may be no limit to the capacities of the edges, and that flow may accumulate at a node if assets are growing. Any transfer of funds is represented by a flow on some edge. Various nodes of the economic network can be organized into groups, such as banks, insurance companies, wage earners, shareholders, government, employers, etc.
(a) Develop a simplified model of the economy along these lines.

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(b) A bank charges interest on its loans. If there is a fixed amount of money in the economy, what does this imply? What can you conclude about the money supply in the economy?
(c) When a new business is created, a new node is added to the network. Where does the flow through this node come from?
(d) Consider the node represented by government. Where does its in-flow come from? Where is its out-flow directed? Consider how government savings bonds operate in the model.
(e) Where does inflation fit into this model?
(f) How do shareholders and the stock market fit into the model?

\subsection*{8.6 Notes}

The max-flow algorithm is one of the most important algorithms in graph theory, with a great many applications to other graph theory problems (such as connectivity and Menger's theorems), and to problems in discrete optimization. The original algorithm is from FORD and FULKERSON [43]. See also FULKERSON [47]. EDMONDS and KARP [39] proved that the use of shortest augmenting paths results in a polynomial time complexity of the Ford-Fulkerson algorithm.
Balanced flows were introduced by KOCAY and STONE [80], and then developed greatly in a series of papers by Fremuth-Paeger and J ungnickel. An excellent summary with many references can be found in FREMUTHPAEGER and JUNGNI CKEL [45].

A great many techniques have been developed to improve the complexity of the basic augmenting path algorithm. See PAPADIMITRI OU and STEIGLITZ [94] for further information. The algorithms to find the connectivity and edge-connectivity of a graph in Sections 8.4 and 8.5 are from EVEN [40].
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9
Hamilton Cycles

\subsection*{9.1 Introduction}

A cycle that contains all vertices of a graph \(G\) is called a hamilton cycle (or hamiltonian cycle). \(G\) is
hamiltonian if it contains a hamilton cycle. For example, Figure 9.1 shows a hamilton cycle in the graph called the truncated tetrahedron. It is easy to see that the graph of the cube is also hamiltonian (see Chapter 1).


FI GURE 9.1

\section*{A hamiltonian graph}

Figure 9.2 shows a non-hamiltonian graph \(H\). It is easy to see that \(H\) is non-hamiltonian, since it is bipartite with an odd number of vertices. Clearly any bipartite graph that is hamiltonian must have an even number of vertices, since a hamilton cycle \(C\) must start and end on the same side of the bipartition. Although \(H\) is nonhamiltonian, it does have a hamilton path, that is, a path containing all its vertices.
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\section*{FI GURE 9.2}

\section*{A non-hamiltonian graph}

The problem of deciding whether a given graph is hamiltonian is only partly solved.
Problem 9.1: HamCycle
Instance: a graph \(G\)
Question: is \(G\) hamiltonian?
This is an example of an NP-complete problem. We will say more about NP-complete problems later. There is no known efficient algorithm for solving the HamCycle problem. Exhaustive search algorithms can take a very long time in general. Randomized algorithms can often find a cycle quickly if \(G\) is hamiltonian, but do not give a definite answer if no cycle is found.
The HamCycle problem is qualitatively different from most other problems in this book. For example, the
questions "is G bipartite, Eulerian, 2-connected, planar?", and, "is a given flow \(f\) maximum?" can all be solved by efficient algorithms. In each case an algorithm and a theoretical solution are available. For the HamCycle problem, there is no efficient algorithm known, and only a partial theoretical solution. A great many graph theoretical problems are NP-complete.
A number of techniques do exist which can help to determine whether a given graph is hamiltonian. A graph with a cut-vertex \(u\) cannot possibly be hamiltonian, since a hamilton cycle \(C\) has no cut-vertex. This idea can be generalized into a helpful lemma.
LEMMA 9.1 If \(G\) is hamiltonian, and \(S \subseteq V(G)\), then \(w(G-S) \leq|S|\).
This lemma says that if we delete \(k=|S|\) vertices from \(G\), the number of
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connected components remaining is at most \(k\). Let \(C\) be a hamilton cycle in \(G\). If we delete \(k\) vertices from \(C\), the cycle \(C\) will be decomposed into at most \(k\) paths. Since \(C\) is a subgraph of \(G\), it follows that \(G-S\) will have at most \(k\) components.
For example, the graph of Figure 9.3 is non-hamiltonian, since the deletion of the three shaded vertices gives four components.


\section*{FI GURE 9.3}

\section*{A non-hamiltonian graph}

The Petersen graph is also non-hamiltonian, but this cannot be proved using Lemma 9.1. Instead we use an exhaustive search method called the multi-path method, see RUBIN [105]. Suppose that \(C\) were a hamilton cycle in the Petersen graph \(G\), as shown in Figure 9.4. \(G\) is composed of an outer and inner pentagon, joined by a perfect matching. Since \(C\) uses exactly two edges at each vertex of \(G\), it follows that \(C\) must use at least three edges of the outer pentagon, for otherwise some vertex on it would be missed by C. Consequently, C uses two adjacent edges of the outer pentagon. Without loss of generality, suppose that it uses the edges uu and uw. This means that \(C\) does not use the edge uy, so we can delete it from \(G\). Deleting uy reduces the degree of \(y\) to two, so that now both remaining edges at \(y\) must be part of \(C\). So the two paths ( \(u, u, w\) ) and \((x, y, z)\) must be part of \(C\), where a path is denoted by a sequence of vertices. This is illustrated in Fiqure 9.4. \(C\) must use two edges at \(w\), so there are two cases. Either \(w t \in C\) or \(w r \in C\). Suppose first that \(w t \in C\). Then since \(w r \notin C\), we can delete wr from \(G\). This reduces the degree of \(r\) to two, so that the remaining edges at \(r\) must be in \(C\). Therefore \(r z \in C\). This uses up two edges at \(z\), so we delete \(s z\), which in turn reduces the degree of \(s\) to two. Consequently the edge \(u s \in C\). But this now creates a cycle \((u, u, w, t, s)\) in \(C\), which is not possible. It follows that the choice
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FI GURE 9.4

\section*{The multi-path method}
\(w t \in C\) was wrong. If we now try \(w r \in C\) instead, a contradiction is again reached, thereby proving that the Petersen graph is non-hamiltonian.
This is called the multi-path method, since the cycle \(C\) is gradually built from a number of paths which are forced by two operations: the deletion of edges which are known not to be in \(C\); and the requirement that both edges at a vertex of degree two be in C. The multi-path method is very effective in testing whether 3regular graphs are hamiltonian, since each time an edge is deleted, the degree of two vertices reduces to two, which then forces some of the structure of \(C\). Graphs of degree four or more are not so readily tested by it. We will say more about the multi-path method later on.

\section*{Exercises}
9.1.1 Decide whether or not the graphs in Figure 9.5 are hamiltonian.
9.1.2 Prove that Qn, the \(n\)-cube, is hamiltonian for all \(n \geq 2\).
9.1.3 Let \(P\) be a hamilton path in a graph \(G\), with endpoints \(u\) and \(u\). Show that \(w(G-S) \leq|S|+1\), for all \(S \subseteq V(G)\), in two ways:
a) By counting the components of \(G-S\).
b) By counting the components of \((G+u u)-S\), and using Lemma 9.1.
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FIGURE 9.5

\section*{Are these hamiltonian?}

\subsection*{9.2 The crossover algorithm}

Suppose that we want to find a hamilton cycle in a connected graph \(G\). Since every vertex of \(G\) must be part of \(C\), we select any vertex \(x\). We then try to build a long path \(P\) starting from \(x\). Initially \(P=(x)\) is a path of length zero. Now execute the following steps:
\(u \leftarrow x ; u \leftarrow x\) " \(P\) is a \(u u\)-path"
while \(\exists w \longrightarrow u\) such that \(w \notin P\)
\[
\text { do }\left\{\begin{array}{l}
P \leftarrow P+u w \\
u \leftarrow w
\end{array}\right.
\]
while \(\exists w \longrightarrow v\) such that \(w \notin P\)
\[
\text { do }\left\{\begin{array}{l}
P \leftarrow P+v w \\
v \leftarrow w
\end{array}\right.
\]

The first loop extends \(P\) from \(u\) and the second loop extends \(P\) from \(u\), until it cannot be extended anymore. At this point we have a uu-path
\[
P=(u, \ldots, x, \ldots, u)
\]
such that the endpoints \(u\) and \(u\) are adjacent only to vertices of \(P\). The length of \(P\) is \(\ell(P)\), the number of edges in \(P\). The vertices of \(P\) are ordered from \(u\) to \(u\). If \(w \in P\), then \(w+\) indicates the vertex following \(w\) (if \(w \neq u\) ). Similarly \(w\) - indicates the vertex preceding \(w(\) if \(w \neq u\) ).
If \(u \longrightarrow v\), then we have a cycle \(C=P+u u\). If \(C\) is a hamilton cycle, we are done. Otherwise, since \(G\) is connected, there is a vertex \(w \in P\) such that

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\(w \longrightarrow y\), where \(y \notin P\). Hence there exists a longer path
\[
P^{*}=P-W W++w y .
\]

This is illustrated in Fiqure 9.6.


\section*{FI GURE 9.6}

\section*{Finding a long path}

If \(u \nprec v\), it may still be possible to find a cycle. Suppose that \(P\) contains a vertex \(w\) such that \(v \longrightarrow w\) and \(u \longrightarrow w^{+}\). This creates a pattern called a crossover, which is shown in Figure 9.7. When a crossover exists, there is a cycle
\[
C=P+u w-w w++u w+
\]
containing all the vertices of \(P\).


FI GURE 9.7
A crossover
Having converted the path \(P\) to a cycle \(C\) using the crossover, we are again in the situation where either \(C\) is a hamilton cycle, or else it contains a vertex \(w \longrightarrow y \notin C\), which allows us to find a longer path \(P^{*}\). We now extend \(P^{*}\) from both endpoints as far as possible, and then look for a crossover again, etc. The algorithm terminates either with a hamilton cycle, or with a long path that has no crossover. The crossover algorithm is summarized in Algorithm 9.2.1.
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\section*{Algorithm 9.2.1: LONGPATH \((G, x)\)}
comment: \(\left\{\begin{array}{l}\text { Find a long path in } G \text { containing } x, \text { using crossovers. } \\ P \text { and } C \text { are linked lists. }\end{array}\right.\)
\(u \leftarrow x ; u \leftarrow x ; P \leftarrow(x)\) "a path of length 0 "
repeat
comment: extend \(P\) from \(u\)
while \(\exists w \longrightarrow u\) such that \(w \notin P\)
\[
\text { do }\left\{\begin{array}{l}
\text { add } w \text { to } P \\
u \leftarrow w
\end{array}\right.
\]
comment: extend \(P\) from \(u\).
> while \(\exists w \longrightarrow v\) such that \(w \notin P\) do \(\left\{\begin{array}{l}\operatorname{add} w \text { to } P \\ v \leftarrow w\end{array}\right.\)

comment: search for a crossover
for all \(w \longrightarrow v\) do if \(u \longrightarrow w^{+}\) comment: a crossover has been found
\[
\left\{\begin{array}{l}
C \leftarrow P+v w-w w^{+}+u w^{+} \\
\text {if } C \text { is a hamilton cycle } \\
\text { then go to } 1 \\
\text { find } z \in C \text { such that } z \longrightarrow y \notin C \\
\text { convert } C+z y \text { into a path } P \text { from } y \text { to } z^{+} \\
u \leftarrow y ; \quad v \leftarrow z^{+}
\end{array}\right.
\]
then
until no crossover was found

\section*{1: comment: \(P\) can be extended no more}

\subsection*{9.2.1 Complexity}

The main operations involved in the algorithm are extending \(P\) from \(u\) and \(u\), converting \(P\) to a cycle \(C\), and finding \(z \in C\) such that \(z \longrightarrow y \notin C\). We assume that the data structures are arranged so that the algorithm can check whether or not a vertex \(w\) is on \(P\) in constant time. This is easy to do with a boolean array. We also assume that the algorithm can test whether or not vertices \(u\) and \(w\) are adjacent in constant time.
- Extending \(P\) from \(u\) requires at most DEG(u) steps, for each \(u\). Since \(P\) can extend at most once for each \(u\), the total number of steps taken to extend \(P\) is at most \(\sum_{u} \operatorname{DEG}(u)=2 \varepsilon\), taken over all iterations of the algorithm.
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- Converting \(P\) to a cycle \(C=P+u w-w w++u w+\) requires reversing a portion of \(P\). This can take up to \(\ell(P)\) steps. As \(\ell(P)\) increases from 0 up to its maximum, this can require at most \(O(n 2)\) steps, taken over all iterations.
- Checking whether \(z \in C\) is adjacent to some \(y \notin C\) requires \(\operatorname{DEG}(z)\) steps for each \(z\). There are \(\ell(C)=\ell(P)+1\) vertices \(z\) to be considered. If at some point in the algorithm it is discovered that some \(z\) is not adjacent to any such \(y\), we need never test that \(z\) again. We flag these vertices to avoid testing them twice.
Thus the total number of steps spent looking for \(z\) and \(y\) is at most \(O\left(n^{2}\right)+\sum_{z} \operatorname{DEG}(z)=O\left(n^{2}+\varepsilon\right)\).
So the total complexity of the algorithm is \(O(n 2+\varepsilon)\). More sophisticated data structures can reduce the \(O(n 2)\) term, but there is likely no reason to do so, since the algorithm is already fast, and it is not guaranteed to find a hamilton cycle in any case.
The crossover algorithm works very well on graphs which have a large number of edges compared to the number of vertices. In some cases we can prove that it will always find a hamilton cycle. On sparse graphs (e.g., 3-regular graphs), it does not perform very well when the number of vertices is more than 30 or so.

LEMMA 9.2 Let \(G\) be a graph on \(n\) vertices such that \(D E G(u)+D E G(u\} \geq n\), for all non-adjacent vertices \(u\) and \(U\). Then the crossover algorithm will always find a hamilton cycle in \(\mathcal{G}\).
PROOF If the crossover algorithm does not find a hamilton cycle, let \(P\) be the last path found. Since \(P\) can't be extended from its endpoints \(u\) and \(u\), it follows that \(u\) and \(u\) are joined only to vertices of \(P\). For each \(w \longrightarrow v\), it must be that \(u \nrightarrow w^{+}\), or a crossover would exist. Now \(u\) is joined to DEG(U) vertices of \(P\). There are thus DEG( \(u\) ) vertices that \(u\) is not joined to. Consequently \(u\) can be adjacent to at most \(\ell(P)-\) DEG \((U)\) vertices, where \(\ell(P) \leq n-1\) is the number of edges of \(P\). So we have
\[
\text { DEG }(u)+\operatorname{DEG}(u) \leq \ell(P) \leq n-1 \text {, }
\]
a contradiction, since we assumed that DEG(u)+DEG(u) \(\geq n\) for all non-adjacent \(u\) and \(u\).
This lemma also shows that graphs which satisfy the condition DEG(u)+ DEG(u) \(\geq n\) are always hamiltonian. Such graphs have many edges, as we shall see. However, the crossover algorithm will often find hamilton cycles or hamilton paths, even when a graph does not satisfy this condition.
The crossover algorithm can be improved enormously by searching for crossovers of higher order. The crossover of Figure 9.7 can be defined to be

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the trail \(Q=(u, w+w, u)\) which starts at \(u\), intersects \(P\) in exactly one edge, and finishes at \(u\). The cycle \(C\) is then given by \(C=P \oplus Q\), where \(\oplus\) indicates the operation of exclusive-OR, applied to the edges of \(P\) and \(Q\). In general, higher order crossovers can be defined as follows.
DEFINITION 9.1: Let \(P\) be a uu-path. A crossover \(Q\) is a uu-trail such that \(V(Q) \subseteq V(P)\) and \(C=P \oplus Q\) is a cycle with \(V(C)=V(P)\). The order of a crossover \(Q\) is the number \(|P \cap Q|\) of edges common to \(P\) and \(Q\). A cross-edge is any edge \(x y \in E(Q)-E(P)\).
So a crossover of order 0 occurs when \(u \longrightarrow v\). Then \(Q=(u, u)\) and \(C=P+u u\). There is only one kind of crossover of order one, which is shown in Figure 9.7. A crossover of order two is illustrated in Figure 9.8. There are five different kinds of crossover of order two, as the reader can verify by constructing them. An algorithm employing crossovers of order higher than one requires a recursive search for crossovers up to a pre-selected maximum order M. It was found by KOCAY and LI [78] that choosing \(M=6\) still gives a fast algorithm, and that it improves the performance of the basic algorithm enormously. This algorithm requires sophisticated data structures for an efficient implementation.

\section*{\(P\)}


\section*{FI GURE 9.8}

A crossover \(Q=(u, w, w+, x, x+, u)\) of order two
Suppose that a path \(P\) is the longest path found by the algorithm, and that it has no crossover. If there is a vertex \(x \notin P\) such that \(x \rightarrow w\), w+, for some \(w \in P\), then we can make a longer path by re-routing \(P\) through \(x\) : \(P^{\prime}=(\ldots, w, x, w+\ldots)\). Similarly, a configuration like Figure 9.9 can also be used to give a longer path. Once \(P\) has been re-routed to a longer path, we can again check for a crossover. When used in combination, crossovers and reroutings will very often find a hamilton cycle in \(G\), if it is hamiltonian, even for sparse graphs \(G\).
A re-routing is very much like a crossover. It is a closed trail \(Q\) whose endpoints are on the uu-path \(P\), such that \(P \oplus Q\) is a uu-path containing all vertices of \(P\). It always results in a longer path. The algorithm that searches for higher order crossovers can be easily modified to search for re-routings as well.
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\section*{FI GURE 9.9}

\section*{Re-routing \(P\)}

\section*{Exercises}
9.2.1 Show that if \(G\) is connected and \(n>2 \delta\), where \(\delta\) is the minimum degree of \(G\), then \(G\) has a path of length at least 2 \(\delta\). This is due to DIRAC [36]. (Hint: Consider a longest path.)
9.2.2 Program the crossover algorithm, and test it on the Petersen graph, on the graphs of Figure 9.5, and on the graph in Figure 9.10. Try it from several different starting vertices.


FI GURE 9.10

\section*{The Lederberg graph}
9.2.3 Let \(G\) be a graph. Show how to create a graph \(G^{\prime}\) from \(G\) by adding one vertex so that \(G\) has a hamilton path if and only if \(G^{\prime}\) has a hamilton cycle.
9.2.4 Let \(G\) be a graph such that \(\operatorname{DEG}(u)+D E G(u) \geq n-1\), for all non-adjacent vertices \(u\) and \(u\). Show that \(G\) has a hamilton path.
9.2.5 Construct all five kinds of crossover of order two.
9.2.6 Construct the crossovers of order three.
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\subsection*{9.3 The Hamilton closure}

Suppose that \(\operatorname{DEG}(u)+\operatorname{DEG}(u) \geq n\) in a graph \(G\), where \(u\) and \(u\) are non-adjacent vertices. Let \(G^{\prime}=G+u v\). If \(G\) is hamiltonian, then so is \(G^{\prime}\). Conversely, if \(G^{\prime}\) is hamiltonian, let \(C\) be a hamilton cycle in \(G^{\prime}\). If \(u v \in C\), then \(P=C-u v\) is a hamilton path in \(G\). Since \(\operatorname{DEG}(u)+D E G(u) \geq n\), we know that \(P\) has a crossover, so that \(G\) has a hamilton cycle, too. Thus we have proved:
LEMMA 9.3 Let DEG(u)+DEG(u) \(\geq n\) in a graph \(G\), for non-adjacent vertices \(u\) and \(u\). Let \(G^{\prime}=G+u u\). Then \(G\) is hamiltonian if and only if \(G^{\prime}\) is.
This lemma says that we can add all edges \(u u\) to \(G\), where \(\operatorname{DEG}(u)+D E G(u) \geq n\), without changing the hamiltonicity of \(G\). We do this successively, for all non-adjacent vertices \(u\) and \(u\).
DEFINITION 9.2: The hamilton closure of \(G\) is \(c H(G)\), the graph obtained by successively adding all edges uu to \(G\), whenever DEG \((u)+\operatorname{DEG}(u) \geq n\), for non-adjacent vertices \(u\) and \(u\).
For example, the hamilton closure of the graph of Figure 9.11 is the complete graph \(K 7\). It must be verified that this definition is valid, namely, no matter in what order the edges \(u v\) are added to \(G\), the resulting closure is the same. We leave this to the reader.


FIGURE 9.11
\(\mathrm{cH}(\mathrm{G})=\mathrm{K} 7\)
Lemma 9.3 tells us that \(\mathrm{cH}(\mathrm{G})\) is hamiltonian if and only if \(G\) is. In particular, if \(\mathrm{cH}(\mathrm{G})\) is a complete graph, then \(G\) is hamiltonian. The hamilton closure can be used to obtain a condition on the degree sequence of \(G\) which will force \(G\) to be hamiltonian.
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where \(d 1 \leq d 2 \leq \ldots \leq d n\). If there is no \(m<n / 2\) such that \(d m \leq m\) and \(d n-m<n-m\), then \(c H(G)\) is complete. PROOF Suppose that \(\mathrm{cH}(\mathrm{G})\) is not complete. Let \(u\) and \(u\) be non-adjacent vertices such that DEG(u)+DEG(u) is as large as possible, where the degree is computed in the closure \(c H(G)\). Then \(\operatorname{DEG}(u)+D E G(u)<n\) by definition of the closure. Let \(m=\operatorname{DEG}(u) \leq \operatorname{DEG}(u)\). So \(u\) is joined to \(m\) vertices. There are \(n-D E G(u)-1\) vertices that \(u\) is not adjacent to (not counting \(u\), since \(v \nrightarrow v\) Each of these has degree \(\leq m\). So the number of vertices with degree \(\leq m\) is at least \(n-\operatorname{DEG}(u)-1\). But \(\operatorname{DEG}(u)+D E G(u)<n\), so that \(m=\operatorname{DEG}(u) \leq n-\operatorname{DEG}(u)-1\). That is, the number of vertices of the closure with degree \(\leq m\) is at least \(m\). Since the deqree sequence of \(\mathrm{cH}(\mathrm{G})\) is at least as biq as that of \(G\), it follows that \(d m \leq m\).


\section*{FI GURE 9.12}

\section*{The degree sequence of \(G\)}

How many vertices have degree \(>\operatorname{DEG}(u)\) ? We know that \(u\) is adjacent to all of them. Therefore, the number of them is at most \(m\), so that there are at most \(m\) vertices after \(u\) in the degree sequence. It follows that \(\mathrm{DEG}(u) \geq d n-m\). But since \(\mathrm{DEG}(u)<n-m\), it follows that \(d n-m<n-m\). Thus, we have found a value \(m\) such that \(d m \leq m\) and \(d n-m<n-m\). Here \(m=\operatorname{DEG}(u) \leq \operatorname{DEG}(u)<n-m\), so that \(m<n / 2\). This contradicts the assumptions of the theorem. Therefore \(\mathrm{cH}(\mathrm{G})\) must be complete under these conditions.
The degree sequence condition of the Bondy-Chvátal theorem is easy to apply. For example, any graph with the degree sequence \((2,2,3,4,5,6,6,6)\) must be hamiltonian, since \(d 1=2>1, d 2=2 \leq 2\), but \(d_{8-2}=6 \nless 6\), and \(d B=3 \leq 3\), but \(d_{8-3}=5 \nless 5\). Thus there is no \(m<8 / 2\) satisfying the condition that \(d m \leq m\) and \(d n-m<n-m\). This is the strongest degree sequence condition possible which forces an arbitrary graph \(G\) to be hamiltonian. Any stronger condition would have to place non-degree sequence restrictions on \(G\). To see that this is so, let \(G\) be any non-hamiltonian graph. Let its degree sequence be \((d 1, d 2, \ldots, d n)\), where \(d 1 \leq d 2 \leq \ldots \leq d n\). Since \(G\) is not hamiltonian, there is a value \(m\) such that \(d m \leq m\) and \(d n-m<n-m\). Construct a new degree sequence by increasing
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each di until the sequence ( \(m, \ldots, m, n-m-1, \ldots, n-m-1, n-1, \ldots, n-1\) ) is obtained, where the first \(m\) degrees are \(m\), the last \(m\) degrees are \(n-1\), and the middle \(n-2 m\) degrees are \(n-m-1\). We construct a nonhamiltonian graph \(C(m, n)\) with this degree sequence.


\section*{FI GURE 9.13}
\(C(m, n)\)
\(C(m, n)\) is composed of three parts, a complete graph \(K m\), a complete graph \(K n-2 m\), and an empty graph \(\bar{K}_{m}\). Every vertex of \(\bar{K}_{m}\) is joined to every vertex of \(K m\), and every vertex of \(K m\) is joined to every vertex of \(K n-2 m\). This is illustrated in Figure 9.13. The vertices of \(\bar{K}_{m}\) have degree \(m\), those of \(K n-2 m\) have degree \(n-m-1\), while those of \(K m\) have degree \(n-1 . C(3,9)\) is illustrated in Figure 9.3 . It is easy to see that ( \(C(m\), n) is always non-hamiltonian, since the deletion of the vertices of \(K m\) leaves \(m+1\) components. By Lemma 9.1, we conclude that \(C(m, n)\) is non-hamiltonian. Yet for every non-hamiltonian graph \(G\) on \(n\) vertices, there is some \(C(m, n)\) whose degree sequence is at least as large as that of \(G\), in the lexicographic order.

\section*{Exercises}
9.3.1 Prove that \(\mathrm{cH}(\mathrm{G})\) is well-defined; that is, the order in which edges \(u \boldsymbol{u}\) are added to \(G\) does not affect the result.
9.3.2 Prove that the crossover algorithm will find a hamilton cycle in \(G\) if \(c H(G)\) is complete or find a counterexample.
9.3.3 Use the Bondy-Chvátal theorem to show that any graph with the degree sequence \((2,3,3,4,5,6,6,6\), 7) is hamiltonian. What about ( \(3,3,4,4,4,4,4,4\) )?
9.3.4 Define the hamilton-path closure to be \(c_{H}^{\prime}(G)\), obtained by adding all edges \(u u\) whenever \(\operatorname{DEG}(u)+\operatorname{DEG}(u) \geq n-1\). Prove that \(G\) has a hamilton path if and only \(c_{H}^{\prime}(G)\) does.
9.3.5 Obtain a condition like the Bondy-Chvátal theorem which will force \(c_{H}^{\prime}(G)\) to be complete.
9.3.6 Construct the graphs \(C(2,8)\) and \(C(4,12)\).
9.3.7 Work out \(\varepsilon(C(m, n))\). Show that \(\varepsilon\) has its smallest value when
\[
\begin{gathered}
m=\frac{n}{3}-\frac{1}{6} \\
\quad \text { page_199 }
\end{gathered}
\]

Page 200 for which
\[
\varepsilon=\frac{2}{3}\binom{n}{2}-\frac{1}{24}
\]
9.3.8 Show that if \(G\) is a graph on \(n \geq 4\) vertices with \(\varepsilon \geq\binom{ n-1}{2}+1\), then \(G\) is hamiltonian.

Notice that according to Exercise 9.3.7, \(C(m, n)\) has approximately two-thirds of the number of edges of the complete graph \(K n\), at the minimum. This means that degree sequence conditions are not very strong. They apply only to graphs with very many edges.

\subsection*{9.4 The extended multi-path algorithm}

The multi-path algorithm tries to build a hamilton cycle \(C\) using a recursive exhaustive search. At any stage of the algorithm, a number of disjoint paths \(S 1, S 2, \ldots, S k\) in \(G\) are given, which are to become part of C. Call them segments of \(C\). Initially, we can take \(k=1\), and the single segment \(S I\) can consist of the starting vertex, that is, a path of length zero. On each iteration a vertex \(u\) is selected, an endpoint of some segment \(P=S i\). Every \(w \longrightarrow u\) is taken in turn, and \(P\) is extended to \(P^{\prime}=P+u w\). Vertex \(u\) may now have degree two in Si. In this case, the remaining edges \(u x\) of \(G\) are deleted. This reduces each \(\operatorname{DEG}(x)\) by one. When \(\operatorname{DEG}(x)=2\), both remaining edges at \(x\) must become part of \(C\). A new segment is created containing \(x\). Thus, the choice of uw can force certain edges to be a part of \(C\). It can also happen that when edges are forced in this way, that an edge connecting the endpoints of two segments is forced, and the two segments must be merged into one. This in turn forces other edges to be deleted, etc. The forcing of edges can be performed using a queue. There are three possible outcomes of this operation:
1. An updated set of segments can be produced.
2. A hamilton cycle can be forced.
3. A small cycle can be forced.

By a small cycle, we mean any cycle smaller than a hamilton cycle. If a small cycle is forced, we know that the extension of \(P\) to \(P+u w\) does not lead to a hamilton cycle. If a hamilton cycle is forced, the algorithm can quit. If a new set of segments is produced, the algorithm proceeds recursively. This can be summarized as follows. We assume a global graph \(G\), and a global boolean variable IsHamiltonian, which is initially false, but is changed to true when a hamilton cycle is discovered.
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Algorithm 9.4.1: MULTIPATH(S)
comment: Search for a ham cycle containing all segments of \(S\)
choose a vertex \(u\), an endpoint of some path \(P \in S\)
for all \(w \longrightarrow u\)
(extend path \(P\) to \(P+u w\)
comment: extending \(P\) to \(P+u w\) may force some edges
FORCEEDGES \((u w)\)
if a hamilton cycle was forced
do
then \(\left\{\begin{array}{l}\text { IsHamiltonian } \leftarrow \text { true } \\ \text { return }\end{array}\right.\)
if a small cycle was not forced
then \(\left\{\begin{array}{l}\text { comment: the segments } S \text { have been updated } \\ \operatorname{MultiPath}(S) \\ \text { if } \text { IsHamiltonian } \\ \text { then return }\end{array}\right.\)
restore \(G\) and \(S\) to their state before \(u w\) was chosen
comment: otherwise no hamilton cycle was found
Suppose that the multi-path algorithm were applied to a disconnected graph \(G\). Although we know that \(G\) is
not hamiltonian, the algorithm could still take a very long time to discover this, for example, the connected components of \(G\) could be complete graphs. More generally, it is quite possible for the operation of forcing edges to delete enough edges so as to disconnect \(G\). Thus the algorithm really is obliged to check that \(G\) is still connected before making a recursive call. This takes \(O(\varepsilon)\) steps. Now we know that a graph with a cutvertex also has no hamilton cycle, and we can test for a cut-vertex at the same time as checking that \(G\) is connected. A depth-first search (DFS) can do both in \(O(\varepsilon)\) steps. Thus we add a DFS to the multi-path algorithm before the recursive call is made. But we can make a still greater improvement. Suppose that the multi-path algorithm were applied to the graph of Figure 9.14. This graph is nonhamiltonian because the deletion of the two shaded vertices leaves three components. In certain cases the algorithm is able to detect this, using the DFS that tests for cut-vertices. Suppose that the segments of \(G\) are the bold edges. Notice that one of the segments contains the shaded vertex \(u\). When the non-segment edges incident on \(u\) are deleted, \(u\) becomes a cut-vertex in the resulting graph. The DFS will detect that \(u\) is a cutvertex, and the algorithm will report that adding the edge uw to the segment does not extend to a hamilton cycle.
Normally the algorithm would then try the next edge incident on \(u\), etc. But it can do more. When the cutvertex \(u\) is discovered, the DFS can count the

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\section*{FI GURE 9.14}

\section*{A non-hamiltonian graph}
number of components of \(G-u\). This will be one plus the number of descendants of \(u\) in the DF-tree. It requires almost no extra work for the DFS to calculate this. For vertex \(u\) in Figure 9.14, the count will be three components. But since this is the result of deleting only two vertices, namely, \(u\) and \(u\), the algorithm can determine that the original \(G\) is non-hamiltonian, and stop the search at that point. More generally, a nonhamiltonian graph like Figure 9.14 can arise at some stage during the algorithm as a result of deleting edges, even though the original \(G\) is hamiltonian. The algorithm must be able to detect which graph in the search tree is found to be non-hamiltonian by this method. We leave it to the reader to work out the details. It is helpful to view vertices like \(u\) in Figure 9.14 which have degree two in some segment as having been deleted from \(G\). Each segment is then replaced by an equivalent single edge connecting its endpoints. The set of segments then becomes a matching in \(G\), which is changing dynamically. For example, when the segments of Figure 9.14 are replaced by matching edges, the resulting graph appears as in Figure 9.15. The procedure which forces edges can keep a count of how many vertices internal to segments have been deleted in this way, at each level in the recursion. When the DFS discovers a cut-vertex, this count is used to find the size of a separating set in \(G\). In cases like this, large portions of the search tree can be avoided.
A bipartite graph like the Herschel graph of Figure 9.2 is also non-hamiltonian, but the algorithm is not likely to delete enough vertices to notice that it has a large separating set. In general, suppose that at some stage in the algorithm \(G-E(S)\) is found to be bipartite, with bipartition \((X, Y)\), where \(S\) is viewed as a matching in \(G\). If there is a hamilton cycle \(C\) in \(G\) using the matching edges \(S\), it must


\section*{FI GURE 9.15}

\section*{Segments viewed as a matching}
look something like Figure 9.16, where the bipartition of \(G-E(S)\) is shown by the shading of the nodes. There are now three kinds of segments: those contained within \(X\), those contained within \(Y\), and those connecting \(X\) to \(Y\). Suppose that there are \(\varepsilon X\) of the first type, and \(\varepsilon Y\) of the second type. The vertices of \(C\) must alternate between \(X\) and \(Y\), except for the \(\varepsilon X\) and \(\varepsilon Y\) edges, which must have endpoints of the same color. If we contract each of these edges to a single node, we obtain perfect alternation around the cycle. Therefore \(|X|\) \(-\varepsilon X=|Y|-\varepsilon Y\) if \(G\) has a hamilton cycle. If this condition is not satisfied, we know that \(G\) is non-hamiltonian, and can break off the search. We again employ the DFS that tests for cut-vertices to simultaneously check whether \(G-E(S)\) is bipartite, and to keep a count of the numbers \(|X|-\varepsilon X\) and \(|Y|-\varepsilon Y\). This requires very little extra work, and is still \(O(\varepsilon)\). In this way, non-hamiltonian graphs derived from bipartite graphs or nearbipartite graphs can often be quickly found to be non-hamiltonian.
In summary, the extended multi-path algorithm adds a DFS before the recursive call. The DFS computes several things:
- Whether \(G\) is connected.
- \(w(G-u)\), for each cut-vertex \(u\).
- Whether \(G-E(S)\) is bipartite.
- \(|X|-\varepsilon X\) and \(|Y|-\varepsilon Y\), if \(G-E(S)\) is bipartite.

It may be possible to add other conditions to detect situations when \(G\) is non-hamiltonian. For example, every hamiltonian graph \(G\) with an even number of vertices \(n\) has two disjoint perfect matchings. If \(n\) is odd, every \(G-u\) has a perfect matching.

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\section*{FI GURE 9.16}

\section*{\(G-E(S)\) is bipartite}

\subsection*{9.4.1 Data structures for the segments}

The extended multi-path algorithm still has exponential worst-case running time. Operations on the segments must be made as fast as possible. The operations that must be performed using segments are, given any
vertex \(u\), to determine which segment contains \(u\), and to find its endpoints; and to merge two segments when their endpoints are joined. One way to do this is with the merge-find data structure. An array Segment[u] is stored, which is an integer, pointing to the representative of the segment containing \(u\). Each segment has two endpoints, which we arbitrarily designate as the right and left endpoints. The right endpoint \(x\) is the segment representative. It is indicated by a negative value of Segment[ \(x\) ]. Its value is \(-y\), where \(y\) is the left endpoint. Thus we find the segment representative by following the pointers, using path compression (see Chapter 2). Segments are merged by adjusting the pointers of their endpoints.

\section*{Exercises}
9.4.1 Program the multi-path algorithm. Use a DFS to test for the conditions mentioned above.
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9.5 Decision problems, NP-completeness

The theory of NP-completeness is phrased in terms of decision problems, that is, problems with a yes or no answer, (e.g., "is G hamiltonian?"). This is so that an algorithm can be modeled as a Turing machine, a theoretical model of computation. Although Turing machines are very simple, they can be constructed to execute all the operations that characterize modern random access computers. Turing machines do not usually produce output, except for yes or no. Thus, a Turing machine can be constructed to read in a graph, and perform an exhaustive search for a hamilton cycle. If a cycle exists, it will be found, and the algorithm will report a yes answer. However, the exhaustive search will tend to take an exponential amount of time in general.
The class of all decision problems contains an important subclass called \(P\), all those which can be solved in polynomial time; that is, the complexity of a problem is bounded by some polynomial in its parameters. For a graph, the parameters will be \(n\) and \(\varepsilon\), the number of vertices and edges.
There is another class of decision problems for which polynomial algorithms are not always known, but which have an additional important property. Namely, if the answer to a problem is yes, then it is possible to write down a solution which can be verified in polynomial time. The HamCycle problem is one of these. If a graph \(G\) has a hamilton cycle \(C\), and the order of vertices on the cycle is written down, it is easy to check in \(n\) steps that \(C\) is indeed a hamilton cycle. So if we are able to guess a solution, we can verify it in polynomial time. We say that we can write a certificate for the problem, if the answer is yes. A great many decision problems have this property that a certificate can be written for them if the answer is yes, and it can be checked in polynomial time. This forms the class NP of non-deterministic polynomial problems. The certificate can be checked in polynomial time, but we do not necessarily have a deterministic way of finding a certificate.
Now it is easy to see that \(\mathrm{P} \subseteq \mathrm{NP}\), since every problem which can be solved in polynomial time has a certificate-we need only write down the steps which the algorithm executed in solving it. It is generally believed that HamCycle is in NP but not in P. There is further evidence to support this conjecture beyond the fact that no one has been able to construct a polynomial-time algorithm to solve HamCycle; namely, it can be shown that the HamCycle problem is one of the NP-complete problems.
To understand what NP-completeness means we need the concept of polynomial transformations. Suppose \(\Pi 1\) and \(\Pi 2\) are both decision problems. A polynomial transformation from \(\Pi 1\) to \(\Pi 2\) is a polynomial-time algorithm which when given any instance 11 of problem \(\Pi 1\) will generate an instance 12 of problem \(\Pi 2\) satisfying:
\(I 1\) is a yes instance of \(\Pi 1\) if and only if \(I 2\) is a yes instance of \(\Pi 2\)
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We use the notation \(\Pi_{1} \propto \Pi_{2}\) to indicate that there is a polynomial transformation from \(\Pi 1\) to \(\Pi 2\). We say that \(\Pi 1\) reduces to \(\Pi 2\). This is because if we can find a polynomial algorithm \(A\) to solve II2, then we can transform \(\Pi 1\) into \(\Pi 2\), and then use \(A\) to solve \(\Pi 2\), thereby giving a solution to \(\Pi 1\).
DEFINITION 9.3: A decision problem is II is NP-complete, if
1. \(\Pi\) is in NP.
2. For any problem \(\Pi^{\prime} \in N P, \Pi^{\prime} \propto \Pi\).

It was Cook who first demonstrated the existence of NP-complete problems. He showed that Problem 9.2, satisfiability of boolean expressions (Sat) is NP-complete. Let \(U\) be a set of \(n\) boolean variables \(u 1, u 2, \ldots\), un with their complements \(\bar{u} 1, \bar{u} 2, \ldots, \bar{u} n\). These variables can only take on the values true and false, such that \(u i\) is true if and only if \(\bar{u} i\) is false, and vice versa. If \(x, y \in U\), then we denote by \(x+y\) the boolean or of \(x\) and \(y\) by \(x y\) the boolean and of \(x\) and \(y\). A clause over \(U\) is a sum of variables in \(U\). For example, \((u 1+\bar{u} 3+\bar{u} 4+u 6)\) is a clause. A boolean expression is a product of clauses. For example \((u 1+\bar{u} 3+\bar{u} 4+u 6)(u 2+u 5)(\bar{u} 7)\) is a boolean expression. A truth assignment \(t\) is an assignment of values true and false to the variables in \(U\). If \(B\) is a boolean expression, then \(t(B)\) is the evaluation of \(B\) with truth
assignment \(t\). For example if
\[
B=(u 1+\bar{u} 3+\bar{u} 4+u 6)(u 2+u 5)(\bar{u} 7)
\]
and
then
\[
t=\left(\begin{array}{ccccccc}
u_{1} & u_{2} & u_{3} & u_{4} & u_{5} & u_{6} & u_{7} \\
\text { true } & \text { false } & \text { false } & \text { true } & \text { true } & \text { false } & \text { false }
\end{array}\right)
\]
then
\(t(B)=(\) true + true+false+false)(false+true)(true)=true.
Not every boolean expression \(B\) has a truth assignment \(t\) such that \(t(B)=\) true. For example there is no way to assign true and false to the variables in the expression \((\bar{u} 1+\bar{u} 2)(u 1)(u 2)\) so that it is true. if there is a truth assignment \(t\) such that \(t(B)=\) true, we say that \(B\) is satisfiable. The satisfiability of boolean expressions problem is
Problem 9.2: Sat
Instance: a set of boolean variables \(U\) and a boolean expression \(B\) over \(U\).
Question:is B satisfiable?
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and was shown by COOK [28] to be NP-complete. See also KARP [70]. The proof of this is beyond the scope of this book; however, a very readable proof can be found in the book by PAPADIMITRIOU and STEIGLITZ [94]. Many problems have subsequently been proved NP-complete, by reducing them either to satisfiability, or to other problems already proved NP-complete.
The importance of the NP-complete problems is that, if a polynomial algorithm for any NP-complete problem II were discovered, then every problem in NP would have a polynomial algorithm; that is, \(\mathrm{P}=\mathrm{NP}\) would hold. Many people have come to the conclusion that this is not very likely, on account of the large number of NPcomplete problems known, all of which are extremely difficult. We will now show that

\section*{Sat \(\propto\) 3-Sat \(\propto\) Vertex Cover \(\propto\) HamCycle}
and thus the HamCycle problem (as well as 3-Sat and Vertex Cover) is an NP-complete problem. Thus if \(\mathrm{P} \neq \mathrm{NP}\), then a polynomial algorithm for the HamCycle problem would not exist. This is why we say that the HamCycle problem is qualitatively different from most other problems in this book.
Among the most useful problems for establishing the NP-completeness of other problems is 3-Sat.
Problem 9.3: 3-Sat
Instance: a set of boolean variables \(U\) and a boolean expression \(B\) over \(U\), in which each clause contains exactly three variables.
Question:is \(B\) satisfiable?
THEOREM 9.5 3-Sat is NP-complete.
PROOF It is easy to see that 3-Sat is in NP. Any truth assignment satisfying the boolean expression \(B\) can be checked in polynomial time by assigning the variables and then evaluating the expression.
We reduce Sat to 3 -Sat as follows. Let \(U\) be a set of boolean variables and \(B=C 1 C 2 \ldots C m\) be an arbitrary boolean expression, so that \(U\) and \(B\) is an instance of Sat. We will extend the variable set set \(U\) to a set \(U^{\prime}\) and replace each clause \(C i\) in \(B\) by a boolean expression \(B i\), such that
(a) \(B i\) is a product of clauses that use exactly three variables of \(U^{\prime}\).
(b) \(B i\) is satisfiable if and only if \(C i\) is.

Then \(B^{\prime}=B 1 B 2 \ldots B m\) will be an instance of 3 -Sat that is satisfiable over \(U^{\prime}\) if and only if \(B\) is satisfiable over \(U\). Let \(C i=(x 1+x 2+x 3+\ldots x k)\). There are three cases.
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Case 1: \(k=1\).
In this case we introduce new variables \(y i\) and \(z i\) and replace \(C i\) with
\[
B_{i}=\left(x_{1}+y_{i}+z_{i}\right)\left(x_{1}+y_{i}+\bar{z}_{i}\right)\left(x_{1}+\bar{y}_{i}+z_{i}\right)\left(x_{1}+\bar{y}_{i}+\bar{z}_{i}\right) .
\]

Case 2: \(k=2\).
In this case we introduce a new variable \(y i\) and replace Ci with
\[
B_{i}=\left(x_{1}+x_{2}+y_{i}\right)\left(x_{1}+x_{2}+\bar{y}_{i}\right) .
\]

Case 3: \(k=3\).
In this case we replace Ci with \(\mathrm{Bi}=\mathrm{Ci}\). Thus we make no change.
Case 4: \(k>3\).
In this case we introduce new variables \(y_{i_{1}}, y_{i_{2}}, \ldots, y_{i_{k-3}}\) and replace Ci with
\[
B_{i}=\left(x_{1}+x_{2}+y_{i_{1}}\right)\left(\bar{y}_{i_{1}}+x_{3}+y_{i_{2}}\right)\left(\bar{y}_{i_{2}}+x_{4}+y_{i_{3}}\right) \cdots\left(\bar{y}_{i_{k-3}}+x_{k-1}+x_{k}\right)
\]

It is routine to verify for each of Cases \(1,2,3\), and 4 , that Bi satisfies (a) and (b), see Exercise 9.5.2. We still must show that
\[
B^{\prime}=B 1 B 2 B 3 \ldots B m
\]
can be constructed with a polynomial time algorithm. If \(C i=(x 1+x 2+x 3+\ldots+x k)\) then \(B i\) contains at most \(4 k\) clauses of three variables and at most \(k+1\) new variables were introduced. Because \(k \leq n\), we conclude that to construct \(B^{\prime}\), at most \(4 m n\) new clauses of three variables are needed, and at most ( \(n+1\) ) \(m\) new variables are introduced. 'Both are polynomial in the size of the instance of Sat. Consequently we can construct \(B\) ' in
polynomial time.
Given a graph \(G\), a \(k\)-element subset \(K \subseteq V(G)\) of vertices is a called a vertex cover of size \(k\) if each edge of \(G\) has at least one end in \(K\). The Vertex Cover decision problem is:

Problem 9.4: Vertex Cover
Instance: a graph \(G\) and positive integer \(k\).
Question: does \(G\) have a vertex cover of size at most \(k\) ?
THEOREM 9.6 Vertex Cover is NP-complete.
PROOF It is easy to see that Vertex Cover is in NP, for if \(K\) is a purported vertex cover of the graph \(G\) of size \(k\), then we simply check each edge of \(G\) to see
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that at least one endpoint is in \(K\). There are \(\varepsilon\) edges to check so this takes time \(O(\varepsilon)\) and we can check in time \(|K| \leq n\) whether or not \(|K| \leq k\).
We now give a polynomial transformation from 3-Sat to Vertex Cover. Let \(B=C 1 C 2 . . . C m\) be a boolean expression over \(U=\{u 1, u 2, \ldots, u n\}\) in which each clause is a sum of exactly three variables. Thus for \(i=1,2, \ldots, n\), \(C i=(x i+y i+z i)\) for some \(z_{i} \in U \cup \bar{U}\), where \(\bar{U}=\{\bar{u} 1, \bar{u} 2, \ldots, \bar{u} n\}\). We construct a graph \(G\) on the vertex set
\[
V=U \cup \bar{U} \cup W,
\]
where \(W=\cup_{i=1}^{m}\left\{a_{i}, b_{i}, c_{i}\right\}\). The edge set of \(G\) is the union of the edges of \(m\) subgraphs \(H i, i=1,2, \ldots, m\), where \(H i\) is the subgraph shown in Figure 9.17. It consists of a triangle (ai,bi,ci), edges from ai, bi, ci to the variables contained in the clause, and edges connecting the variables to their complements. \(G\) has \(2 n+3 m\) vertices and \(n+6 m\) edges and hence can be built in polynomial time. Choose \(k=n+2 m\) to obtain an instance of the Vertex Cover problem for the graph \(G\) constructed.


\section*{FI GURE 9.17}

Subgraph Hicorresponding to clause ( \(x i+y i+z i\) )
We show that \(B\) has a satisfying truth assignment if and only if \(G\) has a vertex cover \(K\) of size \(k=n+2 m\). If \(t\) is a truth assignment, such that
\[
t(B)=\text { true, }
\]
then \(t\) must assign at least one variable \(x i, y i\) or \(z i\) to be true in clause Ci. Assume it is \(x i\). As \(x i\) is adjacent to exactly one vertex, ai, in the triangle \{ai,bi,ci\}, it follows that \(\{x i, b i, c i\}\) is a vertex cover of Hi , and hence
\[
K=\cup_{i=1}^{m}\left\{x_{i}, b_{i}, c_{i}\right\}
\]
is a vertex cover of size \(k\) for \(G\). An example is given in Figure 9.18.


FI GURE 9.18
Graph \(G\) corresponding to the boolean expression

\section*{\(B=(u \mathbf{2}+\bar{u} \mathbf{1}+n \mathbf{4})(u \mathbf{2}+u \mathbf{6}+\bar{u} 5)(u \mathbf{1}+u \mathbf{2}+u \mathbf{3})(\bar{u} \mathbf{2}+\bar{u} \mathbf{6}+\bar{u} 3)\). A vertex cover is}
\(K=\{u 4, a 1, b 1, u 6, a 2, c 2, u 3, a 3, b 3, \overline{u 2}, b 4, c 4 i, u 1, u 5\}\) and \(u 3=u 4=u 6=\) true, \(u 2=\) false, \(u \mathbf{1}, u 5\), assigned arbitrarily is a truth assignment satisfying \(B\).
Conversely suppose that \(K\) is a vertex cover of size \(k=n+2 m\) of \(G\). Then \(K\) must include at least one end of each of the \(n\) edges \(\{u i, \overline{u i}\}, i=1,2, \ldots, n\), accounting for at least \(n\) vertices in \(K\). Also \(K\) must cover the edges of each triangle (aj, bj, cj), and thus must contain at least two of \(\{a j, b j, c j\}\), for each \(j=1,2, \ldots, m\). This accounts for \(2 m\) more vertices, for a total of \(n+2 m=k\) vertices. Hence \(K\) must contain exactly one of the endpoints of each edge \(\{u i, \bar{u} i\}\), for \(i=1,2, \ldots, n\), and exactly two of \(a j, b j, c j\), for each \(j=1,2, \ldots, m\), corresponding to clause \(C j\). For each clause \(C j\), there is exactly one vertex aj, bj or cj of the triangle which is not in \(K\). Call it \(d j\).
Choose the unique variable of \(U \cup \bar{U}\) adjacent to di, and assign it true. Then at least one variable in each clause \(C j\) has been assigned the value true, giving a truth assignment that satisfies \(B\). Any remaining unassigned variables in \(U\) can be assigned true or false arbitrarily.
THEOREM 9.7 HamCycle is NP-complete.
PROOF Let \(G\) be a graph. Given an ordering \(u 1, u 2, \ldots\), , un of vertices of \(G\) we can check whether ( \(u 1, u 2, u 3\), ..., Un) is a hamilton cycle in polynomial time. Thus HamCycle is in NP. To show that HamCycle is NP-complete we transform from Vertex Cover.
Let \(G\) and \(k\) be an instance of Vertex Cover, where \(k\) is a positive integer. We will construct a graph \(G^{\prime}\) such that \(G^{\prime}\) has a hamilton cycle if and only if \(G\) has a vertex cover \(K=\{x 1, x 2, \ldots, x k\}\) of size \(k\). The graph \(G^{\prime}\) will have \(k+12 m\)
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vertices \(V\left(G^{\prime}\right)=K \cup\{(u, e, i): u \in V(G)\) is incident to \(e \in E(G)\) and \(i=1,2, \ldots, 6\}\), where \(m=|E(G)|\). The edges of \(G^{\prime}\) are of three types.
Type 1 edges of \(G^{\prime}\)
The type 1 edges are the \(14 m\) edges among the subgraphs \(H e, e \in E(G)\). We display \(H e\), where \(e=u u\) in Fiqure 9.19.


\section*{FIGURE 9.19}

The subgraph He , where \(e=u u\).

\section*{Type 2 edges of \(\mathbf{G}^{\prime}\)}

For each vertex \(u\) of \(G\) choose a fixed but arbitrary ordering \(e_{v_{1}}, e_{v_{2}}, \ldots, e_{v_{3}}\) of the \(d=D E G(u)\) edges incident to \(u\). The type 2 edges of \(G^{\prime}\) corresponding to \(u\) are:
\[
\left\{\left\{\left(v, e_{v_{i}}, 6\right),\left(v, e_{v_{i+1}}, 1\right)\right\}: i=1,2, \ldots, d-1\right\}
\]

Type 3 edges of \(G^{\prime}\)
The type 3 edges of \(G^{\prime}\) are:
\[
\left\{\left\{x_{i},\left(v, e_{v_{1}}, j\right)\right\}: x_{i} \in K, v \in V(G), j \in\{1,6\}\right\}
\]

The subgraph of \(G^{\prime}\) corresponding to the edges incident to a vertex \(u\) in \(G\) is illustrated in Figure 9.20. Before proving that \(G\) has a vertex cover of size \(k\) if and only if \(G^{\prime}\) has a hamilton cycle \(C=u 1, u 2, \ldots, u n\), we make five observations.
1. C must enter and exit the subgraph \(H e, e=u u\) from the four corners ( \(u, e, 1\) ), ( \(u, e, 6),(u, e, 1),(u, e, 6)\).
2. If \(C\) enters He at ( \(u, e, 1\) ), it must exit at ( \(u, e, 6\) ) and either pass through all the vertices of He or only those vertices with first coordinate \(u\). (In the first case as we shall see, \(u\) will be in the vertex cover of \(G\), and in the latter case both \(u\) and \(u\) will be in the vertex cover of \(G\).)
3. If \(C\) enters He at ( \(u, e, 1\) ), it must exit at ( \(u, e, 6\) ) and either pass through all the vertices of He or only those vertices with first coordinate \(u\). (In the first case as we shall see, \(u\) will be in the vertex cover of \(G\), and in the latter case both \(u\) and \(u\) will be in the vertex cover of \(G\).)

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FI GURE 9.20

\section*{Subgraph of \(G^{\prime}\) corresponding the the edges incident to \(u\) in \(G\)}
4. The vertices \(\{x 1, x 2, \ldots, x k\}\) divide \(C\) into paths. Thus we may assume, relabeling the vertices \(x 1, x 2, \ldots, x k\) if necessary, that \(C=P 1 P 2 \ldots P k\) where \(P i\) is an \(x i\) to \(x i+1\) path, where \(x k+1=x k\).
5. Let \(u i\) be such that \(x i\) is adjacent to ( \(u i, e, j\) ) in Pi where \(j=1\) or 6 . Then Pi contains every vertex ( \(u i, e^{\prime}, h\) ) where e is incident to \(u\).
We claim that the \(k\) vertices \(u 1, u 2, \ldots, u k\) selected in observation 5 are a vertex cover of \(G\). This is because the hamilton cycle C must contain all vertices of each of the subgraphs He for each \(e \in G\); and when \(H e\) is traversed by \(C\), it is traversed by some Pi in \(C\) and that \(P i\) selects an endpoint vi of \(e\).
Conversely, suppose \(K=\left\{v_{1}, v_{2}, \ldots, v_{k}\right\} \subseteq V(G)\) is a vertex cover of \(G\), of size \(k\). To construct a hamilton cycle \(C\) of \(G^{\prime}\), choose for each edge \(e \in E(G)\) the edges of He specified in Figure 9.21 (a), (b), or (c) depending on whether \(\{u, v\} \cup K\) equals \(\{u\},\{u, u\}\), or \(\{u\}\), respectively. (One of these must occur, because \(K\) is a vertex cover.) Also include the edges
\[
\left\{\left(v_{i}, e_{v_{i}}, 6\right),\left(v_{i}, e_{v_{i}}, 1\right)\right\}, i=1,2, \ldots, k
\]
the edges
and the edges
\[
\left\{x_{i},\left(v_{i}, e_{v_{i}}, 1\right)\right\}, i=1,2, \ldots, k
\]
\[
\left\{x_{i+1},\left(v_{i}, e_{v_{i}}, 1\right)\right\}, i=1,2, \ldots, k \text {, where } v_{k+1}=v_{1} .
\]

It is an easy exercise to verify that the included edges form a hamilton cycle in \(G^{\prime}\); see Exercise 9.5.5.
(a)

(b)

(c)


FI GURE 9.21
The three possible ways that a Hamilton cycle can traverse the subgraph He, corresponding to the cases for \(e=\{u, u\}\) in which (a) \(e \cap K=\{u\}\), (b) \(e \cap K=\{u, u\}\), and (c) \(e \cap K=\{u\}\). Exercises
9.5.1 Consider the boolean expression
\[
B=\left(x_{1}+\bar{x}_{2}+x_{3}+\bar{x}_{6}\right)\left(x_{2}+\bar{x}_{4}\right)\left(x_{5}\right)\left(x_{2}+\bar{x}_{4}+x_{5}\right)
\]

Find a boolean expression equivalent to \(B\) in which each clause uses only three variables.
9.5.2 Show for each Case 1, 2, 3, and 4 in Theorem 9.5 that the pair Bi, Ci satisfies
(a) \(B i\) is a product of clauses that use at most three variables in \(U^{\prime}\).
(b) Bi is satisfiable if and only if Ci is.
9.5.3 Consider the boolean expression
\[
\begin{aligned}
& B=(x+y+z)(x+y+\bar{z})(w+\bar{x}+z)(w+\bar{x}+\bar{z})(\bar{w}+\bar{x}+z) \\
& \quad(\bar{w}+\bar{x}+\bar{z})(w+\bar{y}+z)(w+\bar{y}+\bar{z})(\bar{w}+\bar{y}+z)(\bar{w}+\bar{y}+\bar{z})
\end{aligned}
\]
(a) Show that there is no truth assignment that satisfies \(B\).
(b) Construct the graph \(G\) in Theorem 9.6 that corresponds to \(B\).
(c) Show that \(G\) does not have a vertex cover of size 25.
9.5.4 Verify the five observations in Theorem 9.7.
9.5.5 Verify that the included edges in the converse part of Theorem 9.7, do indeed form a hamilton cycle. page_213

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\subsection*{9.6 The traveling salesman problem}

The traveling salesman problem (TSP) is very closely related to the HamCycle problem. A salesman is to visit \(n\) cities \(u 1, u 2, \ldots, u n\). The cost of traveling from ui to \(u j\) is \(W(u i u j)\). Find the cheapest tour which brings him back to his starting point. Figure 9.22 shows an instance of the TSP problem. It is a complete graph Kn with positive integral weights on the edges. The problem asks for a hamilton cycle of minimum cost.


\section*{FI GURE 9.22}

\section*{An instance of the TSP problem.}

It is easy to show that the HamCycle problem can be reduced to the TSP problem. In order to do this, we first must phrase it as a decision problem.

Problem 9.5: TSP Decision
Instance: a weighted complete graph \(K n\), and an integer \(M\),
Question: does \(K n\) have a hamilton cycle of cost \(\leq M\) ?
We can then find the actual minimum by doing a binary search on the range of values \(n \leq M \leq n W m a x\), where Wmax is the maximum edge-weight. Suppose that we had an efficient algorithm for the TSP Decision problem. Let \(G\) be any graph on \(n\) vertices which we want to test for hamiltonicity. Embed \(G\) in a compete graph \(K n\), giving the edges of \(G\) weight 1 , and the edges of \(\bar{G}\) weight 2.
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Now ask whether \(G\) has a TSP tour of cost \(\leq n\). If the answer is yes, then \(G\) is hamiltonian. Otherwise \(G\) is non-hamiltonian.
Since HamCycle is NP-complete, we conclude that the TSP Decision problem is at least as hard as an NPcomplete problem. In a certain sense, it is harder than the NP-complete problems, since the edge weights W(uiuj) are not bounded in size. So it may take many steps just to add two of the weights. However, if we limit the size of the weights to the range of numbers available on a computer with a fixed word length, then the TSP Decision problem is also NP-complete. It is easy to see that TSP Decision \(\in\) NP, since we can write down the sequence of vertices on a cycle \(C\) of cost \(\leq M\) and verify it in \(n\) steps.
One way to approximate a solution is similar to the crossover technique. Choose a hamilton cycle C in Kn arbitrarily. For each edge \(u v \in C\), search for an edge \(w x \in C\) such that \(W(u u)+W(w x)>W(u w)+W(u x)\). If such an edge exists, re-route \(C\) as shown in Figure 9.23. Repeat until no improvement can be made. Do this for several randomly chosen starting cycles, and take the best as an approximation to the optimum.


FI GURE 9.23

\section*{Re-routing a TSP tour}

The cycle \(Q=(u, u, x, w)\) is similar to a crossover. In general, if \(Q\) is any cycle such that \(C \oplus Q\) is a hamilton cycle, and \(W(C \cap Q)>W(Q-C)\), then \(C \oplus Q\) will be a TSP tour of smaller cost than \(C\). We can search for crossovers \(Q\) containing up to \(M\) edges, for some fixed value \(M\), and this will provide a tour which may be
close to the optimum. How close does it come to the optimum?
It is possible to obtain a rough estimate of how good a tour \(C\) is, by using a minimum spanning tree algorithm. Let \(C^{*}\) be an optimum TSP tour. For any vertex \(u, C^{*}-u\) is a spanning tree of \(K n-u\). Let Tu be a minimum spanning tree of \(K n-u\). Then \(W\left(C^{*}-u\right) W(T u)\). Given the path \(C^{*}-u\), we must add back two edges incident on \(u\) to get \(C^{*}\). If we add two edges incident on \(u\) to \(T u\), of minimum possible weight, we will get a graph \(T_{v}^{*}\), such that \(W\left(C^{*}\right) \geq W\left(T_{v}^{*}\right)\). For example, Figure 9.24 shows a minimum spanning tree \(T 3\), of \(K n-3\) for the
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instance of TSP shown in Fiqure 9.22.


\section*{FI GURE 9.24}

\section*{A minimum spanning tree 73 , plus two edges}

The two edges incident on vertex 3 that we add to \(T 3\) have weights 10 and 11 in this case. We thus obtain a bound \(W\left(C^{*}\right) \geq W\left(T_{3}^{*}\right)=42\). We do this for each vertex \(u\), and choose the maximum of the bounds obtained. This is called the spanning tree bound for the TSP:
\[
W\left(C^{*}\right) \geq \operatorname{Max}_{v} W\left(T_{v}^{*}\right)
\]

\section*{Exercises}
9.6.1 Work out the spanning tree bound for the TSP instance of Figure 9.22.
9.6.2 Find a TSP tour \(C\) in the graph of Figure 9.22 by re-routing any starting cycle, until no more improvement is possible. Compare the weight of \(C\) with the result of Exercise 9.6.1.
9.6.3 Construct all possible re-routing patterns (crossovers) containing three or four edges of \(C\).
9.7 The \(\triangle\) TSP

Distances measured on the earth satisfy the triangle inequality, namely, for any three points \(X, Y\), and \(Z\), DIST \((X, Y)+\operatorname{DIST}(Y, Z) \geq\) DIST \((X, Z)\). The triangle traveling salesman problem, denoted \(\Delta\) TSP, refers to instances of the TSP
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satisfying this inequality. When the triangle inequality is known to hold, additional methods are possible.
THEOREM 9.8 Let Knbe an instance of the \(\triangle T S P\), and let \(G\) be any Eulerian spanning subgraph of Kn. If C* is an optimum TSP tour, then \(W\left(C^{*}\right) \leq W(G)\).
PROOF Consider an Euler tour \(H\) in \(G\) starting at any vertex. The sequence of vertices traversed by \(H\) is \(v_{i_{0}}, v_{i_{1}}, v_{i_{2}}, v_{i_{3}}, \cdots\) If \(G\) is a cycle, then \(H\) is a hamilton cycle, so that \(W\left(C^{*}\right) \leq W(G)\), and we are done. Otherwise, \(H\) repeats one or more vertices. Construct a cycle \(C\) from \(H\) by taking the vertices in the order that they appear in \(H\), simply ignoring repeated vertices. Since \(G\) is a spanning subgraph of \(K n\), all vertices will be included in \(C\). For example, if \(G\) is


FI GURE 9.25

\section*{An Eulerian graph \(G\) and TSP tour \(C\)}
the graph of Figure 9.25 , and \(H\) is the Euler tour ( \(1,2,3,4,6,1,3,6,5,4\) ), then the cycle \(C\) obtained is \((1,2,3,4,6,5)\). Because of the triangle inequality, it will turn out that \(W(C) \leq W(G)\). Let the cycle obtained be \(C=(u 1, u 2, \ldots, u n)\), and suppose that the Euler tour \(H\) contains one or more vertices between \(u k\) and \(u k+1\). Without loss of generality, suppose that there are just three vertices \(x, y, z\) between \(u k\) and \(u k+1\). See Figure 9.26. Then because of the triangle inequality, we can write
\[
\begin{aligned}
& W(u k x)+W(x y) \geq W(u k y), \\
& W(u k y)+W(y z) \geq W(u k z),
\end{aligned}
\]
and
Thus
\[
W(u k z)+W(z u k s+1) \geq W(u k u k+1)
\]
\[
\underset{\text { page_217 }}{W(u k u k+1) \leq W(u k x)+W(x y)}+W(z u k+1) .
\]

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\(u_{k}\)
FI GURE 9.26
Applying the triangle inequality
The left side of the inequality contributes to \(W(C)\). The right side contributes to \(W(H)\). It follows that \(W(C) \leq W(G)\), for any Eulerian \(G\).
Notice that the particular cycle \(C\) obtained from \(G\) depends on the Euler tour \(H\) chosen, so that the graph \(G\) will give rise to a number of different hamilton cycles \(C\). In particular, we could construct \(G\) from a minimum spanning tree \(T\), by simply doubling each edge. This gives an Eulerian multigraph \(G\). The method used in the theorem will also work with multigraphs, so we conclude that \(W\left(C^{*}\right) \leq 2 W(T)\). This is called the tree algorithm for the TSP.
LEMMA 9.9 The tree algorithm produces a cycle of cost at most twice the optimum.
PROOF Let \(C\) be the cycle obtained by the tree algorithm, let \(C^{*}\) be an optimum cycle, and let \(T\) be a minimum spanning tree of the instance for \(\triangle T S P\). Since \(C^{*}\) is a spanning subgraph of \(K n\), we conclude that \(W\left(C^{*}\right)>W(T)\). But we know that \(W(C) \leq 2 W(T)<2 W\left(C^{*}\right)\).

\subsection*{9.8 Christofides' algorithm}

Christofides found a way to construct an Eulerian subgraph of smaller weight than \(2 W(T)\). Let \(K n\) be an instance of the \(\Delta \mathrm{TSP}\), and let \(T\) be a minimum spanning tree. Let \(X \subseteq V\left(K_{n}\right)\) be the vertices of \(T\) of odd degree. \(X\) contains an even number of vertices. The subgraph of \(K n\) induced by \(X\) is a complete subgraph. Let \(M\) be a perfect matching in \(X\) of minimum weight. For example, Figure 9.27 shows a minimum spanning tree for the graph of Figure 9.22, together with a
minimum-weight matching \(M\), shown as dashed lines.


\section*{FI GURE 9.27}

\section*{Christofides' algorithm}

This gives a graph \(G=T+M\) which is Eulerian. It is quite possible that \(G\) is a multigraph. We now find an Euler tour in \(G\) and use it to construct a TSP tour \(C\) of cost at most \(W(T)+W(M)\). This is called Christofides' algorithm.
THEOREM 9.10 Let C be the TSP tour produced by Christofides' algorithm and let \(C^{*}\) be an optimum tour. Then
\[
W(C) \leq \frac{3}{2} W\left(C^{*}\right)
\]

PROOF Let \(u 1, U 2, \ldots, u 2 k\) be the vertices of odd degree, and suppose that they appear on \(C^{*}\) in that order. This defines two matchings,
and
\[
M 1=\{u 1 u 2, u 3 u 4, \ldots\}
\]
\[
M 2=\{u 2 u 3, u 4 u 5, \ldots, u 2 k u 1\} .
\]

See Figure 9.28. If \(M\) is the minimum weight matching, we conclude that \(W(M 1), W(M 2) \geq W(M)\). The portion of \(C^{*}\) between \(u i\) and \(u i+1\) satisfies
\[
\begin{gathered}
W\left(C^{*}[u i, u i+1]\right) \geq W(\text { uiui }+1), \\
W\left(C^{*}\right) \geq W(M 1+W(M 2) \geq 2 W(M), \\
\text { page_219 }
\end{gathered}
\]

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or
\[
W(M) \leq \frac{1}{2} W\left(C^{*}\right) .
\]

The cycle \(C\) found by Christofides's algorithm satisfies
\[
W(C) \leq W(T)+W(M)<W\left(C^{*}\right)+\frac{1}{2} W\left(C^{*}\right)
\]
because \(W(T)<W\left(C^{*}\right)\). It follows that
\[
W(C) \leq \frac{3}{2} W\left(C^{*}\right)
\]


FI GURE 9.28
Two matchings M1 and M2
Thus, Christofides' algorithm always comes within \(50 \%\) of the optimum.

\section*{Exercises}
9.8.1 Use the tree algorithm to find a TSP tour for the graph of Figure 9.22.
9.8.2 Solve the same TSP instance using Christofides' algorithm. Compare the values found for \(W(C), W(T)\), and \(W(T+M)\).
9.8.3 Solve the \(\Delta\) TSP instance of Figure 9.29, using Christofides' algorithm. Compute the spanning tree bound as well.

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FI GURE 9.29

\section*{An instance of \(\Delta\) TSP}

\subsection*{9.9 Notes}

An excellent survey of hamiltonian graphs appears in BERMOND [15]. The hamilton closure and the BondyChvátal theorem are from BONDY and MURTY [19]. The extended multi-path algorithm is from KOCAY [77]. A classic book on the theory of NP-completeness is the text by GAREY and JOHNSON [50]. A very readable proof of Cook's theorem, that Satisfiability is NP-complete, appears in PAPADIMITRI OU and STEI GLITZ [94], which also contains an excellent section on Christofides' algorithm. The book by CHRISTOFIDES [26] has an extended chapter on the traveling salesman problem. The book LAWLER, LENSTRA, RINNOOY KAN and SHMOYS [83] is a collection of articles on the traveling salesman problem.
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\subsection*{10.1 I ntroduction}

Directed graphs have already been introduced in the Chapter 8 . If \(G\) is a digraph and \(u, v \in V(G)\), we write \(u \longrightarrow v\)
to indicate that the edge \(u v\) is directed from \(u\) to \(u\). The in-edges at \(u\) are the edges of the form \((u, u)\). The in-degree of \(u\) is \(d\)-(u), the number of in-edges. Similarly the out-edges at \(u\) are all edges of the form \((u, u)\) and the out-degree \(d+(u)\) is the number of out-edges at \(u\). The degree of \(u\) is
\[
\mathrm{DEG}(u)=d+(u)+d-(u) .
\]

Given any undirected graph G, we can assign a direction to each of its edges, giving a digraph called an oriented graph. A digraph is simple if it is an orientation of a simple graph. A digraph is strict if it has no loops, and no two directed edges have the same endpoints. A strict digraph can have edges \((u, u)\) and \((u, u)\), whereas an oriented graph cannot.
Digraphs have extremely wide application, for the social sciences, economics, business management,
operations research, operating systems, compiler design, scheduling problems, combinatorial problems, solving systems of linear equations, and many other areas. We shall describe only a few fundamental concepts in this chapter.

\subsection*{10.2 Activity graphs, critical paths}

Suppose that a large project is broken down into smaller tasks. For example, building a house can be subdivided into many smaller tasks: dig the basement, install the sewer pipes, water pipes, electricity, pour the basement concrete, build

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the frame, floor, roof, cover the roof and walls, install the wiring, plumbing, heat-ing, finish the walls, etc. Some of these tasks must be done in a certain order-the basement must be dug before the concrete can be poured, the wiring must be installed before the walls can be finished, etc. Other tasks can take place at the same time, (e.g., the wiring and plumbing can be installed simultaneously). We can construct a directed graph, called an activity graph, to represent such projects. It has a starting node s, where the project begins, and a completion node \(t\), where it is finished. The subtasks are represented by directed edges. The nodes represent the beginning and end of tasks (the synchronization points between tasks). Figure 10.1 shows an example of an activity graph. Each task takes a certain estimated time to complete, and this is represented by assigning each edge uu a weight WT(uu), being the amount of time required for that task.
What is the minimum amount of time required for the entire project? It will be the length of the longest directed path from start to completion. Any longest directed path from \(s\) to \(t\) is called a critical path. Figure 10.1 shows a critical path in an activity graph.


\section*{FI GURE 10.1}

\section*{An activity graph}

Notice that an activity graph must have no directed cycles. For if a directed cycle existed, it would be impossible to complete the project according to the constraints. Thus, activity graphs are acyclic digraphs.
Activity graphs are applicable to any large project, such as building construction, business projects, or factory assembly lines.
The critical path method (CPM) is a technique for analyzing a project acccording to the longest paths in its activity graph. In order to find a longest path from s
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to \(t\), we proceed very much as in Dijkstra's algorithm (Chapter 2 ), which builds a spanning tree, rooted at s , of shortest paths. To find longest paths instead, the algorithm builds an out-directed spanning tree, rooted at \(s\), of longest directed paths from \(s\) to each vertex \(u\). We store a value \(T[u]\) for each \(u\), being the earliest time
at which tasks starting from \(u\) can begin. \(T[u]\) is the length of a longest su-path. When the algorithm completes, the critical path is the unique path in the spanning tree to vertex \(t\). Notice that in Figure 10.1, the edge from vertex 1 to 3 has length 6 , but that the path \((1,7,3)\) has the longer length of 7 . In order for the algorithm to correctly choose the longer path, it must be sure to assign \(T(7)\) before \(T(3)\). Thus, the vertices must be taken in a certain order. For every edge \((u, u), T(u)\) must be computed before \(T(u)\).

\subsection*{10.3 Topological order}

A topological ordering of an acyclic digraph \(G\) is a permutation \(o\) of
\[
V(G)=\{1,2, \ldots, n\}
\]
such that \(o(u)<o(U)\) whenever \(u \longrightarrow v\). Thus all edges are directed from smaller to higher vertex numbers. Notice that only acyclic digraphs have topological orderings, since a directed cycle cannot be ordered in this way. Topological orderings are easy to find. We present both a breadth-first and a depth-first algorithm. Algorithm 10.3.1 is the the breadth-first algorithm. The topological order is built on the queue. Algorithm 10.3.1 begins by placing all vertices with in-degree 0 on the queue. These are first in the topological order. InDegree[u] is then adjusted so that it counts the in-degree of \(u\) only from vertices not yet on the queue. This is done by decrementing InDegree[u] according to its in-edges from the queue. When InDegree[u]=0,u has no more in-edges, so it too, is added to the queue. When all \(n\) vertices are on the queue, the vertices are in topological order. Notice that if \(G\) has a directed cycle, none of the vertices of the cycle will ever be placed on the queue. In that case, the algorithm will terminate with fewer than \(n\) vertices on the queue. This is easy to detect. Computing InDegree[u] takes \(\sum_{v} d^{-}(v)=\varepsilon_{\text {steps. Each vertex is placed on the queue exactly }}\) once, and its \(d+(u)\) out-edges are taken in turn, taking \(\sum_{u} d^{+}(u)=\varepsilon\) steps. Thus the complexity of the algorithm is \(O(n+\varepsilon)\).
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comment: \(\begin{aligned} & \text { Algorithm 10.3.1: BFTOPSORT(G, } n \text { ) } \\ & \left\{\begin{array}{l}\text { Breadth-first topological sort of } G \\ \text { InDegree }[v] \text { is the in-degree of vertex } v, \text { an array } \\ S \text { canQ }[k] \text { is the } k^{\text {th }} \text { vertex on a queue, an array } \\ \text { Qsize is the number of points on Scan } Q\end{array}\right.\end{aligned}\)
Qsize \(\leftarrow 0\)
\(k \leftarrow 1\)
for \(U \leftarrow 1\) to \(n\)
do \(\left\{\begin{array}{l}\text { compute InDegree }[v] \\ \text { if InDegree }[v]=0\end{array}\right.\)
then \(\left\{\begin{array}{l}\text { Qsize } \leftarrow \text { Qsize }+1 \\ \text { Scan }[Q \text { size }] \leftarrow v\end{array}\right.\)
while \(k \leq\) Qsize
do \(\left\{\begin{array}{l}u \leftarrow \operatorname{ScanQ[k]} \\ \text { for each } v \text { such that } u \longrightarrow v \\ \text { do }\left\{\begin{array}{l}\text { InDegree }[v] \leftarrow \text { InDegree }[v]-1 \\ \text { if InDegree }[v]=0\end{array}\right. \\ \text { then }\left\{\begin{array}{l}\text { Qsize } \leftarrow Q \operatorname{size}+1 \\ \text { ScanQ[Qsize }] \leftarrow v \\ \text { if } Q \operatorname{size}=n \\ \text { then go to } 1\end{array}\right. \\ k \leftarrow k+1\end{array}\right.\)
if Qsizie<n then \(G\) contains a directed cycle
Algorithm 10.3.2 is the depth-first topological sort algorithm and is easier to program, but somewhat subtler. It calls the recursive Procedure DFS(), and we assume that Procedure DFS() has access to the variables of Algorithm DFTOPSORT() as globals. When Procedure DFS(u) is called from DFTOPSORT(), it builds a rooted tree, directed outward from the root u. DFNum[u] gives the order in which the vertices are visited. The depth-first search does a traversal of this tree, using a recursive call to visit all descendants of \(u\) before \(u\) itself is assigned a number NUM[u], its rank in the topological order. Thus, if \(G\) is acyclic, all vertices that can be reached on directed paths out of \(u\) will be
ranked before \(u\) itself is ranked. Thus, for every edge \((u, u)\), the numbering will satisfy \(N U M[u]<N U M[u]\). The first vertex numbered is assigned a Rank of \(n\). The variable Rank is then decremented. So the vertices are numbered 1 to \(n\), in topological order. It is obvious that the complexity of the algorithm is \(O(n+\varepsilon)\).
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Algorithm 10.3.2: \(\operatorname{DFTOPSORT}(G, n)\)
Depth-first topological sort of \(G\)
comment: \(D F N u m[v]\) is the DF-numbering assigned to the vertex \(v\) \(N U M[v]\) is the topological numbering of the vertex \(v\) DFCount, Rank are counters
procedure DFS(U)
comment: extend the depth-first search to vertex \(u\)
DFCount \(\leftarrow\) DFCount+1
DFNum[U] \(\leftarrow\) DFCount
for each \(w\) such that \(v \longrightarrow w\)
do if \(D F N u m[w]=0\) then DFS(w)
NUM[U] \(\leftarrow\) Rank
Rank \(\leftarrow\) Rank-1
main
for \(u \leftarrow 1\) to \(n\)
do
\(\left\{\begin{array}{l}N U M[u] \leftarrow 0\end{array}\right.\)
\(\{\operatorname{DFNum}[u] \leftarrow 0\)
DFCount \(\leftarrow 0\)
Rank \(\leftarrow n\)
for \(u \leftarrow 1\) to \(n\)
do if \(\operatorname{DFNum}[u]=0\) then DFS(u)
The depth-first topological sort does not provide the vertices on a queue in sorted order. Instead it assigns a number to each vertex giving its rank in the topological order. If we need the vertices on a queue, as we likely will, we can construct one from the array NUM by executing a single loop.
for \(u \leftarrow 1\) to \(n\) do \(S c a n Q[N U M[U]] \leftarrow u\)
This works because NUM is a permutation of the numbers 1 to \(n\), and the loop computes the inverse of the permutation. Another method is to compute the inverse array during the DFS simultaneously with the NUM array.
What happens if the depth-first topological sort is given a digraph that is not acyclic? It will still produce a numbering, but it will not be a topological ordering. We will have more to say about this in Section 10.4. Notice that DFS(u) may be called several times from Algorithm 10.3.2. Each time it is called, a rooted tree directed outward from the root is constructed. With undirected graphs, a DFS

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constructs a single rooted spanning tree of \(G\) (see Chapter 6). For directed graphs, a single out-directed tree may not be enough to span all of G. A spanning forest of rooted, out-directed trees is constructed.
We return now to the critical path method. Let \(G\) be an activity graph with \(V(G)=\{1,2, \ldots, n\}\), and suppose that the vertices have been numbered in topological order; that is, \(u<u\) whenever \(u \longrightarrow v\). The start vertex is \(s=1\). We set \(T(1) \leftarrow 0\). We know that vertex 2 has an in-edge only from \(s, s o T(2)\) is assigned the cost of the edge ( 1,2 ). In general, \(u\) can have in-edges only from vertices \(1, \ldots, u-1\), and we can take \(T(u)\) to be
\[
T(v) \leftarrow \operatorname{Max}\{T(u)+\mathrm{WT}(u v): u \longrightarrow v\} .
\]

We also store an array PrevPt, where PrevPt[u] is the point previous to \(u\) on a directed su-path. If \(T(u)\) is computed to be \(T(u)+W T(u u)\) for some \(u\), we simultaneously assign PrevPt[u] \(\leftarrow u\). When the algorithm completes, we can find the critical path by executing \(w \leftarrow \operatorname{PrevPt}[w]\) until \(w=0\), starting with \(w=t(=n)\). The number of steps required to compute the longest paths once the topological sort has been completed is proportional to \(n+\sum_{v} d^{-}(v)=O(n+\varepsilon)\).
The minimum time required to complete the project is \(T(n)\). This can be achieved only if all tasks along the
critical path begin and end on time. These tasks are critical. There may be some slack elsewhere in the system, though, which can be used to advantage. The earliest time at which a node \(u\) in the activity graph can be reached is \(T(u)\), the length of the longest \(s u\)-path. We could also compute the latest time at which node \(u\) must be reached if the project is to finish on time. This is \(T(n)\) minus the length of the longest directed path from \(u\) to \(t\). Let \(T^{\prime}(u)\) be the length of the longest directed path from \(u\) to \(t\). We can compute this in the same way that \(T(U)\) is computed, but beginning with \(t\) instead of \(s\), and working backward. Thus for each node \(u\), we can find the two values \(T(u)\) and \(T(n)-T^{\prime}(u)\), being the earliest and latest times at which node \(u\) can be reached. This slack time can create some flexibility in project management.

\section*{Exercises}
10.3.1 Find a topological ordering of the activity graphs of Figures 10.1 and 10.2. Apply the critical path method to find the longest \(s u\)-paths and vt-paths, for each \(u\). Work out the earliest and latest times for each node \(u\).
10.3.2 Program the breadth-first and depth-first topological sort algorithms. Test them on the graph of Figure 10.1.
10.3.3 Consider the recursive procedure DFS(U) defined above, applied to a directed graph G. Suppose that \(D F S(u)\) has just been called, and that \(A(U)\) is the set of all vertices which are ancestors of \(u\) (the path from \(u\) to the root contains the ancestors of \(u\) ).

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\section*{FI GURE 10.2}

\section*{An activity graph}

Suppose that \(G-A(u)\) contains a directed path from \(u\) to \(w\). Prove that \(w\) will be visited before \(\operatorname{DFS}(u)\) returns. Use induction on the length of the path from \(u\) to \(w\).

\subsection*{10.4 Strong components}

A digraph \(G\) is connected if every pair of vertices \(u\) and \(u\) is connected by a path. This need not be a directed path. The digraph \(G\) is strongly connected if every pair of vertices is connected by a directed path. Thus, if \(G\) is strongly connected, G contains both a uu-path and a uu-path, for every \(u\) and \(u\). It follows that every vertex of \(G\) is contained in a directed cycle. A digraph which is strongly connected is said to be strong. Notice that a strong digraph does not have to be 2-connected. It may contain one or more cut-vertices. By default, the complete digraph \(K 1\) is strong, since it does not have a pair of vertices. If \(G\) is acyclic, then the only strong subgraphs of \(G\) are the individual nodes. But if \(G\) contains any directed cycle, then \(G\) will contain one or more non-trivial strongly connected subgraphs. A subgraph \(H\) is a strong component of \(G\) if it is a maximal strongly connected subgraph; that is, \(H\) is strong, and \(G\) has no larger subgraph containing \(H\) which is also strong. Figure 10.3 shows a digraph \(G\) with four strong components. The edges of the strong components are indicated by thicker lines. Two of the strong components are single vertices, which are shaded black.


\section*{FI GURE 10.3}

\section*{Strong components}

Notice that every vertex of \(G\) is contained in exactly one strong component, but that some edges of \(G\) need not be contained in any strong component. Exercise 10.4.1. shows that this definition of strong components is well-defined.
If G1, G2,..., Gm are the strong components of \(G\), we can construct a new digraph by contracting each strong component into a single vertex.
DEFINITION 10.1: Let \(G 1, G 2, \ldots, G m\) be the strong components of \(G\). The condensation of \(G\) is the digraph whose vertices are \(G 1, G 2, \ldots, G m\), and whose edges are all ordered pairs \((G i, G j)\) such that \(G\) has at least one edge directed from a vertex of \(G i\) to a vertex of \(G j\).
It is proved in Exercise 10.4.3 that the condensation is an acyclic digraph.

\section*{Exercises}
10.4.1 Suppose that \(H\) is a strong subgraph of \(G\) such that \(H\) is contained in two larger strong subgraphs: \(H \leq H 1\) and \(H \leq H 2\), where \(H 1\) and \(H 2\) are both strong. Show that \(H_{1} \cup H_{2}\) is strong. Conclude that the strong components of \(G\) are well-defined.
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10.4.2 Show that an edge \((u, u)\) is contained in a strong component if and only if \((u, u)\) is contained in a directed cycle.
10.4.3 Find the condensation of the digraph of Figure 10.3. Prove that the condensation of a digraph is always acyclic.
10.4.4 The converse of a digraph is obtained by reversing the direction of each edge. A digraph is selfconverse if it is isomorphic to its converse. Find all self-converse simple digraphs on one, two, three, and four vertices.
10.4.5 Show that the condensation of the converse is the converse of the condensation.
10.4.6 Let \(G\) be a self-converse simple digraph, and let \(G^{\prime}\) be the converse of \(G\). Let \(\theta\) be an isomorphism of \(G\) with \(G^{\prime}\), so that \(\theta\) is a permutation of \(V(G)=V\left(G^{\prime}\right)\). Prove that \(\theta\) has at most one cycle of odd length. Find the possible cycle structures of \(\theta\) when \(G\) has at most five vertices. Use this to find all the self-converse digraphs on five vertices.
In this section, we present an algorithm to find the strong components of a digraph \(G\). It is based on a depthfirst search. It is very similar to the DFS used to find the blocks of a graph in Chapter 6, and to the DFS used above to find a topological ordering in an acyclic digraph. When finding a topological ordering, we saw that in a digraph \(G\), Algorithm 10.3.2 constructs a spanning forest of outdirected, rooted trees. Each time DFS \((u)\) is called, a DF-tree rooted at \(u\) is built. The edges of \(G\) can be classified as either tree-edges or fronds. For example, a spanning forest for the graph of Figure 10.3 is shown in Figure 10.4 below. The fronds are shown as dashed edges. Not all the fronds are shown, as can be seen by comparing Figures 10.3 and 10.4. The numbering of the nodes is the DF-numbering.
Let the components of the spanning forest constructed by a depth-first search in a graph \(G\) be denoted T1, \(\mathrm{T} 2, \ldots, \mathrm{Tk}\), where the Ti were constructed in that order. Figure 10.4 has \(k=3\). Each Ti is an out-directed,
rooted tree. Notice that each strong component of \(G\) is contained within some Ti, and that each Ti may contain more than one strong component. Fronds can be directed from a tree Ti to a previous tree Tj, where \(j<i\), but not to a later tree, by nature of the depth-first search.
Given any vertex \(u, u\) is contained in some Ti. The set of ancestors of \(u\) is \(A(u)\), all vertices (except \(u\) )
contained in the path in Ti from \(u\) to the root of Ti. When DFS(u) is called, it will in turn call DFS \((w)\) for several vertices \(w\). The branch of Ti at \(u\) containing \(w\) is the sub-tree built by the recursive call DFS ( \(w\) ). For example, in Figure 10.4, there are two branches at vertex 4, constructed by the recursive calls DFS(5) and DFS(7). If \(x\) is any vertex for which \(v \in A(x)\), we write \(B u(x)\) for the branch at \(u\) containing \(x\). In Figure 10.4 above, we have \(B 4(7)=B 4(8)\) and \(B 4(5)=B 4(6)\).
LEMMA 10.1 Suppose that a depth-first search in \(G\) is visiting vertex \(u\), and that \(G-A(u)\) contains a directed path from \(u\) to \(w\). Then vertex \(w\) will be visited before the algorithm returns from visiting \(u\).

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FI GURE 10.4
A depth-first, rooted, spanning forest
PROOF Exercise 10.4.1.
This lemma allows us to classify the fronds of \(G\) with respect to a depth-first forest.
THEOREM 10.2 Let T1, T2,..., Tkbe the components of a depth-first spanning forest of \(G\), where the Ti were constructed in that order. Let \((x, y)\) be a frond, where \(x \in T_{i}\). Then there are three possibilities:
1. \(y \in T_{j}\), where \(j<i\).
2. \(y \in T_{i}\), and one of \(x\) and \(y\) is an ancestor of the other.
3. \(y \in T_{i}\), and \(x\) and \(y\) are in different branches of a common ancestor \(u\), where \(B u(y)\) was searched before \(B u(x)\).
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PROOF Let \((x, y)\) be a frond, where \(x \in T_{i}\). If \(y \in T_{j}\), where \(j \neq i\), then we must have \(j<i\), for otherwise the previous lemma tells us that \(y\) would be visited before DFS( \(x\) ) returns, so that \(x\) would be an ancestor of \(y\).

Otherwise \(x, y \in T_{i}\). If \(x\) is visited before \(y\), then the previous lemma again tells us that \(x\) would be an ancestor of \(y\). This gives the second case above. Otherwise \(y\) is visited before \(x\). If \(G\) contains a directed \(y x\) path, then we have \(y\) an ancestor of \(x\), again the second case above. Otherwise there is no directed \(y x\)-path. The paths in Ti from \(x\) and \(y\) to the root of Ti first meet in some vertex \(u\). Then \(B u(y)\) was searched before \(B u(x)\), giving Case 3.
We call a frond \((x, y)\) type 1, 2, or 3 according to whether it falls in Case 1, 2, or 3 in Theorem 10.2. Fronds of type 1 cannot be part of any directed cycle since there are no edges from Tj to Ti when \(j<i\). Therefore these fronds are not in any strong component. Consequently each strong component is contained within some Ti. A frond of type 2 creates a directed cycle, so that all edges of Ti on the path connecting \(x\) to \(y\) are in the same strong component. The low-point technique used to find blocks in Chapter 6 will work to find these cycles. A frond ( \(x, y\) ) of type 3 may or may not be part of a directed cycle. Consider the frond \((7,5)\) in Figure 10.4.
Vertices 7 and 5 are in different branches at vertex 4 . Since 4 is an ancestor of 7 , we have a directed path (4, \(7,5)\). If we were to compute low-points, we would know that the low-point of 5 is vertex 3, an ancestor of 4. This would imply the existence of a directed cycle containing \(3,4,7\), and 5 , namely, \((3,4,7,5,6)\). So we find that 7 is in the same strong component as 5 and that the low-point of 7 is also 3.
We can build the strong components of \(G\) on a stack. Define the low-point of a vertex \(u\) to be
LowPt[U]=the smallest DFNum[ \(w\) ],
where either \(w=u\) or \(w \in A(v)\) and \(G\) contains a directed path from \(u\) to \(w\). The main component of Algorithm 10.4.1 to compute the strong componets just initiallizes the variables and calls Proceedure DFS() to build each rooted tree of the spanning forest and to compute the low-points. We assume that Proceedure DFS() has access to the variables of the calling program as globals. The algorithm stores the vertices of each strong component on a stack, stored as an array. As before, we have the DFNum[•] and LowPt[•] arrays. We also store the Stack[•] as an array of vertices. OnStack [ \(u\) ] is true if \(u\) is on the stack. DFCount is a global counter. Top is a global variable giving the index of the current top of the stack.
The Procedure DFS() computes the low-points and builds the stack. The algorithm begins by stacking each vertex that it visits. The vertices on the stack will form the current strong component being constructed. LowPt[ \(u\) ] is initiallized to DFNum[u]. Each w such that \(w \longrightarrow v\) is taken in turn. The statements at point (1) extend the DFS from vertex \(u\) to \(w\). Upon returning from the recursive
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call, LowPt[ \(u\) ] is updated. Since \(v \longrightarrow w\) and \(G\) contains a directed path from \(w\) to LowPt[ \(w\) ], we update LowPt[u] if LowPt[w] is smaller.
The statements at point (2) are executed if \(u w\) is a frond. If DFNum[w]> DFNum[u], it means that \(u\) is an ancestor of \(w\). These fronds are ignored. Otherwise \(w\) was visited before \(u\). If \(w\) is the parent of \(u, u w\) is a tree edge rather than a frond, and is ignored. If \(w\) is no longer on the stack, it means that \(w\) is in a strong component previously constructed. The edge uw cannot be part of a strong component in that case, so it is also ignored. If each of these tests is passed, \(G\) contains a directed path from \(u\) to LowPt[w], which is in the same strong component as \(w\). Therefore \(u\) and \(w\) are in the same strong component. If this value is smaller than LowPt [ \(U\) ], then LowPt [ \(U\) ] is updated. Statement (3) is reached after all \(w\) adjacent to \(u\) have been considered. At this point the value of LowPt[U] is known. If LowPt[U]=DFNum[u], it means that there is no directed path from \(u\) to any ancestor of \(u\). Every vertex of the strong component containing \(u\) has been visited, and so is on the stack. These vertices are then popped off the stack before returning.
The complexity of Algorithm 10.4.1 is easily seen to be \(O(n+\varepsilon)\). For each vertex \(u\), all out-edges \(u w\) are considered, giving
\[
\sum_{v} d^{+}(v)=\varepsilon
\]
steps. Some arrays of length \(n\) are maintained. Each node is stacked once, and removed once from the stack.

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Algorithm 10.4.1: STRONGCOMPONENTS (G,n)
comment: Find the strong components using a depth-first search.
procedure DFS(U)
comment: extend the depth-first search to vertex \(u\)
DFCount \(\leftarrow D F\) Count+1
DFNum[u] \(\leftarrow D F C o u n t\)
LowPt \([\mathrm{U}] \leftarrow\) DFCount "initial value"
\(T o p \leftarrow T o p+1\)
Stack[Top] \(\leftarrow u\) "push u on Stack" OnStadc[u] \(\leftarrow\) true
for each \(w\) such that \(v \longrightarrow w\)
(if \(\operatorname{DFNum}[w]=0\)
then \(\left\{\begin{array}{l}\text { DFS }(w) \\ \text { if } \operatorname{LowPt}[w]<\operatorname{LowPt}[v] \\ \text { then } \operatorname{LowPt}[v] \leftarrow \operatorname{LowPt}[w]\end{array}\right.\)
else \(\left\{\begin{array}{l}\text { if } \operatorname{DFNum}[w]<\operatorname{DFNum}[v] \\ \text { then if } w \neq \text { parent of } v \\ \text { then if } \operatorname{OnStack}[w] \\ \text { then if } \operatorname{LowPt}[w]<\operatorname{LowPt}[v] \\ \text { then LowPt }[v] \leftarrow \operatorname{LowPt}[w]\end{array}\right.\)
if LowPt[U]=DFNum[U] (3)
then \(\left\{\begin{array}{l}\text { comment: }\left\{\begin{array}{l}\text { the points on the stack up to } v \text { form a } \\ \text { strong component - pop them off }\end{array}\right. \\ \text { repeat } \\ w \leftarrow \text { Stack }[\text { Top }] \\ \text { Top TTop-1 } \\ \text { OnStack }[w] \leftarrow \text { false } \\ \text { until } w=v\end{array}\right.\)
main
for \(u \leftarrow 1\) to \(n\)
do \(\left\{\begin{array}{l}\text { DFNum }[u] \leftarrow 0 \\ \text { OnStack }[v] \leftarrow \text { false }\end{array}\right.\)
DFCount \(\leftarrow 0\)
Тор \(\leftarrow 0\)
for \(u \leftarrow 1\) to \(n\)
do if \(\operatorname{DFNum}[u]=0\)
then DFS(u)
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\subsection*{10.4.1 An application to fabrics}

A fabric consists of two sets of strands at right angles to each other, called the warp and weft, woven together. The pattern in which the strands are woven can be represented by a rectangular matrix. Let the horizontal strands be \(h 1, h 2, \ldots, h m\) and let the vertical strands be \(U 1, U 2, \ldots, u n\). The matrix shown below contains an X wherever hi passes under \(u j\), and a blank otherwise. The pattern can be repeated as often as desired.
\begin{tabular}{l|llllll} 
& \(v_{1}\) & \(v_{2}\) & \(v_{3}\) & \(v_{4}\) & \(v_{5}\) & \(v_{6}\) \\
\(h_{1}\) & X & & X & X & & X \\
\(h_{2}\) & & X & & X & X & \\
\(h_{3}\) & X & & X & & & X
\end{tabular}

Suppose that the strand \(h 1\) were lifted. Since it passes under \(u 1, \cup 3, u 4\), and \(u 6\), these vertical strands would also be lifted. But since \(u 1\) passes under \(h 2\), this would in turn lift \(h 2\). Similarly lifting \(h 2\) would cause \(u 2\) to be lifted, which in turn causes \(h 3\) to be lifted. So the fabric hangs together if any strand is lifted.
For some pattern matrices, it is quite possible that the fabric defined does not hang together. For example, in the simplest case, a strand hi could lie under or over every uj, allowing it to be lifted off the fabric, or the fabric could fall apart into two or more pieces. In general, we can form a bipartite directed graph whose vertices are the set of all strands. The edges are
\(\{(u, w)\) : strand \(u\) lies under strand \(w\}\).
Call this the fabric graph. It is an oriented complete bipartite graph. If the fabric graph is strongly connected, then it hangs together, since there is a directed path from any strand to another. If the fabric graph is not strongly connected, then it can be separated into its strong components. Some strong component will lie
completely over another strong component, and be capable of being lifted off.

\section*{Exercises}
10.4.1 Prove Lemma 10.1.
10.4.2 Program the algorithm for strong components, and test it on the digraphs of Figures 10.1, 10.2, 10.3, 10.5 , and 10.6.
10.4.3 Find all digraphs which can be obtained by orienting a cycle of length 5 or 6.
10.4.4 Determine whether the fabric defined by the pattern matrix in Figure 10.7 hangs together.

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FI GURE 10.5
Find the strong components


FI GURE 10.6
Find the strong components
\begin{tabular}{|llllllll|}
\hline & & & X & & X & & \\
X & & & & & & X & \\
X & X & & X & X & X & X & \\
X & & & X & & X & & \\
X & & X & X & X & X & X & X \\
X & X & & X & & & X & X \\
X & & X & X & & X & X & \\
X & & & X & X & X \\
\hline
\end{tabular}

FI GURE 10.7
A pattern matrix

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10.5 Tournaments

In a round-robin tournament with \(n\) teams, each team plays every other team. Assuming that ties are not allowed, we can represent a win for team \(u\) over team \(u\) by a directed edge \((u, u)\). When all games have been played we have a directed complete graph. We say that a tournament is any oriented complete graph. It is easy to see that there are exactly two possible tournaments on three vertices, as shown in Figure 10.8.


\section*{FI GURE 10.8}

\section*{The tournaments on three vertices}

The second of these tournaments has the property that if \(u \longrightarrow v\) and \(v \longrightarrow w\), then \(u \longrightarrow w\). Any tournament which has this property for all vertices \(u, u, w\), is called a transitive tournament. It is easy to see that there is a unique transitive tournament \(T n\) for each \(n \geq 1\). For if \(T n\) is a transitive tournament, there must be a unique node \(u\) that is undefeated. If we delete it, we are left with a transitive tournament on \(n-1\) vertices. We use induction to claim that this is the unique \(T n-1\) When \(u\) is restored, we have the uniqueness of Tn.
If \(G\) is any tournament on \(n\) vertices, it will have a number of strong components. Let \(G^{*}\) denote its condensation. Then since \(G^{*}\) is acyclic, it is a transitive tournament on \(m \leq n\) vertices. We can find a topological ordering of \(V\left(G^{*}\right)\), and this will define an ordering of the strong components of \(G\). We can then make a list of the possible sizes of the strong components of \(G\), ordered according to the topological ordering of \(G^{*}\). We illustrate this for \(n=4\). The possible sizes of the strong components are (1, 1, 1, 1), (1, 3), (3, 1), and (4), since a simple digraph cannot have a strong component with only two vertices. The first ordering corresponds to the transitive tournament \(T 4\). The orderings \((1,3)\) and \((3,1)\) correspond to the first two tournaments of Figure 10.9. It is easy to see that they are unique, since there is only one strong tournament on three vertices, namely the directed cycle.
The third tournament is strongly connected. We leave it to the reader to verify that it is the only strong tournament on four vertices. The following theorem will
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\section*{FI GURE 10.9}

\section*{Non-transitive tournaments on 4 vertices}
be helpful. A digraph is said to be hamiltonian if it contains a directed hamilton cycle.
THEOREM 10.3 Every strong tournament on \(n \geq 3\) vertices is hamiltonian.
PROOF Let \(G\) be a strong tournament, and let \(C=(u 1, u 2, \ldots, u k)\) be the longest directed cycle in \(G\). If \(G\) is non-hamiltonian then \(\ell(C)<n\). Pick any \(v \notin C\). Because \(G\) is a tournament, either \(v \longrightarrow u_{1}\) or else \(u_{1} \longrightarrow v\). Without loss, suppose that \(u_{1} \longrightarrow v\). If \(v \longrightarrow u_{2}\), then we can produce a longer cycle by inserting \(u\) between \(u\) land \(u 2\). Therefore \(u_{2} \longrightarrow v\). If \(v \longrightarrow u_{3}\), then we can produce a longer cycle by inserting u between \(u 2\) and \(u 3\). Therefore \(u_{3} \longrightarrow v\), etc. Eventually we have \(v \longrightarrow u_{i}\) for all ui. This is impossible, since \(G\) is strong. We finish this section with the following theorem ??.

\section*{THEOREM 10.4 (Robbins' Theorem)}

Every 2-connected graph has a strong orientation.
PROOF Let \(G\) be a 2 -connected graph. Then \(G\) contains a cycle \(C\), which has a strong orientation. Let \(H\) be a subgraph of \(G\) with the largest possible number of vertices, such that \(H\) has a strong orientation. If \(u \notin H\), then since \(G\) is 2-connected, we can find two internally disjoint paths \(P\) and \(Q\) connecting \(u\) to \(H\). Orient \(P\) from \(u\) to \(H\), and \(Q\) from \(H\) to \(u\). This gives a strong orientation of a larger subgraph than \(H\), a contradiction. page_239

\section*{Exercises}
10.5.1 Show that there is a unique strong tournament on four vertices.
10.5.2 Find all the tournaments on five vertices. Show that there are exactly 12 tournaments, of which 6 are strong.
10.5.3 Show that every tournament has a hamilton path.
10.5.4 Show that if an odd number of teams play in a round robin tournament, it is possible for all teams to tie for first place. Show that if an even number of teams play, it is not possible for all teams to tie for first place.
10.5.5 Prove the following theorem. Let \(G\) be a digraph on \(n\) vertices such that \(d+(u)+d-(u) \geq n\) whenever \(u \nrightarrow v\). Then \(G\) is strong.
10.5.6 Describe a \(O(n+\varepsilon)\) algorithm to find a strong orientation of a 2-connected graph.
10.5.7 Show that every connected graph has an acyclic orientation.

\subsection*{10.6 2-Satisfiability}

In Chapter 9 we saw that 3-Sat is NP-complete. The related problem 2-Sat \(\in\) P. It has a number of practical applications.

Problem 10.1: 2-Sat
Instance: a set of boolean variables \(U\) and a boolean expression \(B\) over \(U\), in which each clause contains exactly two variables.
Question:is B satisfiable?
Consider the following instance of 2-Sat:
\[
(u 1+u 2)(u 1+\bar{u} 2)(u 2+u 3)(\bar{u} 1+\bar{u} 3)
\]

We want a truth assignment satisfying this expression. If \(u 1=\) false, the first clause tells us that \(u 2=\) true. We could write this implication as \(\bar{u} 1 \rightarrow u 2\). The second clause tells us that if \(u 1=\) false, then \(u 2=\) false. We could write this implication as \(\bar{u} 1 \rightarrow \bar{u} 2\). As this gives a contradiction, we conclude that \(u l=t r u e ~ i s ~ n e c e s s a r y . ~\). Continuing in this line of reasoning quickly gives the solution.
This example shows that a clause \((x+y)\) of an instance of 2-Sat, where \(x, y \in U \cup \bar{U}\), corresponds to two implications \(\bar{x} \rightarrow y\) and \(\bar{y} \rightarrow x\). We can construct a digraph with edges based on these implications.
Given an instance of 2-Sat with variables \(U\) and boolean expression \(B\), construct a digraph \(G\) whose vertex set is \(U \cup \bar{U}\). The edges of \(G\) consist of all ordered
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pairs \((\bar{x}, y)\) and \((\bar{y}, x)\), where \((x+y)\) is a clause of \(B\). \(G\) is called the implication digraph of \(B\). The implication digraph corresponding to the above expression is shown in Figure 10.10. A sequence of implications \(\bar{x} \rightarrow y \rightarrow z\) corresponds to a directed path in \(G\). Thus, directed paths in \(G\) are important. If any variable in a directed path is assigned the value true, then all subsequent variables in the path must also be true. Similarly, if any variable in a directed path is assigned the value false, then all previous variables must also be false. If \(G\) contains a directed cycle \(C\), then all variables of \(C\) must be true, or all must be false. We are talking about the strong components of \(G\).


\section*{FI GURE 10.10}

\section*{The implication digraph corresponding to an instance of 2-Sat}

The graph \(G\) has an antisymmetry-if there is an edge \((x, y)\), then there is also an edge \((\bar{y}, \bar{x})\), as can be seen
from the definition. Therefore the mapping of \(V(G)\) that interchanges every ui and \(\bar{u} i\) reverses the orientation of every edge.
Let G1, \(G 2, \ldots, G m\) be the strong components of \(G\). The variables in any strong component are either all true, or all false. If any strong component contains both ui and \(\bar{u} i\), for any \(i\), then the expression \(B\) is not satisfiable; for \(u i\) and \(\bar{u} i\) cannot be both true, or both false. So if \(B\) is satisfiable, ui and \(\bar{u} i\) are in different strong components.
The antisymmetry maps each strong component \(G j\) to another strong component \(G^{\prime} j\), the complementary strong component, such that \(x\) is in \(G j\) if and only if \(\bar{x}\) is in \(G^{\prime} j\). If all variables of \(G j\) are true, then all variables of \(G^{\prime} j\) must be false, and conversely. This gives the following algorithm for 2-Sat:

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> Algorithm 10.6.1: 2SAT(B,U)
> \{ Given an instance \(B\) of 2-Sat with variables \(U\),
> comment: construct a solution, if there is one.
> construct the implication digraph \(G\) corresponding to \(B\) construct the strong components of \(G\)
> for each \(u \in U\) do if \(u\) and \(\bar{u}\) are in the same strong component
> then return (NonSatisfiable)

construct the condensation of \(G\), and find a topological ordering of it let \(G 1, G 2, \ldots, G m\) be the topological ordering of the strong components
for \(i \leftarrow m\) downto 1
do \(\left\{\begin{array}{l}\text { if the variables of } G_{i} \text { have not been assigned } \\ \text { then }\left\{\begin{array}{l}\text { assign all variables of } G_{i} \text { to be true } \\ \text { assign all variables of } G_{i}^{\prime} \text { to be false }\end{array}\right.\end{array}\right.\)
In the graph of Figure 10.10 there are two strong components-the shaded vertices and the unshaded vertices. The condensation is a digraph with one edge, directed from left to right. The algorithm will assign true to all variables in the shaded strong component, and false to all variables in the unshaded one. This is the unique solution for this instance of 2-Sat.
THEOREM 10.5 Given an instance B of 2-Sat with variables U. Algorithm 2SAT(B,U)finds a solution if and only if a solution exists.
PROOF Every clause \((x+y)\) of \(B\) corresponds to two implications \(\bar{x} \rightarrow y\) and \(\bar{y} \rightarrow x\). The implication digraph \(G\) contains all these implications. Any assignment of truth values to the variables that satisfies all implications satisfies B. If some strong component of \(G\) contains both ui and \(\bar{u} i\) for some variable ui, then there is no solution. The algorithm will detect this. Otherwise, ui and ūi are always in complementary strong components. The algorithm assigns values to the variables such that complementary strong components always have opposite truth values. Therefore, for every ui and ūi exactly one will be true, and one will be false. Consider a variable \(x \in U \cup \bar{U}\). Suppose that \(x\) is in a strong component \(G j\). Its complement \(\bar{x}\) is in \(G^{\prime} j\). Without loss, suppose that \(G^{\prime} j\) precedes \(G j\) in the topological order. Then \(x\) will be assigned true and \(\bar{x}\) will be assigned false. All clauses \((x+y)\) containing \(x\) are thereby satisfied. All clauses \((\bar{x}+z)\) containing \(\bar{x}\) correspond to implications \(x \rightarrow z\) and \(\bar{z} \rightarrow \bar{x}\). It follows that \(z\) is either in the same strong component as \(x\), or else in a strong component following \(G j\), and \(\bar{z}\) is in a strong component preceding \(G^{\prime} j\). In either case, the algorithm has already assigned \(z \leftarrow\) true, so that the clause \((\bar{x}+z)\) is also satisfied. We conclude that page_242

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the truth assignment constructed by the algorithm satisfies \(B\).
This also gives the following theorem.
THEOREM 10.6 Given an instance \(B\) of 2-Sat with variables \(U\). \(B\) is satisfiable if and only if, for every \(u_{i} \in U\), ui and \(\bar{i} i\) are contained in different strong components of the implication digraph.
If \(U\) has \(n\) variables, and \(B\) has \(k\) clauses, the implication graph \(G\) will have \(2 n\) vertices and \(2 k\) edges. It takes \(O(n+k)\) steps to construct \(G\), and \(O(n+k)\) steps to find its strong components, and to construct a topological order of them. It then takes \(O(n)\) steps to assign the truth values. Thus, we have a linear algorithm that
solves 2-Sat.

\section*{Exercises}
10.6.1 Construct the implication digraph for the following instance of 2-Sat.
\((u 1+u 2)(\bar{u} 1+u 3)(\bar{u} 3+\bar{u} 4)(u 1+u 4)(\bar{u} 2+\bar{u} 5)(u 5+\bar{u} 6)(u 2+u 6)(\bar{u} 3+u 4)\)
10.6.2 Solve the previous instance of 2-Sat.
10.6.3 Given an instance of 2-Sat with the additional requirement that \(u 1=\) true. Show how to convert this into an instance of 2-Sat and solve it. Show also how to solve it if \(u 1\) is required to be false.
10.6.4 Consider an instance of Sat in which each clause has exactly two variables, except that one clause has three or more variables. Describe an algorithm to solve it in polynomial time.

\subsection*{10.7 Notes}

An excellent reference for digraphs is BANG-J ENSEN and GUTIN [8]. The algorithms for strong components is from AHO, HOPCROFT, and ULLMAN [1]. Strong components and 2-satisfiability are further examples of the importance and efficiency of the depth-first search. The subject of tournaments is a vast area. A survey can be found in REID and BEINEKE [102]. A good monograph on the topic is MOON [89].
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\section*{Graph Colorings}

\subsection*{11.1 I ntroduction}

A coloring of the vertices of a graph \(G\) is an assignment of colors to the vertices. A coloring is proper if adjacent vertices always have different colors. We shall usually be interested only in proper colorings. It is clear that the complete graph \(K n\) requires \(n\) distinct colors for a proper coloring. Any bipartite graph can be colored in just two colors. More formally,
DEFINITION 11.1: An \(m\)-coloring of \(G\) is a mapping from \(V(G)\) onto the set \(\{1,2, \ldots, m\}\) of \(m\) "colors". The chromatic number of \(G\) is \(\chi(G)\), the minimum value \(m\) such that \(G\) has a proper \(m\)-coloring. If \(\chi(B)=m, G\) is then said to be \(m\)-chromatic.
If \(G\) is bipartite, we know that \(\chi(G)=2\). Moreover, there is a \(O(\varepsilon)\) algorithm to determine whether an arbitrary \(G\) is bipartite, and to construct a 2 -coloring of it. When \(\chi(G) \geq 3\), the problem becomes NP-complete. We show this in Section 11.7. (See Problem 11.1.) A consequence of this is that there is no complete theoretical characterization of colorability. As with the Hamilton cycle problem, there are many interesting techniques, but most problems only have partial solutions. We begin with a simple algorithm for coloring a graph, the sequential algorithm. We will indicate various colorings of \(G\) by the notation \(x 1, x 2, \ldots\), etc. While \(\chi(G)\) represents the chromatic number of \(G\), there may be many colorings of \(G\) that use this many colors. If \(\chi 1\) is a coloring, then \(\chi 1(u)\) represents the color assigned to vertex \(u\) under the coloring \(x 1\).

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Algorithm 11.1.1: SEQUENTIALCOLORI NG(G,n)
comment: Construct a coloring \(X 1\) of a graph \(G\) on \(n\) vertices
mark all vertices "uncolored"
order \(V(G)\) in some sequence \(u 1, U 2, \ldots\), un
for \(i \leftarrow n\) downto 1
do \(\chi 1\) ( \(u i) \leftarrow\) the first color available for \(u i\)
To calculate the first available color for \(u i\), we consider each \(u \rightarrow v i\). If \(u\) is already colored, we mark its color "in use". We then take the first color not in use as \(\chi 1\) (ui), and then reset the color flags before the next iteration. Thus, iteration \(i\) of the for-loop takes \(O\) (DEG(ui)) steps. Algorithm 11.1.1 is then easily seen to be \(O(\varepsilon)\). The number of colors that it uses usually depends on the sequence in which the vertices are taken. If we knew a proper \(\chi(G)\)-coloring before beginning the algorithm, we could order the vertices by color: all vertices of color 1 last, then color 2, etc. Algorithm 11.1.1 would then color \(G\) in exactly \(\chi(G)\) colors. Since we don't know \(\chi(G)\) beforehand, we investigate various orderings of \(V(G)\).
Spanning trees often give useful orderings of the vertices. We choose a vertex \(u 1\) as a root vertex, and build a spanning tree from it. Two immediate possibilities are a breadth-first or depth-first tree. The ordering of the vertices would then be the order in which they are numbered in the spanning tree. This immediately gives the following lemma.
LEMMA 11.1 If \(G\) is a connected, non-regular graph, then \(\chi(G) \leq \Delta(G)\). If \(G\) is regular, then \(\chi(G) \leq \Delta(G)+1\).
PROOF Assume first that \(G\) is non-regular, and choose a vertex \(u 1\) of degree \(<\Delta(G)\) as the root of a spanning tree. Apply Algorithm 11.1.1 to construct a coloring \(x 1\). When each vertex \(u i \neq u 1\) comes to be
colored, the parent of \(u i\) has not yet been colored. Therefore at most DEG(ui)-1 adjacent vertices have already been colored. Hence \(\chi 1(u i) \leq \Delta(G)\). When \(u l\) comes to be colored, all adjacent vertices have already been colored. Since \(\operatorname{DEG}(u 1)<\Delta(G)\), we conclude that \(\chi 1(u 1) \leq \Delta(G)\). Hence \(\chi(G) \leq \Delta(G)\).
If \(G\) is regular, then the proof proceeds as above, except that \(x 1(u 1) \leq \Delta(G)+1\). The conclusion follows. Notice that if \(G\) is regular, only one vertex needs to use color \(\Delta(G)+1\).
If Algorithm 11.1.1, using a breadth-first or depth-first spanning tree ordering, is applied to a complete graph \(K n\), it is easy to see that it will use exactly \(n\) colors. If the algorithm is applied to a cycle Gn, it will always use two colors if \(n\) is even,
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and three colors, if \(n\) is odd. Complete graphs and odd cycles are special cases of Brooks' theorem.
THEOREM 11.2 (Brooks' theorem) Let G be a connected graph that is not a complete graph, and not an odd cycle. Then \(\chi(G) \leq \Delta(G)\).
PROOF Suppose first that \(G\) is 3 -connected. If \(G\) is non-regular, we know the result to be true. Hence we assume that \(G\) is a regular graph. Since \(G\) is not complete, we can choose vertices \(u, u, w\) such that \(u \rightarrow u, w\), but \(v \nrightarrow w\). Construct a spanning tree ordering of \(G-\{u, w\}\), with \(u\) as root, so that \(u 1=u\). Then set \(u n-1=u\), \(u n=w\), and apply Algorithm 11.1.1 to construct a coloring \(x 1\) of \(G\). Vertices \(u\) and \(w\) will both be assigned color 1. Each \(u i \neq u 1\) will have \(x 1(u i) \leq \Delta(G)\). When \(u l\) comes to be colored, the two adjacent vertices \(u\) and \(w\) will have the same color. Hence \(\chi 1(u 1) \leq \Delta(G)\).
Otherwise \(G\) is not 3 -connected. Suppose that \(G\) is 2 -connected. Choose a pair of vertices \(\{u, u\}\), such that \(G-\{u, u\}\) is disconnected. If \(H\) is any connected component of \(G-\{u, u\}\), let Huv be the subgraph of \(G+u u\) induced by \(V(H) \cup\{u, v\}\). Now \(G\) is regular, of degree at least three (or \(G\) would be a cycle). Therefore we can choose \(u\) and \(u\) so that Huu is a non-regular graph. Therefore \(\chi(H u u) \leq \Delta(H u u)\). But \(\Delta(H u u) \leq \Delta(G)\). Color Huu in at most \(\Delta(G)\) colors. Notice that \(u\) and \(u\) have different colors, because \(u v \in E\left(H_{u v}\right)\). We can do the same for every subgraph Kuu so constructed from each connected component of \(G-\{u, u\}\). Furthermore, we can require that \(u\) and \(u\) are colored identically in each Kuu, by permuting colors if necessary. These subgraph colorings determine a coloring of \(G\) with at most \(\Delta(G)\) colors.
If \(G\) is not 2-connected, but has a cut-vertex \(u\), we use an identical argument, deleting only \(u\) in place of \(\{u, u\}\).

\subsection*{11.1.1 Intersecting lines in the plane}

An interesting example of the use of Algorithm 11.1.1 is given by intersecting lines in the plane. Suppose that we are given a collection of \(m\) straight lines in the plane, with no three concurrent. Construct a graph \(G\) whose vertices are the points of intersection of the lines, and whose edges are the line segments connecting the vertices. An example is shown in Figure 11.1. We can use Algorithm 11.1.1 to show that \(\chi(G) \leq 3\). Notice first that because at most two lines are concurrent at a point that we have \(\Delta(G) \leq 4\). Also note that \(G\) can be enclosed by a disc in the plane. Choose a disc in the plane containing \(G\), and choose a line \(\ell\) in the plane that is outside the disc, such that \(\ell\) is not parallel to any of the original \(m\) lines. The dotted line in Figure 11.1 represents \(\ell\). Assign an orientation to the edges of \(G\) by directing each edge toward \(\ell\). This converts \(G\) to an acyclic digraph. Notice
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that each vertex has at most two incident edges oriented outward, and at most two oriented inward. Let u1, \(u 2, \ldots\), un be a topological ordering of \(V(G)\). Apply Algorithm 11.1.1. The vertices of \(G\) have degree two, three, or four. When each ui comes to be colored, at most two adjacent vertices have already been colored-those incident on out-edges from \(u i\). Therefore \(\chi 1(u i) \leq 3\), for each \(u i\). It follows that \(\chi(G) \leq 3\).

\section*{FI GURE 11.1}

\section*{I ntersecting lines in the plane \\ Exercises}
11.1.1 Write computer programs to apply the sequential algorithm using breadth-first and depth-first searches, and compare the colorings obtained.
11.1.2 Show that if a breadth-first or depth-first sequential algorithm is applied to a bipartite graph, that exactly two colors will be used.
11.1.3 Construct a complete bipartite graph \(K n, n\) with bipartition \(X=\{x 1, x 2, \ldots, x n\}, Y=\{y 1, y 2, \ldots, y n\}\), and remove the matching \(M=\{x 1 y 1, \ldots, x n y n\}\), to get \(G=K n, n-M\). Order the vertices \(x 1, y 1, x 2, y 2, \ldots, x n, y n\). Show that Algorithm 11.1.1 with this ordering will use \(n\) colors.
11.1.4 Construct a complete tripartite graph \(K n, n, n\) with tripartition \(X=\{x 1, x 2, \ldots, x n\}, Y=\{y 1, y 2, \ldots, y n\}\), \(Z=\{z 1, z 2, \ldots, z n\}\), and remove the triangles \(T=\{x 1 y 1 z 1, \ldots, x n y n z n\}\), to get \(G=K n, n, n-T\). Order the vertices \(x 1, y 1, z 1, x 2, y 2, z 2, . ., x n, y n, z n\). Show that Algorithm 11.1.1 with this ordering will use \(n\) colors. How many colors will be used by the breadth-first sequential algorithm?

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11.1.5 Show that in the intersecting lines problem, if three or more lines are allowed to be concurrent, that \(x(G)\) can be greater than three.
11.1.6 Let \(G 1\) and \(G 2\) be graphs with \(m\)-colorings \(\chi 1\) and \(\chi 2\), respectively. We say that \(G 1\) and \(G 2\) are colorisomorphic if there is an isomorphism \(\theta\) from \(G 1\) to \(G 2\) that induces a permutation of the colors. More formally, if \(u 1, u_{2} \in V\left(G_{1}\right)\) are mapped by \(\theta\) to \(u 1, v_{2} \in V\left(G_{2}\right)\), respectively, then \(\chi 1(u 1)=\chi 1(u 2)\) if and only if \(\chi 2(u 1)=\chi 2(u 2)\). Show how to construct graphs \(G_{1}^{\prime}\) and \(G_{2}^{\prime}\) such that \(G 1\) and \(G 2\) are color-isomorphic if and only if \(G_{1}^{\prime}\) and \(G_{2}^{\prime}\) are isomorphic.
11.1.7 Determine whether the graphs of Figure 11.2 are color-isomorphic, where the colors are indicated by the numbers.
11.1.8 Let the vertices of \(G\) be listed in the sequence \(u 1, u 2, \ldots, u n\) and apply Algorithm 11.1.1. Show that \(x 1\) (ui) isatmost \(\min \{n-i+1, \mathrm{DEG}(u i)+1\}\). Conclude that \(\chi(G) \leq \max _{i=1}^{n} \min \left\{n-i+1, \operatorname{DEG}\left(v_{i}\right)+1\right\}\).


\section*{FI GURE 11.2}

\section*{Are these color-isomorphic?}

\subsection*{11.2 Cliques}

Let \(G\) be a graph with a proper coloring. The subset of vertices of color \(i\) is said to be the color class with color \(i\). These vertices induce a subgraph with no edges. In \(\bar{G}\) they induce a complete subgraph.
DEFINITION 11.2: A clique in \(G\) is an induced complete subgraph. Thus a clique is a subset of the vertices that are pairwise adjacent. An independent set is

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a subset of \(V(G)\) which induces a subgraph with no edges. A clique is a maximum clique if \(G\) does not contain any larger clique. Similarly, an independent set is a maximum independent set if \(G\) does not contain any larger independent set. An independent set is also called a stable set.
The problem of finding a maximum clique or maximum independent set in \(G\) is NP-complete, as shown in Section 11.7.
Write \(a(G)\) for the number of vertices in a maximum independent set in \(G\), and \(\bar{\alpha}(G)\) for the number of \(\bar{\alpha}(G)\),
vertices in a maximum clique. It is clear that \(\chi(G) \geq\)
since a clique in \(G\) of \(m\) vertices requires at least \(m\) colors. If we are given a coloring \(x 1\) of \(G\) and \(\chi 2\) of \(\bar{G}\), let \(V \max\) be a largest color class in \(G\), and let \(\bar{V}_{\max }\) be a largest color class in \(\bar{G}\). Since \(\bar{V}_{\max }\) induces a clique in \(G\), we have \(\bar{\alpha}(G) \geq\left|\bar{V}_{\max }\right|\). Since each color class is an independent set, \(a(G) \geq|V \max |\).
This gives the bounds
\[
\left|\bar{V}_{\max }\right| \leq \bar{\alpha}(G) \leq \chi(G)
\]
and
\[
\left|V_{\max }\right| \leq \alpha(G) \leq \chi(\bar{G})
\]

Although we don't know \(\chi(G)\) and \(\chi(\bar{G})\), we can use the sequential Algorithm 11.1.1 to construct colorings \(\chi 1\) and \(\chi 2\) of \(G\) and \(\bar{G}\), so as to obtain bounds on the clique number and independent set number of \(G\) and \(\bar{G}\). We write \(\chi 1(G)\) to denote the number of colors used by the coloring \(\chi 1\).
LEMMA 11.3 If \(\chi 1\) and \(\chi 2\) satisfy \(\left|V_{\max }\right|=\chi_{2}(\bar{G})\), then \(\alpha(G)=\chi(\bar{G})=\chi_{2}(\bar{G})\). If \(\chi\) 1and \(\chi 2\) satisfy \(\left|\bar{V}_{\max }\right|=\chi_{1}(G)\), then \(\bar{\alpha}(G)=\chi(G)=\chi_{1}(G)\).
PROOF The inequalities above hold for any proper colorings \(\chi 1\) and \(\chi 2\). If \(\left|V_{\max }\right|=\chi_{2}(\bar{G})\), then the second inequity determines \(a(G)\). Therefore \(\chi(\bar{G})\) is at least as big as this number. But since we have a coloring in \(\chi_{2}(\bar{G})\) colors, this determines \(\chi(\bar{G})\).
Thus, by using Algorithm 11.1 .1 to color both \(G\) and \(\bar{G}\), we can obtain bounds on \(\bar{\alpha}(G)\), and \(\chi(\bar{G})\).
Sometimes this will give us exact values for some of these parameters. When it does not give exact values, it gives colorings \(\chi 1\) and \(\chi 2\), as well as a clique \(\bar{V}_{\max }\) and independent set \(V\) max in \(G\). In general, Algorithm 11.1.1 does not construct colorings that are optimal, or even nearly optimal. There are many variations of Algorithm 11.1.1. Improvements to Algorithm 11.1.1 will give improvements in the above bounds, by decreasing the number of colors used, and increasing the size of the maximum color classes found. One modification that is found to work well in practice is the degree saturation method, which orders the vertices by the "saturation" degree.
Consider a graph \(G\) for which a coloring \(X 1\) is being constructed. Initially all vertices are marked uncolored. On each iteration of the algorithm, another vertex
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is colored. For each vertex, let \(c(u)\) denote the number of distinct colors adjacent to \(u . c(u)\) is called the saturation degree of \(u\). Initially \(c(u)=0\). At any stage of the algorithm the vertices are partially ordered. A vertex \(u\) such that the pair \((c(u), \operatorname{DEG}(u))\) is largest is chosen as the next vertex to color; that is, vertices are compared first by \(c(u)\), then by DEG(u).

Algorithm 11.2.1: DEGREESATURATION(G, n)
comment: Construct a coloring \(X 1\) of a graph \(G\) on \(n\) vertices mark all vertices "uncolored"
initialize \(c(u) \leftarrow 0\), for all \(u\)
for \(\leftarrow \leftarrow 1\) to \(n\)
> do \(\left\{\begin{array}{l}\text { select } u \text { as a vertex with largest }(c(u), \operatorname{DEG}(u)) \\ \chi_{1}(u) \leftarrow \text { the first color available for } u \\ \text { for each } \mathrm{v} \longrightarrow \mathrm{u} \\ \text { do if } v \text { is uncolored, adjust } c(v)\end{array}\right.\)

Algorithm 11.2.1 requires a priority queue in order to efficiently select the next vertex to color. In practice, it is found that it uses significantly fewer colors than Algorithm 11.1.1 with a fixed ordering of the vertices. However, algorithms based on the sequential algorithm are limited in their efficacy. JOHNSON [69] discusses the limitations of many coloring algorithms based on the sequential algorithm.
The first vertex, \(u 1\), that Algorithm 11.2.1 colors will be one of maximum degree. The second vertex, \(u 2\), will be a vertex adjacent to \(u 1\) of maximum possible degree. The third vertex, \(u 3\), will be adjacent to \(u 1\) and \(u 2\), if there is such a vertex, and so on. Thus, Algorithm 11.2.1 begins by constructing a clique of vertices of large degree. The algorithm could save this information, and use it as a lower bound on \(\bar{\alpha}(G)\).
A recursive search algorithm MAXCLI QUE() to find a maximum clique can be based on a similar strategy.
Algorithm 11.2.2 that follows constructs cliques \(C^{\prime}\) The maximum clique found is stored in a variable C. C and
\(C^{\prime}\) are stored as global arrays. It also uses an array \(S\) of vertices eligible to be added to \(C^{\prime}\). The set of all neighbors of \(u\) is denoted by \(N(u)\).
This algorithm builds a clique \(C^{\prime}\) and tries to extend it from a set \(S^{\prime}\) of eligible vertices. When \(\left|C^{\prime}\right|+\left|S^{\prime}\right|\) is smaller than the largest clique found so far, it backtracks. The performance will depend on the ordering of the vertices used. As with Algorithm 11.2.1, the ordering need not be chosen in advance, but can be constructed as the algorithm progresses. For example, ui might be selected as a vertex of largest degree in the subgraph induced by the vertices not yet considered. If a maximum clique is found early in the algorithm, the remaining calls to EXTENDCLIQUE() will finish more quickly. We can estimate the complexity by

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noticing that the loop in MAXCLIQUE() runs at most \(n\) times. After each choice of \(u i\), the set \(S\) will contain at most \(\Delta(G)\) vertices. There are \(2 \Delta(G)\) subsets of a set of size \(\Delta(G)\). The algorithm might construct each subset at most once. Therefore the complexity is at most \(O(n \cdot 2 \Delta(G))\), an exponential value.

\section*{Algorithm 11.2.2: \(\operatorname{MAXCLIQUE}(G, n)\)}
comment: Construct a maximum clique \(C\) in \(G\)
procedure EXTENDCLIQUE \((S, u)\)
comment: \(u\) has just been added to \(C^{\prime}\)-adjust \(S\) and extend the clique
\[
\begin{gathered}
S^{\prime} \leftarrow \operatorname{SNN}(U) \\
\text { if }\left|S^{\prime}\right|=0
\end{gathered}
\]
then \(\left\{\begin{array}{l}\text { if }\left|C^{\prime}\right|>|C| \text { then } C \leftarrow C^{\prime} \\ \text { return }\end{array}\right.\)
while \(\left|S^{\prime}\right|>0\) and \(\left|C^{\prime}\right|+\left|S^{\prime}\right|>|C|\)
do \(\left\{\begin{array}{l}\text { select } u \in S^{\prime} \\ C^{\prime} \leftarrow C^{\prime} \cup u \\ S^{\prime} \leftarrow S^{\prime}-u\end{array}\right.\)
ExtendClique \(\left(S^{\prime}, u\right)\) \(C^{\prime} \leftarrow C^{\prime}-u\)
main
choose an ordering \(u 1, u 2, \ldots\), un of the vertices
\(C \leftarrow \emptyset\) "largest clique found so far"
\(C^{\prime} \leftarrow \emptyset\) "clique currently being constructed"
\(S \leftarrow V(G)\) "vertices eligible for addition to \(C^{\prime \prime}\) "
\(i \leftarrow 1\)
while \(|S|>\mid C\)
do \(\left\{\begin{array}{l}C^{\prime} \leftarrow\left\{v_{i}\right\} \\ S \leftarrow S-v_{i} \\ \text { Extendelique }\left(S, v_{i}\right) \\ i \leftarrow i+1\end{array}\right.\)
comment: \(C\) is now a maximum clique of size \(\bar{\alpha}\)
\[
\bar{\alpha} \leftarrow|C|
\]
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\subsection*{11.3 Mycielski's construction}

A graph \(G\) which contains a 4 -clique necessarily has \(\chi(G) \geq 4\). However, it is possible for a graph with no 4clique to have \(x(G) \geq 4\). Mycielski found a way to construct triangle-free graphs with arbitrarily large chromatic number.
We start with a triangle-free graph \(G\) with \(\chi(G) \geq 3\). Any odd cycle with five or more vertices will do (e.g., \(G=C 5)\). We now extend \(G\) to a graph \(G^{\prime}\) as follows. For each \(v \in V(G)\), we add a vertex \(u^{\prime}\) to \(G^{\prime}\) adjacent to the same vertices of \(G\) that \(u\) is adjacent to. We now add one more vertex \(u 0\) adjacent to each \(u^{\prime}\). Thus, if \(G\) has \(n\) vertices, \(G^{\prime}\) will have \(2 n+1\) vertices. Refer to Figure 11.3.


FI GURE 11.3

\section*{Mycielski's construction}

LEMMA \(11.4 \chi\left(G^{\prime}\right)=\chi(G)+1\). Furthermore, \(G^{\prime}\) has no triangles if \(G\) has none.
PROOF Consider a coloring \(\chi 1\) of \(G^{\prime}\). Since it induces a coloring of \(G\), we conclude that \(\chi\left(G^{\prime}\right) \geq \chi(G)\). Let \(m=x(G)\). Some vertex of \(G\) with color number \(m\) is adjacent to \(m-1\) other colors in \(G\); for otherwise each vertex of color \(m\) could be recolored with a smaller color number. It follows that the vertices \(u\) ' must be colored with at least \(x(G)\) colors. In fact, we could assign each \(x 1\left(u^{\prime}\right)=x 1(u)\). The vertex \(u 0\) is adjacent to \(\chi(G)\) colors, so that \(\chi 1(u 0)=m+1\). It follows that \(\chi\left(G^{\prime}\right)=\chi(G)+1\). It is easy to see that \(G^{\prime}\) has no triangles, because \(G\) has none.

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By iterating Mycielski's construction, we can construct triangle-free graphs with large chromatic numbers. In fact, if we begin the construction with \(G=K 2\), the result is C5. Or we could start Mycielski's construction with a graph which contains triangles, but is \(K 4\)-free, and construct a sequence of \(K 4\)-free graphs with increasing chromatic numbers, and so forth.

\subsection*{11.4 Critical graphs}

Let \(G\) be a graph with \(x(G)=m\). If we remove an edge \(u u\) from \(G\), there are two possibilities, either \(x(G-u u)=m\) or \(x(G-u u)=m-1\). In the latter case, we say that edge \(u v\) is critical.
DEFINITION 11.3: A graph \(G\) is critical if \(\chi(G-u U)=\chi(G)-1\) for all edges \(u v \in E(G)\). If \(\chi(G)=m\), we say that \(G\) is \(m\)-critical.
It is easy to see that every graph contains a critical subgraph. If \(\chi(G-u u)=\chi(G)\) for some edge \(u u\), we can remove uu. Continue deleting edges like this until every edge is critical. The result is a critical subgaph.
Critical graphs have some special properties.
LEMMA 11.5 If \(G\) is \(m\)-critical, then \(\delta(G) \geq m-1\).
PROOF If DEG \((u)<m-1\), choose an edge \(u u\), and color \(G-u u\) with \(m-1\) colors. Since DEG \((u)<m-1\), there are at most \(m-2\) adjacent colors to \(u\) in \(G\). So there is always a color in \(\{1,2, \ldots, m-1\}\) with which \(u\) can be colored to obtain an ( \(m-1\) )-coloring of G , a contradiction.
If \(G\) is an \(m\)-critical graph, then \(G\) has at least \(m\) vertices, and each has degree at least \(m-1\). Therefore every graph with \(\chi(G)=m\) has at least \(m\) vertices of degree \(\geq m-1\).
LEMMA 11.6 Every critical graph with at least three vertices is 2-connected.
PROOF Suppose that \(G\) is an \(m\)-critical graph with a cut-vertex \(u\). Let \(H\) be a connected component of \(G-u\), and let \(H u\) be the subgraph induced by \(V(H) \cup\{v\}\). Color \(H u\) with \(\leq m-1\) colors. Do the same for every such subgraph Hu . Ensure that \(u\) has the same color in each subgraph, by permuting colors if necessary. The result is a coloring of \(G\) in \(\leq m-1\) colors, a contradiction.
The ideas of the Lemma 11.6 can be extended to separating sets in general.
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LEMMA 11.7 Let S be a separating set in an m-critical graph \(G\). Then \(S\) does not induce a clique.
PROOF If \(S\) is a separating set, let \(H\) be a component of \(G-S\), and consider the subgraph \(H S\) induced by \(V(H) \cup S\). It can be colored in \(m-1\) colors. The vertices of \(S\) are colored with \(|S|\) distinct colors, which can be permuted in any desired way. Do the same for every component of \(G-S\). The result is a coloring of \(G\) in
m-1 colors.
It follows from this lemma that if \(\{u, u\}\) is a separating set in an \(m\)-critical graph, that \(u v \notin E(G)\). Suppose that \(\{u, u\}\) is a separating set. Let \(H\) be a component of \(G-\{u, u\}\), and let \(K\) be the remaining components. Construct Huu induced by \(V(H) \cup\{u, v\}\), and Kuu induced by \(V(K) \cup\{u, v\}\). Huu and Kuu can both be colored in \(m-1\) colors. If Huu and Kuu both have ( \(m-1\) )-colorings in which \(u\) and \(u\) have different colors, then we can use these colorings to construct an ( \(m-1\) )-coloring of G. Similarly, if Huu and Kuu both have ( \(m-1\) )-colorings in which \(u\) and \(u\) have the same color, we can again construct an ( \(m-1\) )-coloring of \(G\). We conclude that in one of them, say Huu, \(u\) and \(u\) have the same color in every ( \(m-1\) )-coloring; and that in \(K u u, u\) and \(u\) have different colors in every ( \(m-1\) )-coloring.
Now consider the graph \(H^{\prime}=H u u+u u\). It cannot be colored in \(m-1\) colors, however \(H^{\prime}-u u\) can be. Let \(x y\) be any other edge of \(H^{\prime}\). Then \(G-x y\) can be colored in \(m-1\) colors. Now Kuu is a subgraph of \(G-x y\). Therefore \(u\) and \(u\) have different colors in this coloring. It follows that \(H^{\prime}-x y\) can be colored in \(m-1\) colors. Hence, \(H\) \('=H u u+u v\) is an \(m\)-critical graph.
Now consider the graph \(K^{\prime}=(K u u+u u) \cdot u u\). It cannot be colored in \(m-1\) colors, as this would determine an ( \(m-1\) )-coloring of \(K u u\) in which \(u\) and \(u\) have the same color. Let \(x y\) be any edge of \(K^{\prime}\). It corresponds to an edge \(x^{\prime} y^{\prime}\) of \(G\). \(G-x^{\prime} y^{\prime}\) can be colored in ( \(m-1\) ) colors. Since Huu is a subgraph of \(G\), it follows that \(u\) and \(u\) have the same color. This then determines a coloring of \(K^{\prime}-x y\) in \((m-1)\) colors. Hence \(K^{\prime}=(K u u+u u) \cdot u u\) is also an \(m\)-critical graph.

\section*{Exercises}
11.4.1 Program Algorithm 11.2.1, and compare its performance with a breadth-first or depth-first sequential algorithm.
11.4.2 Program the MAXCLIQUE() algorithm.
11.4.3 Let \(G\) be an \(m\)-critical graph, and let \(v \in V(G)\). Show that \(G\) has an \(m\)-coloring in which \(u\) is the only vertex of color number \(m\).
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11.4.4 Let \(G\) be an \(m\)-critical graph. Apply Mycielski's construction to obtain a graph \(G^{\prime}\). Either prove that \(G^{\prime}\) is \((m+1)\)-critical, or find a counterexample.

\subsection*{11.5 Chromatic polynomials}

Suppose that we wish to properly color the complete graph \(K n\) in at most \(\lambda\) colors, where \(\lambda \geq n\). Choose any ordering of \(V(K n)\). The first vertex can be colored in A choices. The next vertex in \(\lambda-1\) choices, and so on. Thus, the number of ways to color \(K n\) is \(\lambda(\lambda-1)(\lambda-2) \ldots(\lambda-n+1)\). This is a polynomial of degree \(n\) in \(\lambda\). DEFINITION 11.4: The chromatic polynomial of a graph \(G\) is \(n(G, \lambda)\), the number of ways to color \(G\) in \(\leq \lambda\) colors.
In order to show that \(n(G, \lambda)\) is in fact a polynomial in \(\lambda\), we use a method that is familiar from counting spanning trees. We first find \(n(T, \lambda)\) for any tree \(T\), and then give a recurrence for any graph \(G\).
LEMMA 11.8 Let \(T\) be a tree on \(n\) vertices. Then \(n(T, \lambda)=\lambda(\lambda-1) n-1\).
PROOF By induction on \(n\). It is certainly true if \(n=1\) or \(n=2\). Choose a leaf \(u\) of \(T\), and let \(T^{\prime}=T-u . T^{\prime}\) is a tree on \(n-1\) vertices. Therefore \(n\left(T^{\prime}, \lambda\right)=\lambda(\lambda-1) n-2\). In any coloring of \(T^{\prime}\), the vertex adjacent to \(u\) in \(T\) has some color. There are \(\lambda-1\) colors available for \(u\). Every coloring of \(T\) arises in this way. Therefore \(n(T\), \(\lambda)=\lambda(\lambda-1) n-1\).
Suppose now that \(G\) is any graph. Let \(u v \in E(G)\). In practice we will want to choose \(u u\) so that it is an edge on a cycle.
THEOREM \(11.9 n(G, \lambda)=\Pi(G-u u, \lambda)-\Pi(G u u, \lambda)\).
PROOF In each coloring of \(G-u v\) in \(\leq \lambda\) colors, either \(u\) and \(u\) have different colors, or they have the same color. The number of colorings of \(G-u u\) is the sum of these two. If \(u\) and \(u\) have different colors, then we have a coloring of \(G\) in \(\leq \lambda\) colors. Conversely, every coloring of \(G\) in \(\leq \lambda\) colors gives a coloring of \(G-u u\) in which \(u\) and \(u\) have different colors. If \(u\) and \(u\) have the same color in \(G-u u\), then this gives a coloring of \(G \cdot u u\). Any coloring of \(G \cdot u u\) in \(\leq \lambda\) colors determines a coloring of \(G-u u\) in which \(u\) and \(u\) have the same color. We conclude that \(\Pi(G-u u, \lambda)=\Pi(G, \lambda)+\Pi(G \cdot u u, \lambda)\).
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One consequence of this theorem is that \(n(G, \lambda)\) is in fact a polynomial of degree \(n\) in \(\lambda\). Now if \(n>0\), then \(\lambda \mid n(G, \lambda)\), since \(G\) cannot be colored in \(\lambda=0\) colors. Similarly, if \(\varepsilon(G) \neq 0\), we conclude that \(\lambda(\lambda-1) \mid \Pi(G, \lambda)\), since \(G\) cannot be colored in \(\lambda=1\) color. If \(G\) is not bipartite, then it cannot be colored in \(\lambda=2\) colors. In this case \(\lambda(\lambda-1)(\lambda-2) \mid \Pi(G, \lambda)\). In general:

LEMMA 11.10 If \(\chi(G)=m\), then \(\lambda(\lambda-1)(\lambda-2) \ldots(\lambda-m+1) \mid n(G, \lambda)\).
PROOF \(G\) cannot be colored in fewer than \(m\) colors. Therefore \(n(G, \lambda)=0\), for \(\lambda=1,2, \ldots, m-1\).
Notice that if \(G\) contains an \(m\)-clique \(S\), then \(\chi(G) \geq m\), so that \(\lambda(\lambda-1)(\lambda-2) \ldots(\lambda-m+1) n(G, \lambda)\). There are \(\lambda(\lambda-1)(\lambda-2) \ldots(\lambda-m+1)\) ways to color \(S\) in \(\leq \lambda\) colors. The number of ways to complete a coloring of \(G\), given a coloring of \(S\), is therefore \(\Pi(G, \lambda) / \lambda(\lambda-1)(\lambda-2) \ldots(\lambda-m+1)\).
Suppose that \(G\) has a cut-vertex \(u\). Let \(H\) be a connected component of \(G-u\), and let \(H u\) be the subgraph induced by \(V(H) \cup\{v\}\). Let \(K u\) be the subgraph \(G-V(H)\). Every coloring of \(G\) induces colorings of Hu and \(K u\), such that \(u\) has the same color in both. Every coloring of Hu and \(K u\) occurs in this way. Given any coloring of \(H u\), there are \(\Pi(K u, \lambda) / \lambda\) ways to complete the coloring of \(K u\). It follows that \(n(G, \lambda)=\Pi(H u, \lambda) \pi(K u, \lambda) / \lambda\).
More generally, suppose that \(S\) is a separating set of \(G\) which induces an \(m\)-clique. Let \(H\) be a component of \(G-S\), and let \(H S\) be the subgraph induced by \(V(H) \cup S\). Let \(K S\) be the subgraph \(G-V(H)\). We have: LEMMA 11.11 Let \(S\) be a separating set which induces an m-clique in G. Let HSand Ksbe defined as above. Then \(n(G, \lambda)=\Pi(H S, \lambda) \Pi(K S, \lambda) / \lambda(\lambda-1)(\lambda-2) \ldots(\lambda-m+1)\).
PROOF Every coloring of \(G\) induces a coloring of \(H S\) and \(K S\). There are \(n(H S, \lambda)\) ways to color \(H S\). There are \(n(K S, \lambda) / \lambda(\lambda-1)(\lambda-2) \ldots(\lambda-m+1)\) ways to complete a coloring of \(S\) to a coloring of \(K S\). This gives all colorings of \(G\).
There are no efficient means known of computing chromatic polynomials. This is due to the fact that most coloring problems are NP-complete. If \(\Pi(G, \lambda)\) could be efficiently computed, we would only need to evaluate it for \(\lambda=3\) to determine if \(G\) can be 3 -colored.
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\section*{Exercises}
11.5.1 Find \(n(C 2 n, \lambda)\) and \(n(C 2 n+1, \lambda)\).
11.5.2 Find \(n(G, \lambda)\), where \(G\) is the graph of the cube and the graph of the octahedron.
11.5.3 Let \(G^{\prime}\) be constructed from \(G\) by adding a new vertex joined to every vertex of \(G\). Determine \(n\left(G^{\prime}, \lambda\right)\) in terms of \(n(G, \lambda)\).
11.5.4 The wheel \(W n\) is obtained from the cycle Cn by adding a new vertex joined to every vertex of Cn. Find \(n(W n, \lambda)\).
11.5.5 A unicyclic graph \(G\) is a graph formed from a tree \(T\) by adding a single edge connecting two vertices of
T. \(G\) has exactly one cycle. Let \(G\) be a unicyclic graph on \(n\) vertices, such that the unique cycle of \(G\) has
length \(m\). Find \(n(G, \lambda)\).
11.5.6 Let \(G^{\prime}\) be constructed from \(G\) by adding two new adjacent vertices joined to every vertex of \(G\).

Determine \(\Pi\left(G^{\prime}, \lambda\right)\) in terms of \(n(G, \lambda)\).
11.5.7 Let \(G^{\prime}\) be constructed from \(G\) by adding \(k\) new mutually adjacent vertices joined to every vertex of \(G\). Determine \(\Pi\left(G^{\prime}, \lambda\right)\) in terms of \(\Pi(G, \lambda)\).
11.5.8 Let \(G^{\prime}\) be constructed from \(G\) by adding two new non-adjacent vertices joined to every vertex of \(G\). Determine \(\Pi\left(G^{\prime}, \lambda\right)\) in terms of \(\Pi(G, \lambda)\).
11.5.9 Find \(n(K m, m, \lambda)\).
11.5.10 Let \(G\) be a graph, and suppose that \(u v \notin E(G)\). Show that \(n(G, \lambda)=\Pi(G+u u, \lambda)+\pi((G+u u) \cdot u u, \lambda)\). When \(G\) has many edges, this is a faster way to compute \(n(G, \lambda, \lambda)\) than the method of Theorem 11.9.
11.5.11 Calculate \(n(K n-u u, \lambda)\), \(n(K n-u u-u w, \lambda)\), and \(n(K n-u u-w x, \lambda)\), where \(u, u, w, x\) are distinct vertices of \(K n\).
11.5.12 If \(G\) is a connected graph on \(n\) vertices, show that \(n(K n, \lambda) \leq n(G, \lambda) \leq \lambda(\lambda-1) n-1\), for all \(\lambda \geq 0\).
11.5.13 Prove that the coefficient of \(\lambda n\) in \(n(G, \lambda)\) is 1 , and that the coefficients alternate in sign.

\subsection*{11.6 Edge colorings}

A coloring of the edges of a graph \(G\) is an assignment of colors to the edges. More formally, an \(m\)-edgecoloring is a mapping from \(E(G)\) onto a set of \(m\) colors \(\{1,2, \ldots, m\}\). The coloring is proper if adjacent edges always have different colors. The edge-chromatic number or chromatic index of \(G\) is \(\chi^{\prime}(G)\), the minimum value of \(m\) such that \(G\) has a proper \(m\)-edge coloring. Notice that in any proper edge-coloring of \(G\), the edges of each color define a matching in \(G\). Thus, \(X^{\prime}(G)\) can be viewed as the minimum number of matchings into which \(E(G)\) can be partitioned.
When the edges of a multi-graph are colored, all edges with the same endpoints must have different colors.
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Consider a vertex \(u\) of maximum degree in \(G\). There must be DEG( \(u\) ) colors incident on \(u\). Therefore \(X^{\prime}\) \((G) \geq \Delta(G)\). There is a remarkable theorem by Vizing (Theorem 11.14) that states \(\chi^{\prime}(G) \leq \Delta(G)+1\) for simple graphs.

Before we come to the proof of Vizing's theorem, first consider the case when \(G\) is bipartite. A matching Mi saturating all vertices of degree \(\Delta(G)\) can be found with Algorithm 7.2.1 (the Hungarian algorithm). Alternating paths can be used to ensure that each vertex of degree \(\Delta(G)\) is saturated. Thus we know that the bipartite graph \(G\) has a maximum matching saturating every vertex of degree \(\Delta(G)\). This gives an algorithm for edgecoloring a bipartite graph.

\section*{Algorithm 11.6.1: BI PARTITECOLORING(G) \\ comment: Edge-color a graph \(G\) on \(n\) vertices \\ find the degrees of all vertices \\ \(i \leftarrow 1\) \\ repeat}
(find a maximum matching \(M_{i}\) in \(G\)
saturating every vertex of degree \(\Delta(G)\)
assign color \(i\) to the edges of \(M_{i}\)
\(G \leftarrow G-M_{i}\)
\(i \leftarrow i+1\)
until \(\Delta(G)=0\)
It follows from this algorithm, that when \(G\) is bipartite, \(x^{\prime}(G)=\Delta(G)\).
Suppose that \(G\) is a \(k\)-regular graph, edge-colored in \(k\) colors. Then every color occurs at every vertex of \(G\). If \(i\) and \(j\) are any two colors, then the ( \(i, j\) )-subgraph is the subgraph of \(G\) that contains only the edges of color \(i\) and \(j\). Because the ( \(i, j\) )-subgraph is the union of two matchings, it is the disjoint union of alternating cycles.
Let \(U \subseteq V(G)\). Let ni and nj denote the number of edges of colors \(i\) and \(j\), respectively, in the edge-cut [ \(U\), \(V-U\) ]. Each cycle of the ( \(i, j\) )-subgraph intersects [ \(U, V-U\) ] in an even number of edges. Thus we conclude that \(n i+n j\) is even. Therefore \(n i=n j(\bmod 2)\). This gives the following parity lemma:
LEMMA 11.12 (Parity lemma) Let \(G\) be a \(k\)-regular graph, edge-colored in colors \(\{1,2, \ldots, k\}\). Let \(U \subseteq V(G)\). Let ni denote the number of edges of color \(i\) in \([U, V-U]\). Then \(n 1 \equiv n 2 \equiv \ldots \equiv n k(\bmod 2)\).
Vizing's theorem (Theorem 11.14) is based on an algorithm to edge-color a graph in \(\leq \Delta(G)+1\) colors. The proof presented is based on that of
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FOURNIER [42]. It begins with an arbitrary coloring of \(G\) in \(\Delta(G)+1\) colors, and then gradually improves it until it becomes a proper coloring. Given a coloring, let \(c(u)\) denote the number of colors occurring at vertex \(v \in V(G)\). If \(c(u)=\operatorname{DEG}(u)\), for all \(u\), then the coloring is proper. Otherwise \(c(u)<\operatorname{DEG}(u)\), for some \(u\). The sum \(\Sigma u c(U)\) in an indication of how close an arbitrary coloring is to being a proper coloring.
Suppose first that \(G\) is arbitrarily colored in two colors.
LEMMA 11.13 If \(G\) is a graph that is not an odd cycle, then \(G\) has a 2-edge-coloring in which \(c(u) \geq 2\), for all vertices \(u\), with \(D E G(U) \geq 2\).
PROOF If \(G\) is Eulerian, choose an Euler tour, and color the edges alternately blue and red along the tour. If \(G\) is not Eulerian, add a new vertex \(u 0\) adjacent to every odd degree vertex of \(G\). The result is an Eulerian graph. Color it in the same way.
We can use Lemma 11.13 on subgraphs of \(G\). Given a proper coloring of \(G\), the edges of colors \(i\) and \(j\) each define a matching in \(G\). Consider the ( \(i, j\) )-subgraph. Each connected component is a path or an even cycle whose colors alternate. If, however, we begin with an arbitrary coloring of \(G\), then we want to maximize \(\Sigma u c(u)\). If some component of the \((i, j)\)-subgraph is not an odd cycle then by Lemma 11.13, it can be 2colored so that \(c(u) \geq 2\), for all vertices \(u\), with \(D E G(u) \geq 2\).
THEOREM 11.14 (Vizing's theorem) If \(G\) is simple, then \(\chi^{\prime}(G) \leq \Delta(G)+1\).
PROOF We begin by arbitrarily coloring the edges in \(\Delta(G)+1\) colors. We show that when \(\Sigma u c(u)\) is as large as possible, the coloring must be proper. Suppose that the coloring is not proper, and choose a vertex \(u\) with \(c(u)<D E G(u)\). Some color io is missing at \(u\), and some color il occurs at least twice. Let edges uu0 and uu1 have color \(i 1\). If color \(i 0\) is missing at either \(U 0\) or \(u 1\), we can recolor one of these edges with color \(i 0\), thereby increasing \(\Sigma u c(u)\). Hence, we can assume that color i0 occurs at both \(u 0\) and \(u 1\). Some color is missing at \(u 1\); call it \(i 2\). If \(i 2\) is also missing at \(u\), we can recolor \(u u 1\) with color \(i 2\), thereby increasing \(\Sigma u c(u)\). Hence, let \(u u 2\) be an edge of color i2. Some color is missing at \(u 2\); call it i3. If i3 is also missing at \(u\), we can recolor \(u u 1\) with color \(i 2\), and \(u u 2\) with color is, thereby increasing \(\Sigma u c(u)\). It follows that \(i 3 \neq i 0\), so that i0 occurs at \(u 2\). We continue in this way, constructing a sequence of edges \(u u 1, u U 2, \ldots, u u k\) of distinct colors i1, \(i 2 \ldots, i k\), such that color io occurs at each of \(u 1, \ldots, u k\), and color \(i j+1\) does not occur at \(u j\). Refer to Figure 11.4. We continue in this fashion generating a sequence io, i1,..., of distinct colors until we find a color \(i k+1\)

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\section*{FI GURE 11.4}

\section*{Vizing's theorem}
is not missing at \(u\). Thus \(i k+1\) has previously occurred in the sequence. Suppose that \(i k+1=i \ell\), where \(1 \leq \ell \leq k\). Recolor the edges \(u \cup 1, u \cup 2, \ldots, u v \ell-1\) with the colors \(\ell 2, \ell 3, \ldots\), il, respectively. This does not change \(\sum u c(u)\), because each \(c(u j)\) is unchanged. Notice that \(u u \ell-1\) and \(u u \ell\) are now both colored \(i \ell\). Consider the ( \(i 0, i \ell)\) subgraph containing \(u\), It contains \(u \ell-1\) and \(u \ell\). If it is not an odd cycle, it can be recolored so that colors i0 and if both occur at each of the vertices that have degree exceeding one. This would increase \(c(u)\) and therefore \(\Sigma u c(u)\). Hence this subgraph can only be an odd cycle, so that \(G\) contains an ( \(i 0, i \ell\) ) -path from \(u \ell-1\) to \(u\) l.
Now recolor the edges \(u u \ell, u u \ell+1, \ldots, u u k\) with the colors \(i \ell+1, \ldots, i k, i k+1=i \ell\), respectively. Once again \(\Sigma u c(u)\) is unchanged. The ( \(i 0, i \ell\) )-subgraph containing \(u\) now contains \(u \ell-1\) and \(u k\), and must be an odd cycle. Hence \(G\) contains an ( \(i 0, i \ell\) )-path from \(u \ell-1\) to \(u k\). This contradicts the previous path found. It follows that the coloring can only be proper, and that \(x^{\prime} \leq \Delta(G)+1\).
The proof of Vizing's theorem makes use of a color rotation at vertex \(u\), namely, given a sequence of incident edges \(u u 1, u u 2, \ldots, u u k\) of colors \(i 1, i 2, \ldots, i k\), respectively, such that \(u j\) is missing color \(i j+1\), for \(j=1,2, \ldots, k-1\), and \(u k\) is missing color il where \(\ell \in\{1,2, \ldots, k-1\}\). We then recolor \(u u 1, u u 2, \ldots, u u \ell-1\) with colors \(i 2, i 3\), ..., il, respectively. This technique will be used in the edge-coloring algorithm given as Algorithm 11.6.2. The algorithm will build a sequence of vertices \(u 1, U 2, \cup 3, \ldots\) adjacent to \(u\), searching for a color rotation. Graphs \(G\) for which \(x^{\prime}(G)=\Delta(G)\) are said to be Class I graphs. If \(X^{\prime}(G)=\Delta(G)+1\), then \(G\) is a Class II graph. Although there is an efficient algorithm
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to color a graph in at most \(\Delta(G)+1\) colors, it is an NP-complete problem to determine whether an arbitrary graph is of Class I or II. A proof of this remarkable result is presented at the end of the chapter.
If \(G\) is a multigraph with no loops, then the general form of Vizing's theorem states that \(\chi^{\prime}(G) \leq \Delta(G)+\mu(G)\), where \(\mu(G)\) is the maximum edge-multiplicity. It can often happen that \(\mu(G)\) is not known. Shannon's
theorem gives an alternative bound \(\chi^{\prime}(G) \leq\lfloor 3 \Delta(G) / 2\rfloor\). The proof presented is due to ORE [93]. It requires two lemmas. Given a proper edge-coloring of \(G\), we write \(C(u)\) for the set of colors present at \(u\).
LEMMA 11.15 (Uncolored edge lemma) Let \(G\) be a multigraph without loops. Let uu be any edge of \(G\), and let \(G\)-uu be edge-colored with \(k\) colors, and suppose that \(\chi^{\prime}(G)=k+1\). Then:
\[
\begin{gathered}
|C(u) \cup C(v)|=k \\
|C(u) \cap C(v)|=\operatorname{DEG}(u)+\operatorname{DEG}(v)-k+2 \\
|C(u)-C(v)|=k-\operatorname{DEG}(v)+1 \\
|C(v)-C(u)|=k-\operatorname{DEG}(u)+1
\end{gathered}
\]

PROOF Every color missing at \(u\) is present at \(u\), or there would be a color available for \(u u\), thereby making \(X^{\prime}\) \((G)=k\). Therefore all colors are present at one of \(u\) or \(u\). This gives the first equation. The colors present at \(u\) can be counted as
\[
|C(u)-C(v)|+|C(u) \cap C(u)|=\operatorname{DEG}(u)-1 .
\]

Similarly, those present at \(u\) are given by
\[
|C(u)-C(u)|+|C(u) \cap C(v)|=\operatorname{DEG}(u)-1 .
\]

Now
\[
|C(u) \cup C(v)|=|C(u)-C(v)|+|C(v)-C(u)|+|C(u) \cap C(v)|
\]
so that we can solve these equations for \(|C(u) \cap C(u)|\), giving the second equation. The third and fourth equations then result from combining this with the previous two equations.
LEMMA 11.16 (Ore's lemma) Let G be a multigraph without loops. Then:
\[
\chi^{\prime}(G) \leq \operatorname{MAX}\left\{\Delta(G), \frac{1}{2} \operatorname{MAX}_{\{u, v, w\}}\{\operatorname{DEG}(u)+\operatorname{DEG}(v)+\operatorname{DEG}(w)\}\right\}
\]
where the second maximum is over all triples of vertices \(u, u, w\) such that \(v \longrightarrow u \longrightarrow w\).
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PROOF The proof is by induction on \(\varepsilon(G)\). It is clearly true if \(\varepsilon(G) \leq 3\). Suppose it holds for all graph with \(\varepsilon(G) \leq m\) and consider \(G\) with \(\varepsilon(G)=m+1\). Let \(u u\) be any edge. Delete \(u u\) to obtain \(G-u u\), for which the result holds. Let \(X^{\prime}(G-u u)=k\), and consider a proper \(k\)-edge coloring of \(G-u u\). If there is a color available for \(u u\), then \(\chi^{\prime}(G)=k\), and the result holds. Otherwise the uncolored edge lemma (Lemma 11.15) applies to \(G\). Pick a color \(i \in C(u)-C(v)\) and an edge uw of color \(i\). We first show that \(C(v)-C(u) \subseteq C(w)\). Let \(j\) \(j \in C(v)-C(u)\). If color \(j\) is missing at \(w\), then since \(j\) is also missing at \(u\), we can recolor edge \(u w\) with color \(i\), and assign color \(i\) to \(u u\). This results in a \(k\)-edge-coloring of \(G\), a contradiction. Therefore \(j \in C(w)\), so that \(C(v)-C(u) \subseteq C(w)\).
We also show that \(C(u)-C(v) \subseteq C(w)\). We know that \(i \in C(u)-C(v)\). If there is no other color in \(C(u)-C(u)\), we are done. Otherwise, pick also \(\ell \in C(u)-C(v)\). If \(\ell \notin C(w)\), consider the ( \(\ell, j\) )-subgraph \(H\) containing \(u\). If \(H\) does not contain \(u\), we can interchange colors in \(H\), and assign color \(\ell\) to \(u U\), a contradiction. Therefore \(H\) consists of an alternating path from \(u\) to \(u\). Interchange colors in \(H\), recolor edge \(u w\) with color \(\ell\), and assign color \(i\) to edge \(u u\), again a contradiction. We conclude that \(C(u)-C(v) \subseteq C(w)\). We have \(|C(u)-C(u)|+|C(u)-C(u)| \leq|C(w)|\). By the uncolored edge lemma, this means that
DEG \((u) \geq 2 k-\operatorname{DEG}(u)-\operatorname{DEG}(u)+2\). Therefore DEG(u)+DEG(u)+DEG(w) \(\geq 2 k+2\), so that
\(k+1 \leq \frac{1}{2}\left(\operatorname{DEG}(u)+\operatorname{DEG}(v) \sum_{-} \operatorname{DEG}(w)\right)\). The result then holds for \(G\), as required.
THEOREM 11.17 (Shannon's theorem) Let \(G\) be a multigraph without loops. Then \(\chi^{\prime}(G) \leq\lfloor 3 \Delta(G) / 2\rfloor\). PROOF By Ore's lemma:
\[
\chi^{\prime}(G) \leq \operatorname{Max}\left\{\Delta(G), \frac{1}{2}\{\Delta(G)+\Delta(G)+\Delta(G)\}\right\}=3 \Delta(G) / 2
\]

Vizing's theorem results in an algorithm to color the edges of \(G\) in at most \(\Delta(G)+1\) colors. Let the vertices of \(G\) be numbered \(1,2, \ldots, n\), and let the colors be \(\{0,1, \ldots, \Delta(G)\}\). There are two somewhat different ways in which this can be programmed. In both methods, we begin by assigning color 0 to every edge. We then take each vertex \(u\) in turn and gradually improve the coloring local to \(u\). Both methods involve constructing color rotations and alternating paths, as can be expected from the proof of Vizing's theorem.
One method uses alternating paths to improve the coloring local to \(u\). When this does not succeed, a color rotation is used. The second method uses color

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rotations to improve the coloring local to \(u\), and builds alternating paths when the color rotation does not succeed. We present the first method here. The algorithm keeps the adjacency lists in partially sorted orderall edges of any color \(i\) are consecutive, and any colors which occur only once at vertex \(u\) are stored at the end of the adjacency list for \(u\). This is certainly true initially when all edges have color 0 . This allows the algorithm to easily determine whether there is a repeated color at \(u\) by simply taking the first two adjacent vertices.

Algorithm 11.6.2: \(\operatorname{EDGECOLOR(G,n)}\)
comment: \(\left\{\begin{array}{l}\text { Initially the edges all have color } 0 . \\ \text { Convert the coloring to a proper coloring in } \\ \text { at most } \Delta(G)+1 \text { colors. }\end{array}\right.\)
\[
\begin{gathered}
\text { for } u \leftarrow 1 \text { to } n \\
\text { do }\left\{\begin{array}{l}
v \leftarrow \text { 1st vertex joined to } u \\
w \leftarrow \text { 2nd vertex joined to } u \\
\text { while Color }[u v]=\text { Color }[u w] \\
\text { do }\left\{\begin{array}{l}
\text { comment: } u \text { has a repeated color } \\
\text { CoLORBFS }(u) \\
v \leftarrow 1 \text { st vertex joined to } u \\
w \leftarrow 2 \text { nd vertex joined to } u
\end{array}\right.
\end{array}\right.
\end{gathered}
\]

Algorithm 11.6.2 takes the vertices in order \(1,2, \ldots, n\). When vertex \(u\) is considered, it assumes that the subgraph induced by \(\{1,2, \ldots, u-1\}\) has been properly colored in at most \(\Delta(G)+1\) colors. This is certainly true when \(u=1\). If the first two incident edges have the same color \(k\), it attempts to recolor an incident edge of color \(k\) as follows. Let \(i\) be a color missing at vertex \(u\).
1. If any edge \(u v\) of color \(k\) is such that vertex \(u\) is also missing color \(i\), then \(u u\) is assigned color \(i\).
2. Or if any edge \(u u\) of color \(k\) is such that \(u>u\), then \(u u\) is assigned color \(i\). This improves the coloring at \(u\), and does not affect the subgraph induced by \(\{1,2, \ldots, u-1\}\).
3. Otherwise all incident edges \(u v\) of color \(k\) are such that \(u<u\), and \(u\) has exactly one incident edge of color \(i\). The algorithm constructs paths from \(u\) whose edges alternate colors \(k\) and \(i\), using a breadth-first search.
4. If a path \(P\) to a vertex \(u\) is found, such that \(u\) is missing either color \(i\) or \(k\), then the colors on the path are reversed.
5. Or if a path \(P\) to a vertex \(u\) is found, such that \(u>u\), then the colors on the path are reversed. In each of these cases, the coloring is improved at
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vertex \(u\), and a valid coloring of the subgraph induced by \(\{1,2, \ldots, u-1\}\) is maintained.
6. Otherwise, each alternating path of colors \(i\) and \(k\) induces an odd cycle containing \(u\). There must be an even number of incident edges of color \(k\). If there are more than two such edges, let uu be one of them. Some color \(j\) is missing at \(u\). We recolor edge \(u u\) with color \(j\), and reverse the colors on the remainder of the odd cycle containing \(u u\). This improves the coloring at \(u\), and maintains a valid coloring in the subgraph induced by \(\{1,2, . ., u-1\}\).
7. Otherwise, there are exactly two incident edges of color \(k\). The algorithm searches for a color rotation-it constructs a sequence of incident edges uu1, uu2,..., uum of colors \(i=k, i 2, \ldots, i m\), respectively, such that \(u j\) is missing color \(i j+1\), for \(j=1,2, \ldots, m-1\), and \(u m\) is missing color il where \(\ell \in\{1,2, \ldots, m-1\}\). If color il occurs an even number of times at \(u\), then the edges \(u u 1, u u 2, \ldots, u u \ell-1\) are recolored \(i 2, i 3, \ldots, i \ell\). This creates an odd number of edges of color if incident at \(u\). The edges of color il are moved to the front of the adjacency list of \(u\). The next iteration of the algorithm is certain to improve the coloring at \(u\).
8. Otherwise, color il occurs an odd number of times. If it occurs at least three times, the edges of color il are moved to the front of the adjacency list of \(u\). The next iteration of the algorithm is certain to improve the coloring at \(u\).
9. Otherwise, color if occurs exactly once at \(u\). The algorithm builds an alternating path \(P\) of colors \(i\) and il from um. If \(P\) leads to \(u \ell\), we recolor the edges \(u u 1, u u 2, \ldots, u \cup \ell-1\) with colors \(i 2, i 3, \ldots\), il and move the edges of color if to the head of the adjacency list of \(u\). If \(P\) does not lead to \(u \ell\), we recolor the edges \(u u 1\), \(u u 2, \ldots\), u \(u m\) with colors \(i 2, i 3, \ldots, i m\), il, and move the edges of color il to the head of the adjacency list of \(u\). In either case, the next call to ColorBFS is guaranteed to improve the coloring at \(u\).
Thus, it is always possible to improve the coloring at \(u\). Each time that an edge \(u w\) is recolored, it is necessary to adjust the order of the adjacency lists of \(u\) and \(w\) to ensure that they remain partially sorted. If \(u<u\), we know that every color incident on \(u\) occurs exactly once at \(u\). But if \(u>u\), it will be necessary to change some links in the linked list. After the coloring has been completed for vertex \(u=n\), the entire graph has a valid coloring.
The algorithm uses a global array ScanQ for breadth-first searches. The number of vertices currently on the ScanQ is denoted by QSize. The ScanQ accumulates vertices that form alternating paths. The previous vertex to \(u\) in such a path is denoted \(\operatorname{PrevPt[U].~We~set~} \operatorname{PrevPt[U]\leftarrow 0~initially.~When~it~is~necessary~to~reinitialize~the~}\) ScanQ, the entire array is not reset-only the vertices \(u\) currently on the ScanQ have PrevPt[ \(u\) ] reset to 0 , and QSize reset to 0 . The color of edge \(u v\) is indicated by Color \(\langle u v\rangle\) in the pseudo-code, but in an actual program it would
be stored as a field in the adjacency list structure of the graph. The frequency of color \(i\) at a given vertex is stored in an array as Freq[i].

Algorithm 11.6.3: COLORBFS(u)
comment: 2 or more incident edges at \(u\) have the same color-recolor one
\(i \leftarrow a\) missing color at \(u\)
\(u \leftarrow\) first vertex joined to \(u\) \(k \leftarrow\) Color \(\langle u v\rangle\)
comment: initialize the \(\operatorname{ScanQ} Q\) with all adjacent vertices of color \(k\)
QSize \(\leftarrow 0\)
while \(\operatorname{Color}\langle u v\rangle=k\)
do \(\left\{\begin{array}{l}\text { if (color } i \text { is missing at } v) \text { or }(v>u) \\ \text { then }\left\{\begin{array}{l}\text { RECOLOREDGE }(u, v, i) \\ \text { re-initialize the } \operatorname{ScanQ} \text { and PrevPt arrays } \\ \text { return }\end{array}\right. \\ \text { QSize } \leftarrow \text { QSize }+1 \\ \text { ScanQ }[Q S i z e] \leftarrow v \\ \operatorname{PrevPt}[v] \leftarrow u \\ v \leftarrow \text { next vertex adjacent to } u\end{array}\right.\)
comment: \(\left\{\begin{array}{l}\text { each } v \text { in } S c a n Q \text { satisfies Color }\langle u v\rangle=k \text { and } v<u \\ \text { build paths whose edges are alternately colored } i \text { and } k .\end{array}\right.\)
\[
\begin{aligned}
& w \leftarrow \text { ALTERNATINGPATH }(u, i, k) \\
& \text { if } w \neq 0
\end{aligned}
\]
then \(\left\{\begin{array}{l}\operatorname{RECOLORPATH}(w, i, k) \\ \text { return }\end{array}\right.\)
comment: \(\left\{\begin{array}{l}\text { it was not possible to re-color any alternating path. } \\ \text { the vertices on ScanQ induce odd cycles of colors } i \text { and } k .\end{array}\right.\)
\(U \leftarrow S c a n Q[1]\)
if Freq[k]>2
then \(\left\{\begin{array}{l}j \leftarrow \text { a color missing at } v \\ \operatorname{RECOLOREDGE}(u, v, j) \\ \operatorname{RecolORCYCLE}(v, i, k) \\ \text { return }\end{array}\right.\)
comment: \(\left\{\begin{array}{l}\text { if this point is reached, it is necessary } \\ \text { to search for a color rotation }\end{array}\right.\)
COLORROTATION( \(u, i\) )
The algorithm uses two procedures RECOLOREDGE() and RECOLORPATH() to recolor an edge or alternating path. They are described as follows:
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procedure RECOLOREDGE \((u, u, j)\)
comment: Recolor edge uu with color \(j\).
Color \([u v] \leftarrow j\)
adjust the order of the adjacency lists of \(u\) and \(u\) as necessary procedure RECOLORPATH( \(w, i 1, i 2\) )
comment: \(\left\{\begin{array}{l}\text { Recolor the edges of an alternating path beginning at } w \\ \text { The edges currently alternate colors } i_{1} \text { and } i_{2} .\end{array}\right.\)
\(U \leftarrow\) PrevPt[ \(w\) ]
while \(w \neq 0\)
do \(\left\{\begin{array}{l}j \leftarrow i_{1}+i_{2}-\operatorname{Color}[v w] \\ \operatorname{RECOLOREDGE}(v, w, j) \\ w \leftarrow v \\ v \leftarrow \operatorname{PrevPt}[w]\end{array}\right.\)
re-initialize the ScanQ and PrevPt arrays
ALTERNATINGPATH ( \(u, i 1, i 2\) ) is a procedure which builds alternating paths of colors i1 and i2 beginning from the vertices on the ScanQ. When it is called, the vertices \(u\) currently on the ScanQ are all known to have an incident edge of color i1. If uw has color i1, and \(w\) has no incident edge of color \(i 2\), or if \(w>u\), then vertex \(w\) is returned, so that the colors on the path from \(w\) to \(u\) can be reversed. This improves the coloring at \(u\), and maintains the coloring on the subgraph induced by \(\{1,2, \ldots, u-1\}\).
When a suitable alternating path to a vertex \(w\) is found, it can be traced by following the PrevPt[•] values. When an edge uw is recolored, it may be necessary to adjust the ordering of the adjacency lists of \(u\) and \(w\) so that the partial order is maintained.
COLORROTATION ( \(u, i\) ) searches for a color rotation from a vertex \(u\) which is missing color \(i\) by building a sequence of vertices \(u 1, u 2, \ldots, u m\) on the \(S c a n Q\) such that each edge \(u u j\) has color \(i j\) and \(u j\) is missing color \(i j+1\). The pseudocode for it is left to the reader.
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Algorithm 11.6.4: ALTERNATINGPATH( \(u, i 1, i 2\) )
comment: \(\left\{\begin{array}{l}\text { Build alternating paths of colors } i_{1} \text { and } i_{2}, \\ \text { beginning from vertices currently on the ScanQ. } \\ \text { Return the first } w \text { encountered which is missing color } i_{1} \text { or } i_{2} \\ \text { or which satisfies } w>u, \text { if there is such a } w . \\ \text { NewQSize } \leftarrow \text { QSize } \\ j \leftarrow 1 \\ \text { while } j \leq \text { QSize }\end{array}\right.\)
(comment: extend alternating paths using edges of color \(i_{1}\)
repeat
\(\left(v \leftarrow S c a n Q[j] \quad\right.\) " \(j^{\text {th }}\) vertex on \(S c a n Q "\)
\(w \leftarrow\) a vertex of color \(i_{1}\) adjacent to \(v\)
if (color \(i_{2}\) is missing at \(w\) ) or \(w<u\)
then return \((w)\)
do
if \(w \notin \operatorname{ScanQ}\)
then \(\left\{\begin{array}{l}\text { NewQSize } \leftarrow \text { NewQSize }+1 \\ \text { ScanQ }[\text { NewQSize }] \leftarrow w \\ \operatorname{PrevPt}[w] \leftarrow v\end{array}\right.\)
\(j \leftarrow j+1\)
until \(j>\) QSize
QSize \(\leftarrow\) NewQSize
\(\operatorname{swap}\left(i_{1}, i_{2}\right)\)
return (0) "no suitable alternating path was found"

\subsection*{11.6.1 Complexity}

It is somewhat difficult to organize the data structures to allow an efficient implementation of this algorithm. The reason is that the program often needs to find an edge \(u u\) incident on \(u\) of a given color \(i\), or to determine whether \(u\) is missing color \(i\). This can be done in constant time if an \(n \times n\) array is stored for which entry \([u, i]\) is a pointer to a linked list of all incident edges of color \(i\). This array will be quite sparse for most graphs. With this data structure, the algorithm will be quite efficient.
The main program takes each vertex \(u\) in turn. If there is a repeated color at \(u\), COLORBFS( \(u\) ) is called. It will usually succeed in recoloring one of the incident edges at \(u\), taking only a constant number of steps, so that \(O(D E G(u))\) steps are usually needed to make \(c(u)=\operatorname{DEG}(u)\). But as more and more of the graph is colored, it becomes necessary to construct alternating paths to improve the coloring at \(u\). Let \(G(u)\) denote the subgraph induced by \(\{1,2, \ldots, u\}\). The

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alternating paths constructed are contained in \(G(u)\). A breadth-first search to construct the alternating paths takes \(O(\varepsilon(G(u)))\) steps, provided that the edge of color \(i\) incident on a vertex \(u\) can be found in a constant number of steps.
Since \(\varepsilon(G(u)) \leq \varepsilon(G)\), we determine that if a color rotation is not required, that COLORBFS(u) will take at most \(O(\varepsilon)\) steps. If a color rotation is constructed, it requires at most \(O(\mathrm{DEG}(u))\) steps to build the sequence \(u u 1\), \(u u 2, \ldots\), u \(u m\) of colors \(i=k, i 2, \ldots, i m\). The maximum number of steps are executed when an alternating path must be constructed from um, taking \(O(\varepsilon(G(u)))\) steps. If COLORBFS( \(u\) ) does not succeed in improving the coloring at \(u\) on a given call, it is guaranteed to succeed on the following call. We conclude that it takes at most \(O(\operatorname{DEG}(u) \cdot \varepsilon)\) steps to make \(c(u)=\operatorname{DEG}(u)\). The total complexity is therefore at most \(O(\varepsilon 2)\).

\section*{Exercises}
11.6.1 Describe an algorithm using alternating paths in a bipartite graph which finds a maximum matching saturating all vertices of degree \(\Delta(G)\).
11.6.2 Work out the complexity of the bipartite coloring algorithm.
11.6.3 Program the bipartite edge-coloring algorithm.
11.6.4 Show that an edge coloring of \(G\) gives a vertex coloring of \(L(G)\).
11.6.5 Determine \(x^{\prime}\) for the Petersen graph.
11.6.6 Show that when the inverter shown in Figure 11.6 is edge-colored in three colors, one of the two pairs of edges \(\{a, b\},\{c, d\}\) has the same color, and the other pair has different colors.

\subsection*{11.7 NP-completeness}

In this section we show that several coloring-related problems are NP-complete.
Problem 11.1: 3-Colorability
Instance: a graph \(G\).
Question: is \(x(G) \leq 3\) ?
We transform from 3-Sat. Consider an instance of 3-Sat containing variables \(u 1, u 2, \ldots\), un, with complements \(\bar{u} 1, \bar{u} 2, \ldots, \bar{u} n\). Suppose that there are \(m\) clauses, denoted \(c 1, c 2, \ldots, c m\). Construct a graph \(G\) as follows. \(G\) will have vertices \(u 1, u 2, \ldots, u n, \bar{u} 1, \bar{u} 2, \ldots, \bar{u} n, x 1, x 2, \ldots, x n, \bar{x}_{1}, \bar{x}_{2}, \ldots, \bar{x}_{n}\), and

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\(c 1, c 2, \ldots, c m\). Each ui is joined to \(\bar{u} i\) and to \(x i\). Each \(\bar{u} i\) is joined to \(\bar{x}_{i}\). There will also be triangles \(T 1, T 2, \ldots\), \(T m\) where Ti corresponds to clause \(c i\), and vertices denoted \(B\) and \(W\) (blue and white). Each ui and \(\bar{u} i\) is joined to \(W\). Each ci is joined to \(B\). A clause like \(c 1=(u 1+\bar{u} 2+\bar{u} 3)\) has the corresponding vertex \(c 1\) joined to \(x 1, \bar{x}_{2}\), and \(\bar{x}_{3}\). The triangle \(T 1\) corresponding to \(c 1\) has its three corners joined to the same vertices as \(c 1\). The construction is illustrated in Figure 11.5. We will color \(G\) in colors red, white, and blue.


\section*{FI GURE 11.5}

\section*{Graph 3-Colorability is NP-complete}

Suppose first that \(G\) has a proper 3 -coloring. Without loss of generality, we can assume that vertex \(B\) is colored blue, and that vertex \(W\) is colored white. It follows that all vertices ci are colored red. Each triangle Ti must use all three colors, so that Ti has a vertex colored white, which is adjacent to some \(x j\) or \(\bar{x}_{j}\), which must be colored blue (since ci is also adjacent, and it is red). The adjacent uj (or \(\overline{u j}\) ) can then only be colored red. Only one of \(u j\) and \(\bar{u} j\) can be red-the other must be blue, with corresponding \(x j\) or \(\bar{x}_{j}\) white. It follows that every clause ci contains a literal uj or \(\bar{u} j\) that is colored red. Take these as the variables assigned the value true. The result is an assignment of values satisfying every clause.
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Conversely, if there is an assignment of values satisfying every clause, color each \(u j\) or \(\bar{u} j\) red if its value is true, and color its complement blue. If \(u j\) is red, color the corresponding \(x j\) blue; otherwise color it white. The corner of Ti corresponding to \(u j\) or \(\bar{u} j\) is colored white. The other corners of \(T i\) can be red or blue. The result is a 3 -coloring of \(G\).
We conclude that a polynomial algorithm which solves 3-Colorability could also be used to solve 3-Sat. Since
3-Sat is NP-complete, so is 3-Colorability.
Problem 11.2: Clique
Instance: a graph \(G\) and an integer \(k\).
Question: does \(G\) have a clique with \(\geq k\) vertices (i.e., is \(\bar{\alpha}(G) \geq k\) )?
We transform from Problem 9.5.9.4 Vertex Cover. Recall that a vertex cover in a graph \(G\) is a subset
\(U \subseteq V(G)\) such that every edge has at least one endpoint in \(U\). Given an integer \(k\) the Vertex Cover problem asks: does \(G\) have a vertex cover with \(<k\) vertices?
If \(U\) is a vertex cover, the set \(\bar{U}=V(G)-U\) is an independent set in \(G\). Hence \(\bar{U}\) induces a clique in \(\bar{G}\). If \(G\) has \(n\) vertices, and \(U\) has \(\geq m\) vertices, then \(U\) is a vertex cover with \(\leq n-m\) vertices. Thus, given an instance of the Vertex Cover problem, we construct \(\bar{G}\), and ask whether it has a clique with at least \(n-m\) vertices. If the anwer is yes, then \(G\) has a vertex cover with at most \(m\) vertices. We conclude that a polynomial algorithm which solves Clique could also be used to solve Vertex Cover. It follows that Clique is NP-complete.
Problem 11.3: Chromatic Index I nstance: a graph \(G\).

\section*{Question: is \(X^{\prime}(G)=\Delta(G)\) ?}

There is an ingenious construction of Holyer proving that it is NP-complete to determine whether a 3-regular graph is of Class I or II. Holyer's construction is based on the graph shown in Figure 11.6, which he calls an inverting component, or inverter. The inverter was originally discovered by Loupekine. It consists of two 5cycles sharing two common consecutive edges. Edges \(a, b, c, d\), e are attached to five of the vertices. A 5-cycle requires at least three colors. Consider any 3-edge-coloring of the inverter. Apply the parity condition (Lemma 11.13) to the set of seven vertices shown in the diagram. We determine that in any 3-edge coloring, that three of the five edges \(a, b, c, d\), e must have the same color, and that the other two have distinct colors, since 5 can only be written as \(3+1+1\), as a sum of three odd numbers. We then find that \(a, e\), and \(c\) cannot all have the same color, so that
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\section*{FI GURE 11.6}

\section*{The inverter and its schematic representation (Used with permission of the SI AM J ournal of Computing)}
in any 3-edge-coloring, either \(a\) and \(b\) have the same color, and \(c, d\), e have three distinct colors; or by symmetry, \(c\) and \(d\) have the same color, and \(a, b\), e have three distinct colors. The inverter is represented schematically in Figure 11.6, as a circle with two pairs of inputs (or outputs) and a fifth input (or output). Holyer transforms from 3-Sat to Chromatic Index. Consider an instance of 3-Sat with clauses c1, c2,.., cm involving variables \(u 1, u 2, \ldots\), un and their complements \(\bar{u} i\) Each variable is either true or false. A value of true is represented in the edge coloring of an inverter, as an input pair of the same color. A value of false is represented as an input pair of different colors. In every 3-edge-coloring of the inverter, a value of true is inverted to a value of false, or vice versa.
A clause of three variables, for example, \((u i+u j+\bar{u} k)\), is represented by three inverters tied together by a 7cycle, as shown in Figure 11.7. Note the edges marked \(a, b, x, y\). It is easy to see that \(a\) and \(b\) have the same color if and only if \(x\) and \(y\) have the same color. Because of the 7-cycle, at least one of the three inverters must have two outside inputs the same color. This will be used to indicate that at least one of the variables \(u i, u j\), and \(\bar{u} k\) in the above clause must have a value of true (i.e., every clause will be satisfied).
Now a given variable ui, and/or its complement \(\bar{u} i\) will appear in several clauses. It must have the same value in all clauses. In order to ensure this, the 7-cycles corresponding to clauses containing ui or \(u \bar{i}\) are also tied together. Holyer constructs a cycle of inverters, with two inverters for each occurrence of ui or ūi in the clauses. For example, if ui and \(\bar{u} i\) occur a total of four time in clauses c1, c2, c3, c4, the cycle of Figure 11.8 is constructed. Each pair of inverters corresponds to a clause containing ui or \(\bar{i}\) i. Notice that there are a total of six inputs connecting a pair of inverters to the rest of the graph. By the parity lemma, if this graph is 3-edge-colored, the possible color combinations for the six inputs are \(6+0+0,4+2+0\), or \(2+2+2\). If one pair of external inputs represents
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FI GURE 11.7
The representation of a clause \(c \mathbf{1}=(u i+u j+u \bar{k})\) of 3-Sat (Used with permission of the SI AM J ournal of Computing)
true, then by the properties of an inverter, the opposite pair also represents true. Then the parity lemma implies that the remaining two inputs also represent true. It follows that all pairs of inverters in the diagram have all external input pairs representing true. Consequently, if any one pair represents false, they all represent false. This mechanism is used to guarantee that ui has the same value in all clauses.
We now put these ideas together. We are given an instance of 3-Sat. For each clause cj, three inverters are tied together using a 7-cycle, as in Figure 11.7. For each variable ui, a cycle of pairs of inverters is constructed as in Figure 11.8. The input corresponding to \(c j\) in this structure is connected to the input corresponding to ui in the 7 -cycle component corresponding to cj . If the clause contains \(u \bar{i}\) rather than ui, then another inverter is placed between the two before connecting. The result is a graph containing a subgraph for each ui and for each Cj , tied together through their inputs. There are still a number of inputs not connected to anything. In order to complete the construction and have a 3-regular graph, a second copy of this graph is constructed. The corresponding unused inputs of the two copies are connected together. The result is a 3 -regular graph \(G\). In any 3 -edge-coloring of \(G\), every clause is guaranteed to have at least one variable representing true. All occurrences of each variable are guaranteed to have the same value. The result is an assignment of values to the variables solving the 3 -Sat instance. We conclude that if we could determine whether \(\chi^{\prime}(G)=3\), we could solve 3 -Sat. Since
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FI GURE 11.8

\section*{A pair of inverters for each clause containing uior ūi(Used with permission of the SI AM Journal of Computing)}

3-Sat is NP-complete, we conclude that Chromatic Index is also NP-complete.

\subsection*{11.8 Notes}

The DEGREESATURATION() algorithm is from BRELAZ [20]. A very good paper on the limitations of the sequential algorithm is JOHNSON [69]. A very readable survey of chromatic polynomials appears in READ and TUTTE [101]. See also TUTTE [119], where the word chromial is coined for chromatic polynomial. The proofs of the uncolored edge lemma and Ore's lemma (Lemmas 11.15 and 11.16) are based on those of BERGE [14]. A very efficient edge-coloring algorithm based on Vizing's theorem was developed by ARJ OMANDI [7]. The proof of the NP-completeness of Chromatic Index is based on HOLYER [64]. Figures 11.6, 11.7 and 11.8 are modified diagrams based on those appearing in Holyer's paper. They are used with the permission of the SIAM Journal of Computing.

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Planar Graphs

\subsection*{12.1 Introduction}

A graph \(G\) is planar if it can be drawn in the plane such that no two edges intersect, except at a common endpoint. The vertices of \(G\) are represented as points of the plane, and each edge of \(G\) is drawn as a continuous curve connecting the endpoints of the edge. For example, Figure 12.1 shows a planar graph (the cube), and a planar drawing of the same graph. Although the cube is planar, the drawing on the left is not a planar drawing. These drawings both have straight lines representing the edges, but any continuous curve can be used to represent the edges. We shall often find it more convenient to represent the vertices in planar drawings as circles rather than points, but this is just a drawing convenience.


FI GURE 12.1
Two drawings of the cube

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\subsection*{12.2 J ordan curves}

Any closed, continuous, non-self-intersecting curve \(C\) drawn in the plane divides the plane into three regions: the points inside \(C\), the points outside \(C\), and the points of \(C\) itself. This is illustrated in Figure 12.2. We are relying on an intuitive understanding of the words "plane", "curve", "continuous", "region", etc. Exact mathematical definitions of these ideas would require a lengthy excursion into topology. An intuitive understanding should suffice for this chapter. Notice that the interior of \(C\), denoted INT(C), is bounded, since it is enclosed by \(C\), and that the exterior, denoted \(\operatorname{EXT}(C)\), is unbounded, since the plane is unbounded. If \(u\) is any point in \(\operatorname{INT}(C)\), and \(v \in \operatorname{EXT}(C)\), then any continuous curve with endpoints \(u\) and \(u\) must intersect \(C\) in some point. This fact is known as the Jordan curve theorem. It is fundamental to an understanding of planarity.
DEFINITION 12.1: A closed, continuous, non-self-intersecting curve \(C\) in a surface is called a Jordan curve.


FI GURE 12.2

\section*{The J ordan curve theorem}

Let \(G\) be any graph. We would like to construct a planar embedding of \(G\), if possible. That is, we want to map the vertices of \(G\) into distinct points in the plane, and the edges of \(G\) into continuous curves that intersect only at their endpoints. Let \(\psi\) denote such a mapping. We write \(G \psi\) to indicate the image of \(G\) under the mapping \(\psi\). Let \(C\) be any cycle in \(G\). If \(\psi\) is a planar embedding of \(G\), then \(\psi\) maps \(C\) onto \(C \psi\), a J ordan curve in the plane. For example, consider \(G=K 5\). Let \(V(K 5)=\{U, w, x, y, z\}\). If \(K 5\) were planar, the cycle \(C=(x, y, z)\) must embed as a Jordan curve \(C \psi\) in the plane. Vertex \(u\) is either in
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INT \((C \psi)\) or EXT \((C \psi)\). Without loss of generality, we can place \(u\) in INT \((C \psi)\), as in Figure 12.3. The paths \(u x\), \(u y\), and \(u z\) then divide INT ( \(C \psi\) ) into three smaller regions, each bounded by a Jordan curve. We cannot place \(u\) in EXT \((C \psi)\), as we then cannot embed the path \(u u\) without crossing \(C \psi\). We cannot place \(u\) in any of the smaller regions in \(\operatorname{INT}(C \psi)\), for in each case there is a vertex outside the Jordan curve bounding the region that cannot be reached. We conclude that \(K 5\) cannot be embedded in the plane. \(K 5\) is a non-planar graph. We state this as a lemma.


\section*{FI GURE 12.3}

\section*{Embedding K5}

LEMMA \(12.1 K 5\) and \(K 3,3\) are non-planar graphs.
PROOF The proof for \(K 5\) appears above. The proof for \(K 3,3\) is in Exercise 12.3.1.

\subsection*{12.3 Graph minors, subdivisions}

The graphs \(K 5\) and \(K 3,3\) are special graphs for planarity. If we construct a graph from \(K 5\) by replacing one or more edges with a path of length \(\geq 2\), we obtain a subdivision of \(K 5\). We say that the edges of \(K 5\) have been subdivided.
DEFINITION 12.2: Given a graph \(G\), a subdivision of \(G\) is any graph obtained from \(G\) by replacing one or more edges by paths of length two or more.

It is clear that any subdivision of \(K 5\) or \(K 3,3\) is non-planar, because \(K 5\) and \(K 3,3\) are non-planar. It is apparent that vertices of degree two do not affect the planarity of a graph. The inverse operation to subdividing an edge is to contract an edge with an endpoint of degree two.
DEFINITION 12.3: Graphs G1 and G2 are topologically equivalent or homeomorphic, if G1 can be transformed into \(G 2\) by the operations of subdividing edges and/or contracting edges with an endpoint of degree two.
We will denote by TK5 any graph that is topologically equivalent to \(K 5\). Similarly, TK3, 3 denotes any graph that is topologically equivalent to \(K 3,3\). In general, \(T K\) denotes a graph topologically equivalent to \(K\), for any graph \(K\). If \(G\) is a graph containing a subgraph \(T K 5\) or \(T K 3,3\), then \(G\) must be non-planar. Kuratowski's theorem states that this is a necessary and sufficient condition for a graph to be non-planar. We will come to it later.
If \(G\) is a planar graph, and we delete any vertex \(u\) from \(G\), then \(G-u\) is still planar. Similarly, if we delete any edge \(u u\), then \(G-u v\) is still planar. Also, if we contract any edge \(u u\) of \(G\), then \(G-u v\) is still planar. Contracting an edge can create parallel edges or loops. Since parallel edges and loops do not affect the planarity of a graph, loops can be deleted, and parallel edges can be replaced by a single edge, if desired.
DEFINITION 12.4: Let \(H\) be a graph obtained from \(G\) by any sequence of deleting vertices and/or edges, and/or contracting edges. \(H\) is said to be a minor of \(G\).
Notice that if \(G\) contains a subgraph \(T K 5, K 5\) is a minor of \(G\), even though \(K 5\) need not be a subgraph of \(G\). For we can delete all vertices and edges which do not belong to the subgraph TK5, and then contract edges to obtain \(K 5\). Similarly, if \(G\) has a subgraph \(T K 3,3\), then \(K 3,3\) is a minor of \(G\), but need not be a subgraph. Any graph having \(K 5\) or \(K 3,3\) as a minor is non-planar. A special case of minors is when a graph \(K\) is subdivided to obtain \(G\).
DEFINITION 12.5: Let \(G\) contain a subgraph that is a subdivision of a graph \(K\), where \(\delta(K) \geq 3\). Then \(K\) is said to be a topological minor of \(G\).
LEMMA 12.2 If \(H\) is a minor of \(K\), and \(K\) is a minor of \(G\), then \(H\) is a minor of \(G\).
PROOF This follows from the definition.
A consequence of this lemma is that the relation of being a graph minor is a partial order on the set of all graphs.
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The inverse operation to contracting an edge whose endpoints have degree three or more is splitting a vertex. DEFINITION 12.6: Let \(G\) be any graph with a vertex \(u\) of degree at least three. Let \(u\) be adjacent to vertices \(\{u 1, u 2, \ldots, u k\}\). Construct a graph \(G_{v}^{+}\)by splitting vertex \(u\) : replace \(u\) with two new vertices \(u 1\) and \(u 2\). J oin \(u 1\) to \(\ell 1 \geq 2\) of \(\{u 1, u 2, \ldots, u k\}\), and join \(u 2\) to \(\ell 2 \geq 2\) of them, such that together, \(u 1\) and \(u 2\) are adjacent to all of these vertices. Then join \(u 1\) to \(U 2\).
In any graph \(G_{v}^{+}\)resulting from splitting vertex \(u, u 1\) and \(U 2\) both have degree at least three, and \(G_{v}^{+} \cdot v_{1} v_{2}=G\), that is, splitting vertex \(u\) is an inverse operation to contracting the edge \(u 1 u 2\). Notice that \(G\) is a minor of \(G_{v}^{+}\). Splitting a vertex is illustrated for \(G=K 5\) in Figure 12.4. The following lemma shows that \(K 5\) and \(K 3,3\) are very closely related graphs.


\section*{FI GURE 12.4}

\section*{Splitting a vertex of \(K 5\)}

LEMMA 12.3 Let \(G\) be any graph obtained by splitting a vertex of K5. Then \(G\) contains a subgraph TK3 ,3.
PROOF Let \(U 1\) and \(U 2\) be the two vertices resulting from splitting a vertex of \(K 5\). Each has at least degree
three. Consider \(U 1\). It is joined to \(U 2\). Together, \(U 1\) and \(U 2\) are joined to the remaining four vertices of \(G\), and each is joined to at least two of these vertices. Therefore we can choose a partition of these four vertices into \(x, y\) and \(w, z\) such that \(v_{1} \longrightarrow x, y\) and \(v_{2} \longrightarrow w, z\). Then \(G\) contains a \(K 3,3\) with bipartition \(u 1, w, z\) and \(u 2\), \(x, y\), as illustrated in Figure 12.4.
In the previous example, it was convenient to form a minor \(K 5\) of \(G\) by first deleting a subset of vertices
and/or edges, and then contracting a sequence of edges to obtain \(K 5\). All minors of \(G\) can be obtained in this way, as shown by the following lemma:

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LEMMA 12.4 Suppose that G has a minor H. Then H can be obtained by first deleting a subset of vertices and/or edges of \(G\), and then contracting a sequence of edges.
PROOF Let \(G 0, G 1, \ldots, G k\) be a sequence of graphs obtained from \(G\), where \(G 0=G\) and \(G k=H\), such that each \(G i\), where \(i \geq 1\), is obtained from Gi-1 by deleting a vertex, deleting an edge, or contracting an edge. If all deletions occur before contractions, there is nothing to prove. So let Gi be the first graph obtained from Gi-1 by the deletion of an edge or vertex. Without loss of generality we can assume that \(i \geq 2\), and that \(G 1, \ldots, G i-1\) were obtained by contracting edges \(e 1, \ldots, e i-1\), where \(e j\) is an edge of \(G j-1\). Let \(e i-1=u 1 u 2\).
Suppose first that \(G i=G i-1-U\), for some vertex \(u\). If \(u\) is the result of identifying \(u 1\) and \(u 2\) when ei-1 is contracted, then we can replace \(G i-1, G i\) in the sequence of graphs by \(G_{i-1}^{\prime}, G_{i}^{\prime}\), where \(G_{i-1}^{\prime}\) and \(G_{i}^{\prime}\) are obtained by deleting \(u 1\), and then \(u 2\) from \(G i-2\). If \(u\) is not the result of identifying \(u 1\) and \(u 2\), we can interchange the order of \(G i\) and \(G i-1\) by deleting \(u\) before contracting ei-1. In each case, we obtain an equivalent sequence of graphs with only \(i-1\) edge contractions preceding a deletion. The number of edges contracted does not increase, and the final result is still H .
Suppose now that \(G i=G i-1-u u\), for some edge \(u u\). We know that \(G i-1=G i-2 \cdot e i-1\). We can reverse the order of the two operations, and delete uu before contracting ei-1, thereby replacing \(G i-1\), \(G i\) with graphs
\(G_{i-1}^{\prime}, G_{i}^{\prime}\). Again we obtain an equivalent sequence of graphs with only \(i-1\) edge contractions preceding a deletion. We repeat this as many times as required until all deletions precede all contractions.
It follows that when constructing a minor \(H\) of graph \(G\), we can start with a subgraph of \(G\) and apply a sequence of edge-contractions only to obtain \(H\). Often this is used as the definition of graph minor. Consider the situation when a vertex \(u\) of degree three in \(G\) is split into \(u 1 u 2\). If \(u\) is adjacent to vertices \(x, y\), \(z\) in \(G\), then in \(G_{v}^{+}, u 1\) is adjacent to at least two of \(x, y, z\), and via \(u 2\) there is always a path from \(u 1\) to the third vertex. This is used in the following theorem, and also in Exercise 12.3.5.
THEOREM 12.5 If \(G\) has a minor K3, 3, then \(G\) contains a subgraph TK3,3. If \(G\) has a minor \(K 5\), then \(G\) contains a subgraph TK5 or TK3, 3.
PROOF Suppose that \(G\) has a minor \(K 5\) or \(K 3,3\). If no edges were contracted to obtain this minor, then it is also a subgraph of \(G\). Otherwise let \(G 0, G 1, \ldots, G k\) be a sequence of graphs obtained from \(G\), where \(G 0\) is a subgraph of \(G\), edge ei of \(G i-1\) is contracted to obtain \(G i\), and \(G k\) is either \(K 5\) or \(K 3,3\).
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If each ei has an endpoint of degree two, then we can reverse the contractions by subdividing edges, resulting in a \(T K 5\) or \(T K 3,3\) in \(G\), as required. Otherwise let ei be the edge with largest \(i\), with both endpoints of at least degree three. All edges contracted subsequent to \(G i\) have an endpoint of degree two, so that \(G i\) has a subgraph TK5 or TK3, 3. Gi-1 can be obtained by splitting a vertex \(u\) of \(G i\). If \(u\) is a vertex of \(T K 5\), then by Lemma 12.3, \(G i-1\) contains \(T K 3,3\). If \(u\) is a vertex of \(T K 3,3\), then \(G i-1\) also contains \(T K 3\), 3 . If \(u\) is a vertex of neither \(T K 5\) nor \(T K 3,3\), then \(G i-1\) still contains \(T K 5\) or \(T K 3,3\). In each case we find that \(G 0\) must have a subgraph TK5 or TK3, 3.
DEFINITION 12.7: Given a subgraph \(T K\) of \(G\), equal to \(T K 5\) or \(T K 3,3\). The vertices of \(T K\) which correspond to vertices of \(K 5\) or \(K 3,3\) are called the corners of \(T K\). The other vertices of \(T K\) are called inner verftices of \(T K\). Suppose that \(G\) is a non-planar graph with a subgraph \(T K 5\). Let \(u 1, u 2, \ldots, u 5\) be the corners of \(T K 5\). Each ui has degree four in TK5; the inner vertices of TK5 have degree two. Let Pij be the path in TK5 connecting ui to \(u j\). Consider the situation where \(G\) contains a path \(P\) from \(x \in P_{i j}\) to \(y \in P_{k \ell}\), where \(x\) and \(y\) are inner vertices. This is illustrated in Figure 12.5, where a vertex \(x \in P_{45}\) is connected by a path to \(y \in P_{23}\). We see that in this case, \(G\) contains a TK3,3.


FI GURE 12.5
TK5and TK3, 3
THEOREM 12.6 Let \(G\) contain a subgraph TK5, with corners \(u 1, u 2, . .\), u5connected by paths Pij. If \(G\) has vertices \(x\) and \(y\) such that \(x\) is an inner vertex of \(P i j\), and \(y \in P_{k \ell}\) but \(y \notin P_{i j}\), where \(P i j \neq P k \ell\), then \(G\) contains a TK3,3.
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PROOF One case of the proof is illustrated in Figure 12.5. The remaining cases are done in Exercise 12.3.2. A consequence of Theorem 12.6 is that nearly any graph that has a subgraph \(T K 5\) also has a subgraph \(T K 3\), 3. This theorem will be useful in embedding algorithms for non-planar graphs in Chapter 13. It also permits a recursive characterization of graphs which contain TK5 but not TK3, 3.

\section*{Exercises}
12.3.1 Show that \(K 3,3\) is non-planar, using the J ordan curve theorem.
12.3.2 Complete the proof of Theorem 12.6.
12.3.3 Characterize the class of 2-connected graphs which contain TK5 but not TK3,3.
12.3.4 Construct a \(O(\varepsilon)\) algorithm which accepts as input a graph \(G\) and a subgraph \(T K 5\) with corners u1, u2, ..., \(u 5\), and finds a \(T K 3,3\) containing \(u 1, \cup 2, \ldots, u 5\) if one exists.
12.3.5 Let \(K\) be a graph such that \(\Delta(K) \leq 3\). Show that \(K\) is a minor of \(G\) if and only if \(G\) has a subgraph \(T K\).

\subsection*{12.4 Euler's formula}

Let \(G\) be a connected planar graph with \(n\) vertices and \(\varepsilon\) edges, embedded in the plane by a mapping \(\psi\). If we remove all the points of the image \(G \psi\) from the plane, the plane falls apart into several connected regions. This is equivalent to cutting the plane along the edges of \(G \psi\). For example, if we cut the plane along the edges of the planar embedding of the cube in Figure 12.1, there are six regions, one of which is unbounded. DEFINITION 12.8: The faces of an embedding \(G \psi\) are the connected regions that remain when the plane is cut along the edges of \(G \psi\). The unbounded region is called the outer face.
Notice that if \(u u\) is an edge of \(G\) contained in a cycle \(C\), then ( \(u u\) ) \(\psi\) is a portion of a J ordan curve \(C \psi\). The face on one side of \((u \cup) \psi\) is in INT \((C \psi)\) and the face on the other side is in EXT \((C \psi)\). These faces are therefore distinct. But if uu is an edge not contained in any cycle, then it is a cut-edge, and the same face appears on each side of \((u v) \psi\). For example, in a tree, every edge is a cut-edge, and there is only one face, the outer face.
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We view the plane as an oriented surface, which can be viewed from "above" or "below". Given an embedding \(G \psi\) in the plane, we will view it consistently from one side, which we can assume to be "above" the plane. If we then view an embedding \(G \psi\) from "below" the surface, it will appear to have been reversed. Therefore we choose one orientation ("above") for all embeddings.
The boundary of a face \(F\) of an embedding \(G \psi\) is a closed curve in the plane. It is the image under \(\psi\) of a closed walk \(C\) in \(G\). We can walk along \(C \psi\) so that the interior of \(C \psi\) is to our right-hand side. We will call this a clockwise direction and thereby assign an orientation to \(C\). We shall always choose a clockwise orientation for traversing the boundaries of faces, so that the face \(F\) will be to our right-hand side.
DEFINITION 12.9: An oriented closed walk \(C\) in \(G\) bounding a face of \(G \psi\) is called a facial walk of \(G \psi\) (or facial cycle if \(C\) is a cycle).
Notice that if \(C\) contains a cut-edge \(u u\), then \(u u\) will appear twice on \(C\). The two occurrences of \(u u\) on \(C\) will have opposite orientations. All other edges appear at most once on C. As we will mostly be interested in 2-
connected graphs \(G\), facial walks will almost always be cycles.
DEFINITION 12.10: The degree of a face \(F\) is DEG(F), the length of its facial walk.
Notice that a cut-edge appearing on the facial walk of \(F\) will contribute two to its degree.
THEOREM 12.7 (Euler's formula) Let \(G\) be a connected planar graph with \(n\) vertices and \(\varepsilon\) edges. Let \(G \psi\) have \(f\) faces, where \(\psi\) is a planar embedding of \(G\). Then
\(n+f-\varepsilon=2\)
PROOF The proof is by induction on \(\varepsilon-n\). Every connected graph has a spanning tree. If \(\varepsilon-n=-1\), then \(G\) is a tree. It is clear that every tree has a planar embedding. Since a tree has no cycles, there is only one face, the outer face, so \(f=1\). Euler's formula is then seen to hold for all embeddings of \(G\).
Now suppose that \(\varepsilon-n=k \geq 0\). Choose any cycle \(C\) in \(G\), and any edge \(u v \in C\). Let the faces on the two sides of (uu) \(\psi\) be \(F 1\) and \(F 2\). Consider \(G^{\prime}=G-u u\), with \(n^{\prime}, \varepsilon^{\prime}\) and \(f^{\prime}\) vertices, edges, and faces, respectively. Clearly \(n^{\prime}=n\) and \(\varepsilon^{\prime}=\varepsilon-1\). \(G^{\prime}\) is connected and \(\psi\) is a planar embedding of it. One of the faces of \(G^{\prime} \psi\) is \(F_{1} \cup F_{2}\). The other faces of \(G^{\prime} \psi\) are those of \(G \psi\). Therefore \(f^{\prime}=f-1\) Euler's formula follows by induction.
It follows from Euler's formula that all planar embeddings of a connected graph \(G\) have the same number of faces. Hence we will refer to \(f(G)\) as the number of
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\section*{FI GURE 12.6}

\section*{Two embeddings of a graph}
faces of \(G\), without specifying an embedding. In Figure 12.6 there is an example of a graph with two distinct embeddings. The embeddings have the same number of faces, but the actual faces and their boundaries are different.

\subsection*{12.5 Rotation systems}

Once we have an embedding \(G \psi\), we can choose any vertex \(v \in V(G)\), and walk around \(u \psi\) in a small, clockwise circle. We encounter the incident edges in a certain cyclic order. For example, in the embedding on the left of Figure 12.6, the edges incident on vertex 1 have the clockwise cyclic order (12, 17, 18, 16). Those incident on vertex 2 have the order (23, 27, 21). Those incident vertex 3 have the order (34, 37, 32), etc. In Figure 12.6, the edges are drawn as straight lines. It is conceivable that the embedding \(\psi\) could assign wildly behaved functions, like \(\sin (1 / x)\) to the curves representing the edges of \(G\). Each edge may then be encountered many times when walking around \(u \psi\) in a small circle, no matter how small the circle is chosen. We will assume that this does not occur, and that \(\psi\) assigns well behaved functions (like straight lines or gradual curves) to the edges. In fact, for graphs embedded on the plane, we shall see that it is always possible to draw the edges as straight lines. For a more complete treatment, the reader is referred to the books of GROSS and TUCKER [54] or MOHAR and THOMASSEN [88].
DEFINITION 12.11: Let \(G \psi\) be an embedding in the plane of a loopless connected graph G. A rotation system \(p\) for \(G\) is a mapping from \(V(G)\) to the set of permutations of \(E(G)\), such that for each \(v \in V(G), p(U)\) is the cyclic per-
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mutation of edges incident on \(u\), obtained by walking around \(u \psi\) in a clockwise direction.
Notice that if \(G\) has a loop \(u u\), then as we walk around \(u \psi\), we will cross the loop twice. Therefore \(p(u)\) will contain \(u u\) twice. In order to extend the definition to be correct for graphs with loops, we must ensure that for each loop \(u u\), that \(p(U)\) contains two corresponding "loops" (uU)1 and (uU)2.
Suppose we are given the rotation system \(p\) determined by an embedding \(G \psi\). We can then easily find the facial cycles of \(G \psi\). The following fundamental algorithm shows how to do this. Let \(u \in V(G)\), and let e be
any edge incident on \(u\). We are assuming that given an edge \(e^{\prime}=u u\) in \(p(u)\), we can find the corresponding edge \(e^{\prime \prime}=u u\) in \(p(U)\). A data structure of linked lists can do this easily in constant time. If \(G\) is a simple graph, then the rotation system is completely specified by the cyclic adjacency lists. Given a planar embedding \(G \psi\), we will assume that the adjacency lists are always given in cyclic order, so that the rotation system \(p\) corresponding to \(\psi\) is available.

Algorithm 12.5.1: \(\operatorname{FACIALCYCLE}(G \psi, u, e)\)
comment: \(\left\{\begin{array}{l}\text { Given an embedding } G^{\psi} \text { with corresponding } \\ \text { rotation system } p \text { and vertex } u \text { with incident edge } e,\end{array}\right.\) find the facial cycle containing \(e\).
\(e^{\prime} \leftarrow e\)
repeat
\(\left\{\begin{array}{l}\text { comment: } e^{\prime} \text { currently equals } u v, \text { for some } v \\ v \leftarrow \text { other end of } e^{\prime} \\ e^{\prime \prime} \leftarrow \text { edge of } p(v) \text { corresponding to } e^{\prime} \\ \text { comment: } e^{\prime \prime} \text { currently equals } v u \\ e^{\prime} \leftarrow \text { edge preceding } e^{\prime \prime} \text { in } p(v) \\ u \leftarrow v \\ \text { until } e^{\prime}=e\end{array}\right.\)

LEMMA 12.8 The sequence of edges traversed by \(\operatorname{FACl} \operatorname{ALCYCLE}(G \psi, u, e)\) forms the facial cycle of the face to the right of \(\mathrm{e} \psi\).
PROOF Let \(F\) be the face to the right of \(e \psi\). Let \(e=u u\). As we walk along \(e \psi\) from \(u \psi\) to \(u \psi\), the face \(F\) is to our right-hand side. When we reach \(u \psi\), we are on the image of an edge \(u u\) in \(p(u)\). Since \(p(u)\) has a clockwise cyclic order, the next edge in the facial cycle is the one preceding \(u u\) in \(p(u)\). This is the one chosen by the algorithm. The algorithm repeats this process until it arrives back at the starting edge \(e\).
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Algorithm \(\mathrm{FACl} \operatorname{ALCYCLE}()\) is very simple, but it is tremendously important. It is used in nearly all algorithms dealing with graph embeddings. Notice that its running time is \(O(\varepsilon)\) and that it can be used to find all the facial cycles of \(G \psi\) in \(O(\varepsilon)\) time.
COROLLARY 12.9 The facial cycles of an embedding \(G \psi\) are completely determined by its rotation system.
PROOF All facial cycles can be determined by executing Algorithm \(\operatorname{FAClALCYCLE}(G \psi, u\), e), such that each edge \(e\) is visited at most twice, once for the face on each side of \(e\). The rotation system is the only information about \(\psi\) that is needed.
Thus, it turns out that planar embeddings are essentially combinatorial, as the facial cycles of an embedding are completely determined by the set of cyclic permutations of incident edges of the vertices. Later we will see that for 3 -connected planar graphs, the rotation system is unique, up to orientation. If \(p\) is the rotation system corresponding to an embedding \(\psi\), we will often write \(G p\) instead of \(G \psi\). We call \(G \psi\) a topological embedding, and Gp a combinatorial embedding. The combinatorial embedding determines the facial cycles of the embedding, but it does not give an actual drawing of \(G\) in the plane.

\subsection*{12.6 Dual graphs}

Consider an embedding \(G \psi\), illustrated by the cube in Figure 12.7. Let its faces be listed as \(\{F 1, F 2, \ldots, F f\}\). Two faces Fi and Fj are adjacent if they share a common edge (uu) \(\psi\) on their boundaries. We can construct a planar graph \(G \psi^{*}\) by placing a new vertex fi in each region \(F i\), for \(i=1,2, \ldots, f\). Whenever two faces Fi and Fj share an edge (uu) \(\psi\) on their boundaries, we draw a continuous curve from fi to fj, passing through (uu) \(\psi\) in exactly one interior point, and intersecting \(G \psi\) in only this point. This is illustrated for an embedding of the cube in Figure 12.7. We call \(G \psi^{*}\) a planar dual of \(G \psi\).
LEMMA 12.10 Let \(G \psi\) be a planar embedding with a planar dual \(G \psi^{*}\). Let \(G \psi^{* *}\) be any planar dual of \(G \psi^{*}\). Then \(G^{\psi * *} \cong G^{\psi}\).
PROOF Let the faces of \(G \psi\) be \(F 1, F 2, \ldots, F f\), and let \(f i\) be the vertex of \(G \psi^{*}\) corresponding to Fi. Consider any vertex \(u\) of \(G\) and its cyclic permutation \(p(u)=(u u 1, u u 2, \ldots, u u k)\) of incident edges. Each edge (uul) \(\psi\) separates two
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faces \(F i\) and \(F j\), and so is crossed by a curve connecting fi to \(f j\). As we successively take the edges uul of \(p(u)\), we traverse these curves, thereby constructing a facial boundary of \(G \psi^{*}\). Vertex \(u \psi\) is contained in the
region interior to this facial boundary. We conclude that each face of \(G \psi^{*}\) contains exactly one \(u \psi\), and that the edges ( \(\left.u u^{\ell}\right) \psi\) are curves connecting the vertices \(u \psi\) and \(v_{\ell}^{\psi}\) located inside the faces of \(G \psi^{*}\). That is, the planar dual construction applied to \(G \psi^{*}\) gives back \(G \psi\). Equivalently, \(G^{\psi * *} \cong G^{\psi}\).


\section*{FI GURE 12.7}

\section*{Constructing a dual graph}

Now the adjacencies of the planar dual are determined completely by common edges of the facial cycles of \(\mathfrak{G} \psi\), and these are determined by the rotation sy stem \(p\) of \(\mathcal{G} \psi\). Therefore we define the combinatorial planar dual in terms of the rotation system.
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DEFINITION 12.12: Let \(p\) be the rotation system for \(G\) corresponding to a planar embedding \(\psi\). Let the facial cycles of \(G p\) be \(F=\{F 1, F 2, \ldots, F f\}\). The combinatorial planar dual of \(G p\) is denoted \(G p^{*}\). The vertex set of \(G p^{*}\) is \(\{F 1, F 2, \ldots, F f\}\). The edges of \(G p^{*}\) are defined by a rotation system, also denoted \(p\), and given as follows.
Consider a facial cycle
\[
F i=(u 1, u 2, \ldots, u k),
\]
traversed in a clockwise direction. Each edge uful+1 is contained in exactly two facial cycles, which are adjacent in \(G p^{*}\). As we walk along the facial cycle, the face corresponding to Fi appears on the right-hand side of \((\cup \ell \cup \ell+1) \psi\). On the left-hand side is the face corresponding to \(F \ell^{\prime}\), where \(F \ell^{\prime}\) is the unique facial cycle containing edge \(u \ell+1 u \ell\). We then take
\[
p\left(F_{i}\right)=\left(F_{1^{\prime}}, F_{2^{\prime}}, \ldots, F_{k^{\prime}}\right) .
\]

It is easy to see that FiFj occurs in \(p(F i)\) if and only if FjFi occurs in \(p(F j)\). Thus the definition is valid. If Fi contains a cut-edge \(u u\), then the same face occurs on both sides of \((u u) \psi\). Gp* will then contain a loop FiFi. Since \(u u\) appears twice on the facial cycle, FiFi will occur twice in \(p\) (Fi).
The graph \(\mathrm{Gp}^{*}\) constructed by this definition is always isomorphic to the planar dual \(G \psi^{*}\) constructed above, because of the correspondence between faces and facial cycles. Therefore the rotation system constructed for \(G p^{*}\) always corresponds to a planar embedding \(G \psi^{*}\). It follows from the above lemma that \(G^{p * *} \cong G^{p}\). Now let \(G p\) denote a combinatorial planar embedding of \(G\), and let \(\{F 1, F 2, \ldots, F f\}\) be the facial cycles of \(G p\). The degree of \(F i\) is \(\mathrm{DEG}(F i)\), the length of the walk. Let \(G p^{*}\) be the dual of \(G p\), and write \(n^{*}, \varepsilon^{*}\), and \(f^{*}\) for the numbers of vertices, edges, and faces, respectively, of \(G p^{*}\).
\[
\sum_{i=1}^{f} \operatorname{DEG}\left(F_{i}\right)=2 \varepsilon(G) .
\]

PROOF Each edge \(u u\) of \(G\) is incident on two faces of \(G p\).
LEMMA \(12.12 n^{*}=f, f^{*}=n\), and \(\varepsilon^{*}=\varepsilon\).
PROOF \(n^{*}=f\) follows from the definition of \(G p^{*}\). Since \(G^{p * *} \cong G^{p}\), we have \(f^{*}=n\). Each edge of \(G^{*}\) corresponds to exactly one edge of \(G p\), and every edge of \(G p\) corresponds to exactly one edge of \(G p^{*}\). Therefore \(\varepsilon^{*}=\varepsilon\).
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Algorithm 12.6.1: CONSTRUCTDUAL(Gp)
comment: \(\left\{\begin{array}{l}\text { Given a graph } G \text { with a planar rotation system } p, \\ \text { construct the combinatorial dual } G^{p *} .\end{array}\right.\)
\(n F a c e s \leftarrow 0\)
for all edges uu
do FaceNumber \(\langle u v\rangle \leftarrow 0\)
for all vertices \(u\)
do \(\left\{\begin{array}{l}\text { for each } u v \text { in } p(u) \\ \text { do }\left\{\begin{array}{l}\text { if FaceNumber }\langle u v\rangle=0 \\ \text { then }\left\{\begin{array}{l}n F a c e s ~ \\ \text { traverse the facial cycle } F \text { to the right of } u v \\ \text { using FACIALCYCLE }\left(G^{p}, u, u v\right) \text { and store } \\ n F a c e s \text { in the FaceNumber of each edge of } F\end{array}\right.\end{array}\right.\end{array}\right.\)
comment: \(\left\{\begin{array}{l}\text { we have numbered all the faces } \\ \text { now construct the edges of the dual }\end{array}\right.\)
for all vertices \(u\)
do \(\left\{\begin{array}{l}\text { for each } u v \text { in } p(u) \\ \text { do }\left\{\begin{array}{l}i \leftarrow \text { FaceNumber }\langle u v\rangle \\ \text { if face } i \text { has not been traversed yet } \\ \text { then }\left\{\begin{array}{l}\text { traverse the facial cycle } F \text { to the right of } u v \\ \text { using FACIALCYCLE }\left(G^{p}, u, u v\right), \\ \text { and for each edge } x y \text { of } F \\ \text { let } j=\text { FaceNumber }\langle y x\rangle \text { in } p(y) \\ \text { append } i j \text { to } p(i) \text { in } G^{p *}\end{array}\right.\end{array}\right.\end{array}\right.\)
Algorithm 12.6.1 is a simple algorithm to construct the dual in \(O(\varepsilon)\) time. We assume that the faces of \(G p\) are numbered \(1,2, \ldots, f\); that the rotation system \(p\) is represented as cyclic linked lists; and that the linked list node corresponding to \(u v\) in \(p(u)\) contains a field called the FaceNumber, used to indicate which face of \(G p\) is on the right-hand side of \(u u\) as it is traversed from \(u\) to \(u\). We will denote this by FaceNumber \(\langle u v\rangle\), although it is not stored as an array. We will also use a variable nFaces to count the faces of Gp.
Algorithm 12.6.1 uses FACIALCYCLE() (Algorithm 12.5.1) to walk around the facial cycles of Gp and number them. Notice that \(\operatorname{FACIALCYCLE}()\) only requires the rotation system \(p\) rather than the topological embedding \(\psi\). Each edge of \(G\) is traversed exactly twice, taking a total of \(O(\varepsilon)\) steps. It then traverses the facial cycles again, constructing the rotation system of \(G p^{*}\), using the face numbers which were previously stored. This again takes \(O(\varepsilon)\) steps. Thus, the dual graph is completely determined by the combinatorial embedding Gp.

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\subsection*{12.7 Platonic solids, polyhedra}

The cube is a 3-regular planar graph whose faces all have degree four. So its dual is 4 -regular. It can be seen from Figure 12.7 that the dual of the cube is the octahedron. Let \(G\) be a connected \(k\)-regular planar graph whose dual is \(\ell\)-regular, where \(k, \ell \geq 3\). These graphs are called graphs of the Platonic solids. Then \(k n=2 \varepsilon\) and \(\ell f=2 \varepsilon\). Substituting this into Euler's formula and dividing by \(\varepsilon\) gives
\[
\frac{1}{k}+\frac{1}{\ell}=\frac{1}{2}+\frac{2}{\varepsilon}
\]

If we consider graphs with \(\varepsilon \geq 4\) edges, we have
\[
\frac{1}{2}<\frac{1}{k}+\frac{1}{\ell} \leq 1
\]

As the number of integers satisfying this inequality is limited, this can be used to find all such graphs. They are the graphs of the regular polyhedra-the tetrahedron, cube, octahedron, dodecahedron, and icosahedron. In general, a polyhedron is a geometric solid whose faces are polygons, that is, regions of a plane bounded by a finite sequence of line segments. Each edge of a polyhedron is common to exactly two polygonal faces. So if we consider an edge \(u u 1\) incident on a vertex \(u\), there are two polygons, \(P 1\) and \(P 2\) incident on \(u u 1\). Now \(P 2\) has two edges incident on \(u\). Let the other be \(u \cup 2\). But this edge is also incident on two polygons, \(P 2\) and \(P 3\). Since \(P 3\) has two edges incident on \(u\), we obtain another edge uu3, etc. Continuing in this way, we get a sequence \(u 1, u 2, \ldots\) of vertices adjacent to \(u\), until we return to P1.
DEFINITION 12.13: A polyhedron is a connected collection of polygons such that
1. Each edge is contained in exactly two polygons.
2. Polygons do not intersect, except on a common edge.
3. Any two polygons intersect in at most one edge.
4. The polygons incident on a vertex form a single cycle.

The fourth condition is to prevent a polyhedron created by identifying two vertices of otherwise disjoint polyhedra.
DEFINITION 12.14: The skeleton of a polyhedron is the graph whose vertices are the vertices of the polyhedron, such that vertices \(u\) and \(u\) are adjacent in the graph if and only if \(u v\) is an edge of the polyhedron.
DEFINITION 12.15: A regular polygon, denoted \(\{p\}\), is a planar polygon with \(p\) sides all of equal length. A regular polyhedron, denoted \(\{p, q\}\), is a poly-
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hedron whose faces are polygons \(\{p\}\), such that exactly \(q\) polygons are incident on each vertex. The symbols \(\{p\}\) and \(\{p, q\}\) are called the Schläfli symbols for the polygon and polyhedron, respectively. Notice that given a vertex \(u\) of a regular polyhedron \(\{p, q\}\), the midpoints of the edges incident on \(u\) form a regular polygon \(\{q\}\).
A polyhedron is convex if its interior is a convex region; that is, given any two points \(P\) and \(Q\) in the interior, the line segment connecting \(P\) to \(Q\) is completely contained inside the polyhedron. There is a remarkable theorem of Steinitz characterizing the skeletons of convex polyhedra.
THEOREM 12.13 (Steinitz's theorem) A graph \(G\) is the skeleton of a convex polyhedron if and only if \(G\) is planar and 3-connected.
A proof of Steinitz's theorem can be found in the book by GRÜNBAUM [56] or ZIEGLER [129]. It is too lengthy to include here.

\section*{Exercises}
12.7.1 Find the planar dual of the line graph of \(K 4\). Find the line graph of the cube, and find its planar dual. 12.7.2 Find all \(k\)-regular planar graphs whose duals are \(\ell\)-regular, for all possible values of \(k\) and \(\ell\).
12.7.3 Find the dual of the multigraph constructed from \(K 4\) by doubling each edge. Refer to Figure 12.8.


FI GURE 12.8
Find the dual graph
12.7.4 A planar graph \(G\) is self-dual if it is isomorphic to its planar dual. Find a self-dual planar graph on \(n\) vertices, for all \(n \geq 4\).
12.7.5 Program the algorithm CONSTRUCTDUAL(), and test it on the graphs ofExer cises 12.7.1 and 12.7.2. 12.7.6 Show that a planar graph is bipartite if and only if its dual is Eulerian.

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12.7.7 Find the Schläfli symbols of the tetrahedron, cube, octahedron, dodecahedron, and icosahedron.
12.7.8 Given a Schläfli symbol \(\{p, q\}\) for a polyhedron, find a formula for the number of vertices, edges, and faces of the polyhedron in terms of \(p\) and \(q\).

\subsection*{12.8 Triangulations}

A planar embedding \(G p\) whose faces all have degree three is called a triangulation. If \(G p\) is a triangulation, then \(3 f=2 \varepsilon\). Substituting into Euler's formula gives:
LEMMA 12.14 A triangulation Gpsatisfies \(\varepsilon=3 n-6\) and \(f=2 n-4\).
If \(G p\) is not a triangulation, and has no multiple edges or loops, then every face has at least degree three. We can convert \(G p\) to a triangulation by adding some diagonal edges to faces of degree four or more. This gives the following:
LEMMA 12.15 A simple planar graph \(G\) has \(\varepsilon \leq 3 n-6\).
For example, since \(K 5\) has \(\varepsilon=10>3 n-6\), we can conclude that \(K 5\) is non-planar.
One consequence of the above lemma is that \(O(\varepsilon)\) algorithms on planar graphs are also \(O(n)\) algorithms. For example, a DFS or BFS in a planar graph takes \(O(n)\) steps. This will be useful in algorithms for testing planarity, or for drawing or coloring a planar graph, or constructing dual graphs.
Given a graph \(G\), we can subdivide any edges of \(G\) without affecting the planarity of \(G\). Therefore, we will assume that \(G\) has no vertices of degree two. We will also take \(G\) to be 2 -edge-connected, so that there are no vertices of degree one. Let \(G\) be a simple planar graph, and let ni be the number of vertices of degree \(i\), for \(i=3,4, \ldots\) Counting the edges of \(G\) gives
\[
3 n 3+4 n 4+5 n 5+\ldots=2 \varepsilon<6 n-12
\]

Counting the vertices of \(G\) gives
\[
n 3+n 4+n 5+\ldots=n
\]

Multiply the second equation by 6 and subtract the two equations to obtain:
LEMMA 12.16 A simple planar graph with no vertices of degree one or two satisfies \(3 n 3+2 n 4+n 5 \geq 12+n 7+2 n 8+3 n 9+\ldots\)
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COROLLARY 12.17 A simple planar graph with no vertices of degree one or two has a vertex of degree three, four, or five.
PROOF The values ni are non-negative integers.
The above corollary results in a technique for reducing a planar triangulation on \(n\) vertices to one on \(n-1\) vertices that is fundamental for understanding the structure of planar graphs, and for handling them algorithmically.

Algorithm 12.8.1: REDUCEGRAPH(Gp)
comment: \(\left\{\begin{array}{l}\text { Given a simple planar triangulation } G \text { on } n>4 \text { vertices } \\ \text { with rotation system } p . \\ \text { Construct a planar triangulation } G^{\prime} \text { on } n-1 \text { vertices. }\end{array}\right.\)
if there is a vertex \(u\) with \(\operatorname{DEG}(u)=3\)
then \(\left\{\begin{array}{l}\text { let } p(u)=(u x, u y, u z) \\ G^{\prime} \leftarrow G-u \\ \text { return }\left(G^{\prime}\right)\end{array}\right.\)
if there is a vertex \(u\) with \(\operatorname{DEG}(u)=4\)
(let \(p(u)=(u w, u x, u y, u z)\)
if \(w \not y\)
then \(\left\{\right.\) then \(\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+w y \\ w y \text { replaces } w u \text { in } p(w) \text { and } y w \text { replaces } y u \text { in } p(y)\end{array}\right.\)
else \(\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+x z \\ x z \text { replaces } x z\end{array}\right.\)
return \(\left(G^{\prime}\right)\)
comment: otherwise, DEG(u)=5
let \(p(u)=(u u, u w, u x, u y, u z)\)
if \(v \nrightarrow x\) and \(v \nrightarrow y\)
then \(\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+v x+v y \\ v x, v y \text { replace } v u \text { in } p(v) \\ x v \text { replaces } x u \text { in } p(x) \text { and } y v \text { replaces } y u \text { in } p(y)\end{array}\right.\)
else if \(v \longrightarrow x\)
then \(\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+w y+w z \\ w y, w z \text { replace } y u \text { in } p(w) \\ y w \text { replaces } y u \text { in } p(y) \text { and } z w \text { replaces } z u \text { in } p(z)\end{array}\right.\)
else if \(v \longrightarrow y\)
then \(\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+z w+z x \\ z w, z x \text { replace } z u \text { in } p(z) \\ w z \text { replaces } w u \text { in } p(w) \text { and } z w \text { replaces } z u \text { in } p(z)\end{array}\right.\)
return ( \(G^{\prime}\) )

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Here \(G n\) is a triangulation on \(n \geq 4\) vertices. If \(n=4\), then \(K 4\) is the only possibility. So we assume that \(n>4\). We know that Gn always has a vertex of degree three, four, or five.
THEOREM 12.18 Given a simple planar triangulation Gnon n>4 vertices, Algorithm REDUCEGRAPH() constructs a simple planar triangulation \(G n-1\) on \(n-1\) vertices.
PROOF We know that \(G n\) has a vertex of degree three, four, or five. If there is a vertex \(u\) of degree three, let \(p(u)=(u x, u y, u z)\), as illustrated in Figure 12.9. Since \(G n\) is a triangulation, we know that \(x y, y z\), and \(z x\) are edges of \(G n\). Consequently \(G n-u\) is also a trianqulation.

\(G_{n}\)

\(G_{n-1}\)

\section*{FI GURE 12.9}

DEG \((u)=3\)
If there is a vertex \(u\) of degree four, let \(p(u)=(u w, u x, u y, u z)\), as illustrated in Figure 12.10. Since \(G n\) is a triangulation, we know that \(w x, x y, y z\), and \(z w\) are edges of \(G n\). When \(u\) is deleted, one face of \(G n-u\) is a quadrilateral. If \(w \nmid y\), we can add the edge wy to get a planar triangulation \(G n-u+w y\). If \(w \longrightarrow y\), then edge \(w y\) is exterior to the quadrilateral face. Consequently \(x \nrightarrow z\), so that we can add the edge \(x z\) to get a planar triangulation \(G n-u+x z\). In either case we get a planar triangulation \(G n-1\).
If there is a vertex \(u\) of degree five, let \(p(u)=(u u, u w, u x, u y, u z)\), as illustrated in Figure 12.11. Since \(G n\) is a triangulation, we know that \(u w, w x, x y, y z\), and \(z u\) are edges of \(G\). When \(u\) is deleted, one face of \(G n-u\) is a pentagon. If \(v \nrightarrow x\) and \(v \nrightarrow y\), we can add the edges \(u x\) and \(u y\) to get a planar triangulation \(G n-u+u x+u y\). Otherwise, if \(v \longrightarrow x\), then edge \(u x\) is exterior to the pentagonal face. Consequently, \(w \nrightarrow y\) and \(w \nsim\), so that we can add the edges \(w y\) and \(w z\) to get a planar triangulation \(G n-u+w y+w z\).
Otherwise, if \(v \longrightarrow y\), then edge \(u y\) is exterior to the pentagonal face, and we can proceed as above to get a triangulation \(G n-u+z w+z x\). The proof is complete.
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FI GURE 12.10
DEG(u)=4


\section*{FI GURE 12.11}

DEG \((u)=5\)
Note that Algorithm REDUCEGRAPH() requires the degrees of the vertices. These can be computed in \(O\) (n) time and stored in an array. Once the degrees are known, the reduction from \(G n\) to \(G n-1\) takes constant time. Usually, this algorithm will be applied recursively to reduce a planar triangulation \(G n\) to \(G 4\), which must equal \(K 4\). If the algorithm is being used to find a planar embedding of \(G\) n, or to color it, the graph will then be rebuilt in reverse order.

\subsection*{12.9 The sphere}

The plane can be mapped onto the surface of the sphere by a simple transformation called stereographic projection. Place a sphere on the surface of the plane, so that it is tangent at the south pole. See Figure 12.12. Now from any point \(P\) on the plane, construct a straight-line \(L\) to the north pole of the sphere. \(L\) will intersect the surface of the sphere in some point. Call it \(P^{\prime}\). This transformation maps any
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point \(P\) in the plane to a point \(P^{\prime}\) on the surface of the sphere. The mapping is clearly invertible and continuous. The only point on the sphere to which no point of the plane is mapped is the north pole.


\section*{FI GURE 12.12}

\section*{Mapping the plane to the sphere}

If \(\mathcal{G} \psi\) is an embedding of a graph on the plane, then stereographic projection will map the points of \(G \psi\) onto an embedding of \(G\) on the surface of the sphere. Conversely, if we are given an embedding of \(G\) on the
surface of the sphere, we can roll the sphere to ensure that the north pole is not a point of the embedding. Then use stereographic projection to map the surface of the sphere onto the plane, thereby obtaining an embedding of \(G\) on the plane. Consequently, embedding graphs on the plane is equivalent to embedding them on the sphere.
When a graph is embedded on the surface of the sphere, the faces are the regions that remain when the sphere is cut along the edges of \(G\). There is no outer face. Every face is bounded. However, the face that contains the north pole will become the outer face when the embedding is projected onto the plane. By rolling the sphere to place any desired face at the top, we can make any face the outer face. We state this as a lemma.
LEMMA 12.19 A planar graph can be drawn so that any facial cycle, any edge, or any vertex appears on the boundary of the outer face.
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\subsection*{12.10 Whitney's theorem}

The plane and sphere are oriented surfaces. Consider a graph \(G\) embedded on the plane as \(G p\), where \(p\) gives the clockwise orientation of the edges incident on each vertex. We are assuming that the plane is viewed from above. If we now view the plane from below, the clockwise orientation of each \(p(u)\) will appear counter clockwise. If an embedding \(G p\) is projected onto the sphere, then each \(p(u)\) will appear clockwise if viewed from inside the sphere, but counter clockwise if viewed from outside the sphere. Given a rotation system \(p\), we write \(\bar{p}\) for the rotation system obtained by reversing the cyclic order of each \(p(u)\). So \(\bar{p}(u)=\) \(p(u)^{-1}\). The embeddings \(G p\) and \(G^{\bar{p}}\) are usually considered equivalent.
DEFINITION 12.16: Let \(G^{p_{1}}\) and \(G^{p_{2}}\) be two embeddings of a graph \(G\), with rotation systems \(p 1\) and \(p 2\), respectively. \(G^{p_{1}}\) and \(G^{p_{2}}\) are isomorphic embeddings if there is an automorphism of \(G\) which transforms \(p 1\) into \(p 2 . G^{p_{1}}\) and \(G^{p_{2}}\) are equivalent embeddings if there is an automorphism of \(G\) which transforms \(p 1\) into either \(p 2\) or \(\bar{p}_{2}\).
An automorphism of \(G\) will permute the vertices of \(G\), and consequently alter the edges in the cyclic permutations of a rotation system. If \(\theta\) is an automorphism of \(G\), then \(p 1(u)=(e 1, e 2, . ., e k)\) is transformed by \(\theta\) into \(\theta(p 1(u))=(\theta(e 1), \theta(e 2), \ldots, \theta(e k))\). If this equals \(p 2(\theta(u))\), for all vertices \(u\), then \(G^{p_{1}}\) and \(G^{p_{2}}\) are isomorphic. Isomorphic rotation systems are equivalent.
Two isomorphic rotation systems for \(K 4\) are illustrated in Figure 12.13. Here \(p_{2}=\overline{p_{1}}\). it is easy to see that if we take \(\theta=(3,4)\), then \(\theta(p 1(u))=p 2(\theta(u))\), for all \(u=1,2,3,4\).
A triangulation \(G\) on 7 points is shown in Figure 12.14. Two rotation systems for it, \(p 1\) and \(p 2\) are also given below. Here we also have \(p_{2}=\overline{p_{1}}\). However, there is no automorphism of \(G\) that will transform \(p 1\) into \(p 2\). This can be verified using the degrees of the vertices. Vertex 2 is the only vertex of degree six. Hence any automorphism \(\theta\) must fix 2 . So \(\theta(p 1(2))\) must equal \(p 2(2)\). The only vertices of degree four are vertices 1 and 5. Therefore either \(\theta(1)=1\) or \(\theta(1)=5\). If \(\theta(1)=1\), then from \(p 2(2)\) we see that \(\theta(3)=6\). This is impossible as vertices 3 and 6 have different degrees. If \(\theta(1)=5\), then from \(p 2(2)\) we have \(\theta(3)=4\), which is also impossible, as vertices 3 and 4 have different degrees.
So \(G^{p_{1}}\) and \(G^{p_{2}}\) are equivalent embeddings that are non-isomorphic. This can only occur if there is an automorphism of \(G\) mapping \(p 1\) to \(\bar{p}_{2}\), but no automorphism mapping \(p 1\) to \(p 2\); that is, \(G^{p_{2}}\) is obtained by "flipping" \(G^{p_{1}}\) upside down. The example of Figure 12.13 shows that this is not possible with \(K 4\), but is possible with the triangulation of Figure 12.14.
DEFINITION 12.17: A planar graph \(G\) is orientable if there exists a planar rotation system \(p\) such that \(G^{p} \not \not G^{\bar{p}}\). Otherwise \(G\) is non-orientable.
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\[
\begin{aligned}
& p_{1}(1)=(12,13,14) \\
& p_{1}(2)=(21,24,23) \\
& p_{1}(3)=(31,32,34) \\
& p_{1}(4)=(41,43,42)
\end{aligned}
\]

\[
\begin{aligned}
& p_{2}(1)=(12,14,13) \\
& p_{2}(2)=(21,23,24) \\
& p_{2}(3)=(31,34,32) \\
& p_{2}(4)=(41,42,43)
\end{aligned}
\]

\section*{FI GURE 12.13}

\section*{Two isomorphic rotation systems for K4}

So \(K 4\) is non-orientable, but the graph of Figure 12.14 is orientable. An example of a 2-connected graph with two inequivalent planar embeddings is shown in Figure 12.6. Whitney's theorem states that if \(G\) is 3 connected, this cannot happen. Let \(C\) be a cycle in a connected graph \(G\). \(C\) is a separating cycle if \(G-V(C)\) is disconnected.
THEOREM 12.20 (Whitney's theorem) Let \(G\) be a 3-connected planar graph. Let \(p\) be any planar rotation system for \(G\). The facial cycles of \(G p\) are the induced, non-separating cycles of \(G\).
PROOF Let \(C\) be an induced, non-separating cycle of \(G\). In any planar embedding of \(G\), \(C\) corresponds to a \(J\) ordan curve in the plane. There can be vertices of \(G\) in the interior or exterior of the J ordan curve, but not both, because \(C\) is non-separating. Without loss of generality, assume that any vertices of \(G-C\) are in the exterior of the J ordan curve. It follows that the interior of the J ordan curve is a face, so that \(C\) is a facial cycle of \(G\).
Conversely, let \(C\) be a facial cycle. Without loss of generality, we can assume that \(C\) corresponds to a J ordan curve whose interior is a face. If \(G\) contains an edge \(u u\) which is a chord of \(C\), then \(u\) and \(u\) divide \(C\) into two paths \(C[u, u]\) and \(C[u, u]\). Since uu must be embedded exterior to \(C\), there can be no path from an is disconnected, a contradiction, as \(G\) is 3 -connected. Consequently, \(C\) is an interior vertex of \(C[u, u]\) to an interior vertex of \(C[u, u]\). Therefore \(G-\{u, u\}\) induced cycle. Let \(x\) be any vertex of \(G-C\). If \(x\) is the only vertex of \(G-C\), then \(G-C=x\), so that \(C\) is a non-separating cycle, and we are done. Otherwise page_298

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\[
\begin{aligned}
& p_{1}(1)=(12,16,14,13) \\
& p_{1}(2)=(21,23,24,25,27,26) \\
& p_{1}(3)=(31,34,32) \\
& p_{1}(4)=(41,46,45,42,43) \\
& p_{1}(5)=(52,54,56,57) \\
& p_{1}(6)=(61,62,67,65,64) \\
& p_{1}(7)=(72,75,76)
\end{aligned}
\]

\[
\begin{aligned}
& p_{2}(1)=(12,13,14,16) \\
& p_{2}(2)=(21,26,27,25,24,23) \\
& p_{2}(3)=(31,32,34) \\
& p_{2}(4)=(41,43,42,45,46) \\
& p_{2}(5)=(52,57,56,54) \\
& p_{2}(6)=(61,64,65,67,62) \\
& p_{2}(7)=(72,76,75)
\end{aligned}
\]

\section*{FI GURE 12.14}

\section*{Two equivalent, non-isomorphic rotation systems}
let \(y\) be another vertex of \(G-C\). Since \(G\) is 3 -connected, \(G\) contains at least three internally disjoint xy-paths. At most two of these paths can intersect \(C\). See Figure 12.15. Therefore \(G-C\) contains an \(x y\)-path, for all \(x, y\). It follows that \(C\) is a non-separating cycle of \(G\).
One consequence of Whitney's theorem is that, if \(G\) is a 3 -connected planar graph, an embedding can be found purely from the cycles of \(G\). If we can identify an induced, non-separating cycle \(C\), we can then assign an orientation to \(C\). Each edge of \(C\) will be contained in another induced, non-separating cycle, so that the orientation of adjacent cycles will thereby also be determined. Continuing in this way, a complete rotation system for \(G\) can be constructed. This rotation system can then be used to construct a dual graph.
DEFINITION 12.18: Let \(G\) be a 3-connected planar graph. The abstract dual is \(G^{*}\), a dual graph determined by the induced non-separating cycles of \(G\).
The abstract dual is the same as \(G p^{*}\), where \(p\) is a rotation system determined by the induced, nonseparating cycles, but it can be constructed without reference to a rotation system.

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\section*{FI GURE 12.15}

\section*{Whitney's theorem}

Up to equivalence of embeddings, a 3-connected planar graph has just one rotation system, so that the abstract dual is uniquely defined. Orientable 3-connected planar graphs have just two rotation systems (which are inverses of each other). Non-orientable 3-connected planar graphs have just one rotation system. Whitney's theorem does not provide an algorithm for determining whether a graph is planar, as the characterization of planarity in terms of induced, non-separating cycles does not lead to an efficient algorithm. There are too many cycles in a graph to effectively find them all and determine whether they are induced, non-separating cycles.

\subsection*{12.11 Medial digraphs}

Let \(G\) be a loopless graph with a planar rotation system \(p\). Note that \(G\) is allowed to be a multigraph. We construct a digraph representing Gp.
DEFINITION 12.19: The medial digraph \(M(G)\) is obtained from \(G\) by subdividing every edge \(u u\) of \(G\) with a vertex xuu. The edges of \(M(G)\) consist of all arcs of the form ( \(u, x u u\) ), (xuu, \(u\) ), and (xuu, xuw), where \(u u\) and uw are consecutive edges in \(p(u)\), with uw following \(u u\).
An example of a medial digraph is shown in Figure 12.16 . Since \(G\) is planar, \(M(G)\) will also be planar. Notice that \(G^{\bar{p}}\) is the digraph converse of \(G p\), obtained by reversing all directed edges. If \(G\) is non-orientable, then \(G^{\bar{p}} \cong G^{p}\). Otherwise \(G^{\bar{p}} \not \approx G^{p}\). This provides a means of determining whether an embedding is orientable. It also gives information about the automorphisms of the planar embedding.


FI GURE 12.16
A medial digraph Exercises
12.11.1 Find all triangulations on 4, 5, 6, and 7 vertices.
12.11.2 Prove that \(G^{p_{1}}\) and \(G^{p_{2}}\) are isomorphic embeddings if and only if \(M\left(G^{p_{1}}\right)\) and \(M\left(G^{p_{2}}\right)\) are isomorphic digraphs.
12.11.3 Determine whether the platonic solids are orientable.
12.11.4 Prove that a planar embedding \(G p\) is orientable if and only if \(G p *\) is orientable.
12.11.5 Determine which of the triangulations on 5,6 , and 7 vertices are orientable.
12.11. 6 Determine the graph \(G\) for which \(M(G)\) is shown in Figure 12.16.

\subsection*{12.12 The 4-color problem}

Given a geographic map drawn in the plane, how many colors are needed such that the map can be colored so that any two regions sharing a common border have different colors? In 1852, it was conjectured by Francis Guthrie that four colors suffice. This simple problem turned out to be very difficult to solve. Several flawed "proofs" were presented. Much of the development of graph theory originated in attempts to solve this conjecture. See AIGNER [2] for a development of graph theory based on the 4-color problem. In 1976, Appel and Haken announced a proof of the conjecture. Their proof was based on the results of a computer program that had to be guaranteed bug-free. A second computer proof

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by ALLAI RE [3] appeared in 1977. Each of these approaches relied on showing that any planar graph contains one of a number of configurations, and that for each configuration, a proper coloring of a smaller (reduced) graph can be extended to a proper coloring of the initial graph. The computer programs generated all irreducible configurations, and colored them. In the Appel-Haken proof, there were approximately 1800 irreducible configurations. The uncertainty was whether all irreducible configurations had indeed been correctly generated. In 1995, ROBERTSON, SANDERS, SEYMOUR, and THOMAS [104] presented another proof, also based on a computer program, but considerably simpler than the original, requiring only 633 irreducible configurations.
In this section, we present the main ideas of Kempe's 1879 "proof" of the 4-color theorem.
Given a geographic map drawn in the plane, one can construct a dual graph, by placing a vertex in the interior of each region, and joining vertices by edges if they correspond to adjacent regions. Coloring the regions of the map is then equivalent to coloring the vertices of the dual, so that adjacent vertices are of different colors. Consequently, we shall be concerned with coloring the vertices of a planar graph.
THEOREM 12.21 (4-Color theorem) Every planar graph can be properly colored with four colors.
If \(G\) is any simple planar graph, then it is always possible to extend \(G\) to a simple triangulation, by adding diagonal edges in non-triangular faces. Therefore, if we can prove that all simple planar triangulations are 4colorable, the result will be true for all planar graphs. Hence we assume that we are given a planar triangulation \(G n\) on \(n\) vertices. We attempt to prove the 4-color theorem (Theorem 12.21) by induction on \(n\). The colors can be chosen as the numbers \(\{1,2,3,4\}\). Given a coloring of \(G\), then the subgraph induced by any two colors \(i\) and \(j\) is bipartite. We denote it by Kij.
DEFINITION 12.20: Given any 4-coloring of a planar graph G, each connected component of Kij is called a Kempe component. The component containing a vertex \(x\) is denoted \(\operatorname{Kij}(x)\). A path in Kij between vertices \(u\)
and \(u\) is called a Kempe chain.
Notice that if we interchange the colors \(i\) and \(j\) in any Kempe component, we obtain another coloring of \(G\). Now let \(G n\) be a simple triangulation on \(n\) vertices. If \(n=4\), then \(G n=K 4\). It is clear that Theorem 12.21 is true in this case. Assume that \(n>4\). By Corollary 12.17, we know that \(G n\) has a vertex of degree three, four, or five. Let \(u\) be such a vertex. Using Algorithm 12.8.1, we reduce \(G n\) to a simple planar triangulation \(G n-1\) by deleting \(u\) and adding up to two diagonals in the resulting

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face. We assume as an induction hypothesis, that \(G n-1\) has a 4 -coloring. There are three cases.
Case 1. \(\operatorname{DEG}(u)=3\).
Let the three adjacent vertices to \(u\) be \((x, y, z)\). They all have different colors. Therefore there is a fourth color available for \(u\), giving a coloring of \(G\) n.
Case 2. \(\operatorname{DEG}(u)=4\).
Let the four vertices adjacent to \(u\) in \(G n\) be ( \(w, x, y, z\) ), with a diagonal wy in \(G n-1\). It is clear that \(w, x\), and \(y\) have different colors. If \(x\) and \(z\) have the same color, then a fourth color is available for \(u\). Otherwise, let \(w\), \(x, y, z\) be colored 1, 2, 3, 4, respectively. There may be a Kempe chain from \(x\) to \(z\). If there is no Kempe chain, interchange colors in the Kempe component \(K 24(x)\), so that \(x\) and \(z\) now both have color 4. If there is a Kempe chain from \(x\) to \(z\), there can be no Kempe chain from \(w\) to \(y\), for it would have to intersect the \(x z\) Kempe chain. Interchange colors in \(K 13(w)\), so that \(w\) and \(z\) now both have color 3 . In each case there is a fourth color available for \(u\).


FI GURE 12.17
Kempe chains
Case 3. \(\operatorname{DEG}(u)=5\).
Let the five vertices adjacent to \(u\) in \(G n\) be ( \(u, w, x, y, z\) ), with diagonals \(u x\) and \(u y\) in \(G n-1\). It is clear that \(u, x\), and \(y\) have different colors. Since we have a 4-coloring of \(G n-1\), the pentagon ( \(u, w, x, y, z\) ) is colored in either 3 or 4 colors. If it is colored in three colors, there is a fourth color available for \(u\). If it is colored in four colors, then without loss of generality, we can take these colors to be ( \(1,2,3,4,2\) ), respectively. If \(K 13(u)\) contains no \(u x\)-Kempe chain, then we can interchange colors in \(K 13(U)\), so that \(u\) and \(x\) are now both colored 3 . Color 1 is then available for \(u\). If \(K 14(U)\) contains
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no uy-Kempe chain, then we can interchange colors in \(K\) 14(U), so that \(u\) and \(y\) are now both colored 4 . Color 1 is again available for \(u\). Otherwise there is a Kempe chain Pux connecting \(u\) to \(x\) and a Kempe chain Puy connecting \(u\) to \(y\). It follows that \(K 24(w)\) contains no wy-Kempe chain, as it would have to intersect Pux in K13(u). Similarly, K23(z) contains no uz-Kempe chain, as it would have to intersect Puy in K14(u). If Pux and Puy intersect only in vertex \(u\), then we can interchange colors in both K24(w) and K23(z), thereby giving w color 4 and \(z\) color 3 . This makes color 2 available for \(u\). The difficulty is that Pux and Puy can intersect in several vertices. Interchanging colors in \(K 24(w)\) can affect the other Kempe chains, as shown in Figure 12.18, where the pentagon ( \(u, w, x, y, z\) ) is drawn as the outer face.


FI GURE 12.18

\section*{Intersecting Kempe chains}

Although this attempted proof of Theorem 12.21 fails at this point, we can use these same ideas to prove the following.
THEOREM 12.22 (5-Color theorem) Any planar graph can be colored in five colors.
PROOF See Exercise 12.12.1.
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Appel and Haken's proof of the 4-color theorem is based on the important concept of reducibility. Given a graph \(G\), a reducible configuration \(H\) is a subgraph of \(G\) with the property that \(H\) can be reduced to a smaller subgraph \(H^{\prime}\), such that a 4 -coloring of \(H^{\prime}\) can be extended to all of \(H\) and \(G\). If every planar graph contained a reducible configuration, then every planar graph could be 4-colored. Appel and Haken's proof was essentially a computer program to construct all irreducible configurations, and to show that they could be 4colored. The difficulty with this approach is being certain that the computer program is correctly constructing all irreducible configurations. The reader is referred to SAATY and KAI NEN [106] or WOODALL and WILSON [128] for more information on reducibility.

\section*{Exercises}
12.12.1 Prove Theorem 12.22, the 5 -color theorem.
12.12.2 Let \(G\) be a planar triangulation with a separating 3 -cycle ( \(u, u, w\) ). Let \(H\) and \(K\) be the two connected subgraphs of \(G\) that intersect in exactly \((u, u, w)\), such that \(G=H \cup K\). Show how to construct a 4-coloring of \(G\) from 4 -colorings of \(H\) and \(K\).
12.12.3 Let \(G\) be a planar triangulation with a separating 4 -cycle ( \(u, u, w, x\) ). Let \(H\) and \(K\) be the two connected subgraphs of \(G\) that intersect in exactly \((u, u, w, x)\), such that \(G=H \cup K\). Show how to construct a 4-coloring of \(G\) from 4 -colorings of the triangulations \(H+u w\) and \(K+u w\). Hint: \(u, u\), and \(w\) can be assumed to have the same colors in \(H\) and \(K\). If \(x\) is colored differently in \(H\) and \(K\), look for an \(x u\)-Kempe chain, try interchanging colors in \(K i j(x)\), or try coloring \(H+U x\) and \(K+U x\).
12.12.4 All lakes are blue. Usually all bodies of water are colored blue on a map. Construct a planar graph with two non-adjacent vertices that must be blue, such that the graph cannot be colored in four colors subject to this requirement.

\subsection*{12.13 Straight-line drawings}

Every simple planar graph can be drawn in the plane with no edges crossing, so that each edge is a straight line. Read's Algorithm is a linear-time algorithm for doing this. It is based on the triangulation reduction. Suppose that \(G n\) is a triangulation on \(n\) vertices that has been reduced to a triangulation \(G n-1\) on \(n-1\) vertices, by deleting a vertex \(u\) as in Algorithm 12.8.1, and adding up to two edges \(e\) and \(e^{\prime}\). Suppose that a straight-line embedding of \(G n-1\) has already been computed. If \(\mathrm{DEG}(u)=3\), let \(x, y, z\) be the adjacent vertices. We can place \(u\) inside the triangle ( \(x, y, z\) ) to obtain a straight-line embedding of

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Gn. If \(\operatorname{DEG}(u)=4\), the edge \(e\) is a diagonal of a quadrilateral in \(G n-1\). We can place \(u\) on the line representing \(e\) to obtain a straight-line embedding of \(G n\).
Suppose now that \(\operatorname{DEG}(u)=5\). The edges \(e=u x\) and \(e^{\prime}=u y\) are diagonals of a pentagon in \(G n-1\). This pentagon may have several possible polygonal shapes, which are illustrated in Figure 12.19. The triangle ( \(u\), \(x, y\) ) is completely contained inside the pentagon. Inside ( \(u, x, y\) ), there is a "visible" region, shown shaded gray. The visible region can be calculated, by extending the lines of the adjacent triangles with sides ux and \(u y\), and intersecting the half-planes with the triangle ( \(u, x, y\) ). Vertex \(u\) can then be placed inside the visible region to obtain a straight-line embedding of \(G n\). Thus in each case, a straight-line embedding of Gn-1 can be extended to \(G n\).
This gives:
THEOREM 12.23 (Fáry's theorem) Every planar graph has a straight-line embedding.


FI GURE 12.19
The "visible" region
Read's algorithm begins by triangulating \(G\) if \(\varepsilon<3 n-6\). It then deletes a
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sequence of vertices \(u 1, u 2, \ldots, u n-4\) to reduce \(G\) to \(K 4\). It next assigns a planar coordinatization to the vertices of \(K 4\), and then restores the deleted vertices in reverse order. For each vertex ui deleted, it is necessary to store ui and its degree, so that it can later be correctly restored to the graph. Finally, the triangulating edges are removed. The result is a straight-line embedding of \(G\).

Algorithm 12.13.1: READSALGORITHM(GP)
comment: \(\left\{\begin{array}{l}\text { Given a simple planar graph } G \text { on } n \geq 4 \text { vertices } \\ \text { with rotation system } p, \text { construct a straight-line } \\ \text { drawing of } G \text { in the plane. }\end{array}\right.\)
triangulate \(G\) without creating multiple edges or loops
mark all triangulating edges as "virtual" edges
\(i \leftarrow 1\)
while \(n>4\)
do \(\left\{\begin{array}{l}G \leftarrow \text { REDUCEGRAPH }(G) \\ u_{i} \leftarrow \text { the vertex that was deleted } \\ i \leftarrow i+1 \\ n \leftarrow n-1\end{array}\right.\)
comment: \(G\) is now \(K 4\)
assign pre-chosen coordinates to the vertices of \(K 4\)
for \(i=n-4\) downto 1
do \(\left\{\begin{array}{l}\text { calculate the visible region for } u_{i} \\ \text { restore } u_{i} \text { to } G\end{array}\right.\)
remove all virtual edges from \(G\)
It is easy to see that Read's algorithm is \(O(n)\). It takes \(O(n)\) steps to compute the degrees of \(G\), and to triangulate \(G\). It takes \(O(n)\) steps to reduce \(G\) to \(K 4\), and then \(O(n)\) steps to rebuild \(G\). Read's algorithm can be modified by initially choosing any facial cycle \(F\) of \(G\), and assigning coordinates to the vertices of \(F\) so that they form a regular convex polygon. The reduction to \(K 4\) is then modified so that vertices of the outer facial cycle \(F\) are never deleted. The result is a planar embedding with the given facial cycle as the outer face. The embeddings produced by Read's algorithm are usually not convex embeddings. Tutte has shown how to produce a straight-line embedding of a graph such that all faces are convex regions, by solving linear equations. Consider any face of \(\mathcal{G}\), with facial cycle ( \(u 1, u 2, \ldots, u k\) ). We begin by assigning coordinates to \(u 1\), \(u 2, \ldots, u k\) such that they form a convex polygon in the plane. This will be the outer face of a planar embedding of \(G\).
Tutte then looks for a coordinatization of the remaining vertices with the special property: the coordinates of \(u i\), where \(i>k\), are the average of the coordinates of

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all adjacent vertices. A coordinatization with this property is called a barycentric coordinatization. We can express it in terms of matrices as follows.
Let \(A\) be the adjacency matrix of \(G\) such that the first \(k\) rows and columns correspond to vertices \(u 1, u 2, \ldots\), \(u k\). Let \(D\) be the \(n \times n\) diagonal matrix such that entry Dii equals 1 , if \(i \leq k\). If \(i>k\), entry Dii equals DEG(ui). Let \(X\) be the vector of \(x\)-coordinates of the vertices, and let \(Y\) be the vector of \(y\)-coordinates. Construct a matrix \(B\) from \(A\) by replacing the first \(k\) rows with zeroes. Then the first \(k\) entries of \(B X\) are zero. But if \(i>k\), the ith entry is the sum of the \(x\)-coordinates of vertices adjacent to ui. Let \(X k\) denote the vector whose first \(n\) entries are the \(x\)-coordinates of \(u 1, \ldots, u k\), and whose remaining entries are zero. \(Y k\) is similarly defined. Then the barycentric condition can be written as
\[
D X=X k+B X, D Y=Y k+B Y
\]

These equations can be written as \((D-B) X=X k\) and \((D-B) Y=Y k\). Consider the matrix \(D-B\).
LEMMA 12.24 The matrix \(D-B\) is invertible.
PROOF Consider the determinant \(\operatorname{det}(D-B)\). The first \(k\) rows of \(D-B\) look like an identity matrix. Expanding the determinant along the first \(k\) rows gives \(\operatorname{det}(D-B)=\operatorname{det}(K)\) where \(K\) is the matrix formed by the last \(n-k\) rows and columns. K looks very much like a Kirchhoff matrix, except that the degrees are not quite right. In fact, if we construct a graph \(G^{\prime}\) from \(G\) by identifying \(u 1, U 2, \ldots, u k\) into a single vertex \(u 0\), and deleting the loops created, then \(K\) is formed from the Kirchhoff matrix \(K\left(G^{\prime}\right)\) by deleting the row and column corresponding to \(u 0\). It follows that \((K)= \pm \tau\left(G^{\prime}\right)\), by the matrix-tree theorem. Since \(G^{\prime}\) is a connected graph, \(\operatorname{det}(K) \neq 0\), so that \(D-B\) is invertible.
It follows that the barycentric equations have a unique solution, for any assignment of the coordinates \(X k\) and \(Y k\). Tutte has shown that if \(G\) is a 3-connected planar graph, this solution has remarkable properties: if we begin with a convex polygon for the outer face, the solution is a planar coordinatization with straight lines, such that no edges cross. All faces, except the outer face, are convex regions. No three vertices of any facial cycle are collinear.
It is fairly easy to see that a planar barycentric coordinatization of a 3-connected graph must have convex faces. For consider a non-convex face, as in Figure 12.20. Vertex \(u\) is a corner at which a polygonal face is non-convex. Clearly vertex \(u\) is not on the outer face. Let the two adjacent vertices of the polygon be \(u\) and \(w\). Since \(G\) is 3 -connected, \(u\) has at least another adjacent vertex. All other vertices adjacent to \(u\) must be in the angle between the lines \(u u\) and \(u w\) as indicated in the diagram, since \(G\) is 3 -connected, and there are no crossing edges. But then all

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vertices adjacent to \(u\) are to the right of the dotted line, which is impossible in a barycentric coordinatization.


\section*{FI GURE 12.20}

\section*{A non-convex face}

\subsection*{12.14 Kuratowski's theorem}

In this section we will prove Kuratowski's theorem. The proof presented is based on a proof by KLOTZ [73]. It uses induction on \(\varepsilon(G)\).
If \(G\) is a disconnected graph, then \(G\) is planar if and only if each connected component of \(G\) is planar.
Therefore we assume that \(G\) is connected. If \(G\) is a separable graph that is planar, let \(H\) be a block of \(G\)
containing a cut-vertex \(u\). \(H\) is also planar, since \(G\) is. We can delete \(H-u\) from \(G\), and find a planar
embedding of the result. We then choose a planar embedding of \(H\) with \(u\) on the outer face, and embed \(H\)
into a face of \(G\) having \(u\) on its boundary. This gives:
LEMMA 12.25 A separable graph is planar if and only if all its blocks are planar.
So there is no loss in generality in starting with a 2-connected graph \(G\).
THEOREM 12.26 (Kuratowski's theorem) A graph \(G\) is planar if and only if it contains no subgraph TK3, 3 or TK5.

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PROOF It is clear that if \(G\) is planar, then it contains no subgraph \(T K 3,3\) or \(T K 5\). To prove the converse, we show that if \(G\) is non-planar, then it must contain \(T K 3,3\) or \(T K 5\). We assume that \(G\) is a simple, 2-connected graph with \(\varepsilon\) edges. To start the induction, notice that if \(\varepsilon \leq 6\), the result is true, as all graphs with \(\varepsilon \leq 6\) are planar. Suppose that the theorem is true for all graphs with at most \(\varepsilon-1\) edges. Let \(G\) be non-planar, and let \(a b \in E(G)\) be any edge of \(G\). Let \(G^{\prime}=G-a b\). If \(G^{\prime}\) is non-planar, then by the induction hypothesis, it contains a \(T K 3,3\) or \(T K 5\), which is also a subgraph of \(G\). Therefore we assume that \(G^{\prime}\) is planar. Let \(\kappa(a, b)\) denote the number of internally disjoint \(a b\)-paths in \(G^{\prime}\). Since \(G\) is 2 -connected, we know that \(\kappa(a, b) \geq 1\).
Case 1. \(\kappa(a, b)=1\).
\(G^{\prime}\) has a cut-vertex \(u\) contained in every \(a b\)-path. Add the edges \(a u\) and \(b u\) to \(G^{\prime}\), if they are not already present, to get a graph \(H\), with cut-vertex \(u\). Let \(H a\) and \(H b\) be the blocks of \(H\) containing \(a\) and \(b\),
respectively. If one of Ha or Hb is non-planar, say Ha , then by the induction hypothesis, it contains a \(T K 3,3\) or \(T K 5\). This subgraph must use the edge \(a u\), as \(G^{\prime}\) is planar. Replace the edge au by a path consisting of the edge ab plus a bu-path in \(H b\). The result is a \(T K 3,3\) or \(T K 5\) in \(G\). If \(H a\) and \(H b\) are both planar, choose planar embeddings of them with edges \(a u\) and \(b u\) on the outer face. Glue them together at vertex \(u\), remove the edges \(a u\) and bu that were added, and restore \(a b\) to obtain a planar embedding of \(G\), a contradiction.
Case 2. \(\kappa(a, b)=2\).
Let \(P 1\), and \(P 2\) be two internally disjoint ab-paths in \(G^{\prime}\). Since \(\kappa(a, b)=2\), there is a vertex \(u \in P_{1}\) and \(v \in P_{2}\) such that all ab-paths contain at least one of \(\{u, u\}\), and \(G^{\prime}-\{u, u\}\) is disconnected. If \(K a\) and \(K b\) denote the connected components of \(G^{\prime}-\{u, u\}\) containing \(a\) and \(b\), respectively, let \(G_{a}^{\prime}\) be the subgraph of \(G\) ' induced by \(K_{a} \cup\{u, v\}\), let \(G_{b}^{\prime}\) be the subgraph induced by \(K_{b} \cup\{u, v\}\). Now add a vertex \(x\) to \(G_{a}^{\prime}\), adjacent to \(u, u\), and a to obtain a graph Ha. Similarly, add \(y\) to \(G_{b}^{\prime}\) adjacent to \(u, u\), and \(b\) to obtain a graph \(H b\). Suppose first that \(H a\) and \(H b\) are both planar. As vertex \(x\) has degree three in Ha, there are three faces incident on \(x\). Embed Ha in the plane so that the face with edges \(u x\) and \(x u\) on the boundary is the outer face. Embed Hb so that edges uy and \(y \mathrm{u}\) are on the boundary of the outer face. Now glue Ha and Hb together at vertices \(u\) and \(u\), delete vertices \(x\) and \(y\), and add the edge \(a b\) to obtain a planar embedding of \(G\). Since \(G\) is non-planar, we conclude that at least one of \(H a\) and \(H b\) must be non-planar. Suppose that Ha is non-planar. It must contain a subgraph TK5 or TK3, 3. If the TK5 or TK3, 3 does not contain \(x\), then it is also contained in \(G\), and we are done. Otherwise the TK5 or TK3, 3 contains \(x\). Now Hb is 2-connected (since \(G\) is), so that it contains internally disjoint

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paths Pbu and Pbu connecting \(b\) to \(u\) and \(u\), respectively. These paths, plus the edge \(a b\), can be used to replace the edges \(u x, u x\), and \(a x\) in Ha to obtain a \(T K 5\) or TK3, 3 in \(G\).
Case 3. \(\kappa(a, b) \geq 3\).
Let \(P 1, P 2\), and \(P 3\) be three internally disjoint ab-paths in \(G^{\prime}\). Consider a planar embedding of \(G^{\prime}\). Each pair of paths \(P_{1} \cup P_{2}, P_{1} \cup P_{3}\), and \(P_{2} \cup P_{3}\) creates a cycle, which embeds as a J ordan curve in the plane. Without loss of generality, assume that the path \(P 2\) is contained in the interior of the cycle \(P_{1} \cup P_{3}\), as in Figure 12.21. The edge ab could be placed either in the interior of \(P_{1} \cup P_{2}\) or \(P_{2} \cup P_{3}\), or else in the exterior of \(P_{1} \cup P_{3}\). As \(G\) is non-planar, each of these regions must contain a path from an interior vertex of \(P i\) to an interior vertex of \(P j\). Let \(P 12\) be a path from \(u 1\) on \(P 1\) to \(U 2\) on \(P 2\). Let \(P 13\) be a path from \(u 1\) on \(P 1\) to \(u 3\) on \(P 3\). Let \(P 23\) be a path from \(U 2\) on \(P 2\) to \(u 3\) on \(P 3\). If \(u 1 \neq u 1\), contract the edges of \(P 1\) between them. Do the same for \(u 2, u 2\) on \(P 2\) and \(u 3, u 3\) on \(P 3\). Adding the edge \(a b\) to the resultant graph then results in a \(T K 5\) minor. By Theorem 12.5, G contains either a TK5 or TK3, 3.


\section*{FI GURE 12.21}

A K5minor
We can also state Kuratowski's theorem in terms of minors.
THEOREM 12.27 (Wagner's theorem) A graph \(G\) is planar if and only if it does not have \(K 3,3\) or \(K 5\) as a minor.
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PROOF It is clear that if \(G\) is planar, then it does not have \(K 3,3\) or \(K 5\) as a minor. Conversely, if \(G\) does not have \(K 3,3\) or \(K 5\) as a minor, then it cannot have a subgraph TK3, 3 or TK5. By Kuratowski's theorem, \(G\) is planar.
The graphs K5 and K3, 3 are called Kuratowski graphs.

\section*{Exercises}
12.14.1 Find a TK3, 3 or \(T K 5\) in the Petersen graph.
12.14.2 Find a TK3, 3 or \(T K 5\) in the graph of Figure 12.22.
12.14.3 Show that if \(G\) is a non-planar 3 -connected graph, then either \(G=K 5\), or else \(G\) contains a \(T K 3,3\). 12.14.4 Let \(G\) be a graph with a separating set \(\{u, u\}\). Let \(H^{\prime}\) be a connected component of \(G-\{u\), \(u\}\), and let \(H\) denote the graph induced by \(V\left(H^{\prime}\right) \cup\{u, v\}\). Let \(K\) be the graph induced by \(V(G)-V\left(H^{\prime}\right)\), so that \(G=H \cup K\), and \(H\) and \(K\) intersect only in \(\{u, u\}\). Let \(H+=H+u u\) and \(K+=K+u u\). Show that \(G\) is planar if and only if \(H+\) and \(K+\) are planar.
12.14.5 Let \(G\) be a 3-connected graph with at least five vertices. Show that \(G\) contains an edge \(x y\) such that \(G \cdot x y\) is 3 -connected. Hint: If \(G \cdot x y\) is not 3 -connected, choose \(x y\) so that the subgraph \(K\) of the preceding exercise is as large as possible, and find a contradiction. This was proved by THOMASSEN [114].
12.14.6 Suppose that a coordinatization in the plane of an arbitrary 3-connected graph \(G\) on \(n\) vertices is obtained by solving the barycentric equations ( \(D-B\) ) \(X=X k\) and \((D-B) Y=Y k\). Describe an \(O(n 3)\) algorithm which determines whether \(G\) is planar, and constructs a rotation system, using the coordinates \(X\) and \(Y\).

\subsection*{12.15 The Hopcroft-Tarjan algorithm}

A number of different algorithms for planarity-testing have been developed. The first linear-time algorithm was the Hopcroft-Tarjan planarity algorithm. Given a 2 -connected graph \(G\) on \(n\) vertices, it determines whether \(G\) is planar in \(O(n)\) time. If \(G\) is found to be planar, it can be extended to construct a rotation
system, too. If \(G\) is found to be non-planar, it can also be modified to construct a \(T K 5\) or \(T K 3,3\) subgraph, although this is somewhat more difficult. We present a simplified version of the Hopcroft-Tarjan algorithm here.
There is no loss in generality in starting with a 2-connected graph \(G\). Suppose first that \(G\) is hamiltonian, and that we are given a hamilton cycle \(C\), and number the vertices of \(C\) consecutively as \(1,2, \ldots, n\). The remaining edges of \(G\) are
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FI GURE 12.22
Find a TK3, 3
chords of \(C\). For each vertex \(u\), order the adjacent vertices ( \(u 1, u 2, \ldots, u k\) ) so that \(u 1<u 2<\ldots<u k\). We then start at vertex \(u=n\), and follow the cycle \(C\) back to vertex 1 . As we proceed, we will place each chord \(u u\) either inside \(C\) or outside \(C\). When we have returned to vertex 1 , we will have constructed a planar embedding of \(G\). We draw the cycle \(C\) as in Figure 12.23 , with the path from 1 to n starting near the top, and moving down the page.
Consider the example of Figure 12.23. The algorithm stores two linked lists of chords, one for the inside of \(C\), and one for the outside of \(C\). We denote these as \(L i\) and \(L o\), respectively. Each linked list defines a sequence of chords [ \(u 1 u 1, u 2 \cup 2, u 3 \cup 3, \ldots]\) as they are added to the embedding. The inside of \(C\) appears in the diagram to the left of the path from 1 to \(n\). The outside of \(C\) appears to the right of the path. In the example, the algorithm begins at vertex \(n=7\) with adjacent vertices (1,3,4, 6). It first places the "chord" (7, 1), which really completes the cycle, on the inside linked list. It then places chords \((7,3)\) and \((7,4)\) also on Li . The inside linked list is now \([(7,1),(7,3),(7,4)]\). The chord \((7,4)\) is called the leading chord in the linked list. The next chord to be inserted is to be placed after it. To determine whether the next chord ( \(u, u\) ) fits on the inside, it need only compare its endpoint \(u\) with the upper endpoint of the current leading chord of Li. After placing the chords incident on vertex 7 , the algorithm moves to vertex 6 , where it sees the chord \((6,1)\). This will not fit on the inside (because \(1<4\) ), but is easily placed on the outside linked list. It then moves to vertex 5,


\section*{FI GURE 12.23}

The Hopcroft-Tarjan algorithm
and places \((5,2)\) also on Lo. The outside linked list is then \([(6,1),(5,2)]\), where \((5,2)\) is the leading chord. It then moves to vertex 4 , where it sees the chord \((4,1)\). When the algorithm moves up the cycle to vertex 4 , the leading chord of \(L i\) is moved past \((7,4)\) to \((7,3)\), because the chord \((4,1)\) is above \((7,4)\). It then determines that \((4,1)\) will not fit on the inside (because \(1<3\), where \((7,3)\) is the leading chord of \(L i\) ); and that \((4,1)\) will not fit on the outside (because \(1<2\), where \((5,2)\) is the leading chord of Lo). Therefore \(G\) is nonplanar. In fact the cycle \(C\), together with the three chords \((7,3),(5,2)\), and \((4,1)\) form a subgraph \(T K 3,3\), which we know to be non-planar.

Algorithm 12.15.1: BASICHT(G, C)
comment:
Given a 2 -connected graph \(G\) with a hamilton cycle
\(C=(1,2, \ldots, n)\). Determine whether \(G\) is planar.
for \(u \leftarrow n\) downto 1
(suppose that \(u\) is adjacent to \(\left(v_{1}, v_{2}, \ldots, v_{k}\right)\) for \(j \leftarrow 1\) to \(k\)

We must still describe what SWITCHSIDES \((u, u)\) does. Its purpose is to determine whether some chords of Li and Lo can be interchanged to allow ( \(u, u\) ) to be embedded. Its description will follow.
This simplified version of the Hopcroft-Tarjan algorithm contains the main features of the complete algorithm, but is much easier to understand. The algorithm stores two linked lists, Li and Lo. Each list has a leading chord-the chord after which the next chord is to be placed when inserted in the list as the sequence of chords is extended. Initially the leading chord will be the last chord in the linked list, but this will change as the algorithm progresses. It is convenient to initialize both linked lists with dummy chords, so that each list has at least one chord. Each chord stored in a linked list will be represented by a pair, denoted (LowerPt, UpperPt), where UpperPt is the endpoint with the smaller value-it is above the LowerPt in the diagram. As the chords are placed in the linked lists, a linear order is thereby defined on the chords in each list. Given a chord \((u, u)\), where \(u>u\), the next chord after \((u, u)\) is the chord nested immediately inside \((u, u)\), if there is such a chord. If there is no such chord, then the next chord after \((u, u)\) is the first chord below \((u, u)\) if there is one; that is, the first chord whose UpperPt \(\geq u\). To determine whether a chord \((u, u)\) fits either inside or outside \(C\), we need only compare \(u\) with the leading chord's UpperPt. It fits if \(u \geq\) UpperPt. As \(u\) moves up the cycle, we must adjust the pointer to the leading chord on both sides
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to ensure that the leading chord always satisfies UpperPt<u.
So we know how to store the chords, and how to determine whether a chord ( \(u, u\) ) fits in either side. If it fits, we insert it in the appropriate linked list and continue with the next chord. What if \((u, u)\) will not fit in either side?

\subsection*{12.15.1 Bundles}

Two chords (u1,u1) and ( \(u 2, u 2\) ) are said to be in conflict if either \(u 1>u 2>u 1>u 2\) or \(u 2>u 1>u 2>u 1\). Conflicting chords cannot both be placed on the same side of \(C\). We define a conflict graph \(K\) whose vertices are the set of all chords currently in the linked lists. Two chords are adjacent in \(K\) if they are in conflict. A set of chords corresponding to a connected component of the conflict graph is called a bundle (originally called a "block" in HT's paper; however, the term "block" has another graph theoretical meaning). The conflict graph must alway s be bipartite, as two chords in the same linked list must never be in conflict. Therefore each bundle \(B\) is also bipartite-the bipartition of a bundle consists of the chords inside the cycle, denoted \(B i\), and the chords outside the cycle, denoted Bo. A typical bundle is shown as the shaded area in Figure 12.24. The two shaded zones represent Bi and Bo for one bundle. Any chord inside a shaded zone conflicts with all chords in the opposite shaded zone. Notice that if the conflict graph is not bipartite, that it is impossible to assign the chords to the inside and outside of C. Consequently, \(G\) must be non-planar in such a case.
Now the conflict graph \(K\) is changing dynamically as the algorithm progresses. Therefore the bundles are also changing. However, they change in a simple way. Each chord is initially a bundle by itself, until it is found to conflict with another chord. In a 1-chord bundle, one of Bi and Bo will be empty. The algorithm will store the bundles in a stack.
Consider the situation where no conflicts have yet been discovered, so that all chords so far have been placed on Li . The algorithm is visiting vertex \(u\) on \(C\), attempting to place a chord ( \(u, u\) ). The leading inside chord satisfies LowerPt \(\geq u>\) UpperPt. It belongs to a 1-chord bundle, the current bundle. The algorithm first attempts
to nest \((u, u)\) inside the leading inside chord. If it fits, then \((u, u)\) becomes the new leading inside chord, and a new 1-chord bundle is created containing only \(u u\), which becomes the current bundle-the bundles are nested. The innermost bundle is always the current bundle, and is stored at the top of the bundle stack. If \((u, u)\) will not fit on the inside, it conflicts with the leading inside chord. ( \(u, u\) ) may conflict with several chords of the inside linked list. They will be consecutive chords of Li, preceding the leading chord. These chords initially belong to different bundles, but will become merged into the current bundle \(B\), thereby forming \(B i\), when ( \(u, u\) ) is placed in the outside list. Bo will consist of \((u, u)\). If \(B\) denotes the current bundle, we will call Bi and Bo the current inside
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FI GURE 12.24

\section*{Bundles of chords}
and outside bundles, although they are part of the same bundle.
At this point we can say that the current inside and outside bundles consist of one or more consecutive
chords in the two linked lists. This will be true at each point of the algorithm.
So in order to represent a bundle B, we require two pointers into each of Li and Lo, being the first and last chords of \(L i\) that belong to \(B i\) and the first and last chords of Lo that belong to Bo. We could switch the chords of Bi to Lo and the chords of Bo to Li by reassigning four pointers. Because the bundles are nested, we store them as a stack.
When a chord \((u, u)\) is placed in one of the linked lists, and it does not conflict with the leading chord on either side, a new current bundle \(B\) containing \((u, u)\) is created, nested inside the previous current bundle. The current bundle is the one at the top of the stack. As vertex \(u\) moves up the cycle, we eventually have \(u \leq\) UpperPt for the uppermost chord in the current bundle. The current bundle \(B\) is then no longer relevant, as no more chords can conflict with it. Therefore \(B\)
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is removed from the stack and deleted. The next bundle \(B^{\prime}\) on the stack becomes the current bundle. When this happens, the chords belonging to \(B i\) or \(B o\) often occur in \(L i\) or Lo, respectively, as a consecutive subsequence of chords contained within \(B_{i}^{\prime}\) or \(B_{o}^{\prime}\). When \(B\) is deleted, the effect is to merge the chords of \(B\) into \(B^{\prime}\). Then \(B^{\prime}\) becomes the new current bundle. Its chords are again consecutive chords of Li and Lo .

\subsection*{12.15.2 Switching bundles}

We have a chord \((u, u)\) that will not fit on the inside or outside of \(C\). What do we do? There are several possible ways in which this can arise. Two of them are illustrated in Figure 12.25. The current bundle is shown shaded in gray.

In the left diagram of Figure 12.25, ( \(u, u\) ) conflicts with one or more chords in \(B i\) and in Bo. If we form a subgraph consisting of the cycle \(C\), the edge ( \(u, u\) ), and a conflicting chord from each of \(B i\) and \(B o\), we have a subgraph TK3, 3 which we know is non-planar. In the right diagram of Figure \(12.25,(u, u)\) conflicts with one or more chords in Bo. It does not conflict with any chords in Bi but it does conflict with the leading chord ( \(a\), b) of \(L i\), which is in a previous inside bundle. If we interchange the the chords of Bi and Bo , there will be room to place \((u, u)\) in Lo. This can be done in constant time, because the chords in each bundle are consecutive chords in the linked list. We only need to change a constant number of pointers to transfer a sequence of chords from Li to Lo, and vice versa.
A third situation also exists which is nearly the mirror image of the right diagram of Figure 12.25, in which ( \(u\), u) conflicts with one or more chords in Bi, but does not conflict with any chords in Bo, and does conflict with the leading chord \((a, b)\) of \(L o\), which is in a previous outside bundle. It can be handled in a similar way. Suppose that a situation similar to the right diagram of Figure 12.25 exists. \(B\) is the current bundle, and chord ( \(u, u\) ) conflicts with a chord of Bo, but not with Bi. The leading chord of Lo is in Bo. The leading chord of Li is ( \(a, b\) ), which is not in \(B i\). Since every chord is contained in some bundle, \((a, b)\) is in a previous bundle on the stack. The bundles are nested, so that \(B\) is nested within a bundle \(B^{\prime}\), which may in turn be nested within in a bundle \(B^{\prime \prime}\), etc. Without loss of generality, suppose that there are at least three bundles on the stack, which begins \(B, B^{\prime}, B^{\prime \prime}\), and that \((a, b)\) is in \(B^{\prime \prime}\).
Now \(B\) is nested within \(B^{\prime}\), which is nested within \(B^{\prime \prime}\). Since the leading chord of \(L i\) is in \(B^{\prime \prime}\), it follows that \((u, u)\) does not conflict with any chord of \(B_{i}^{\prime}\), and that \((u, u)\) does conflict with some chord of \(B_{o}^{\prime}\). So \((u, u)\) conflicts with a chord of both Bo and \(B_{o}^{\prime}\). If \((u, u)\) can be embedded, Bo and \(B_{o}^{\prime}\) must be merged into one bundle, and they must both be on the same side of \(C\). They will be in the same part of the bipartition of \(K\). Therefore the algorithm merges \(B\) and \(B^{\prime}\). Call the result \(B\). It then discovers that interchanging the chords of \(B i\) and Bo allows

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\section*{FI GURE 12.25}

\section*{Switching bundles}
\((u, u)\) to be placed on \(L o\). Since \((u, u)\) conflicts with \((a, b)\) in \(L i\), the bundles \(B\) and \(B^{\prime \prime}\) are then also merged. The properties of bundles are summarized as follows:
1. The bundles correspond to connected components of the conflict graph.
2. The bundles are nested inside each other, and consequently stored on a stack.
3. The current bundle is the one on the top of the stack.
4. The chords of each Bi and Bo form a contiguous subsequence of \(L i\) and \(L o\), respectively.

The description of \(\operatorname{SWITCHSIDES}(u, u)\) Algorithm 12.15.2 can now be given. It is called when chord \((u, u)\) conflicts with both the leading chord of \(L i\) and the leading chord of Lo. The algorithm needs to know whether the leading chord of \(L i\) is within \(B i\), and whether the leading chord of \(L o\) is within Bo. This can be done by comparing the endpoints of the leading chords with the first and last chords of Bi and Bo. One of the leading chords is always in the current bundle. The other must be in a previous bundle, or the conflict graph will be non-bipartite, and \(G\) will be non-planar.
When Bi and Bo are interchanged, the leading chords in Li and Lo also change. Suppose that the leading
chord of \(L o\) is within Bo, but that the leading chord of

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Li is in a previous bundle, as in Figure 12.25. The new leading chord of Lo can easily be found by taking the first chord of Lo following Bo. The new leading chord of \(L i\) is the former leading chord of Lo.
This procedure merges the bundles on the stack until either a non-bipartite conflict graph is found, in which case it returns zero, or until it becomes possible to place \((u, u)\) in one of \(L i\) or Lo. Notice that swapping the chords of Bi and Bo takes a constant number of steps, and that merging the current bundle with the previous bundle on the stack also takes a constant number of steps. The total number of bundles is at most the number of edges of \(G\), so that the total number of steps required by \(\operatorname{SWITCHSIDES}(u, u)\) is \(O(n)\), summed over all iterations.
If SWITCHSIDES \((u, u)\) returns either 1 or -1 , then it is possible to place \((u, u)\) inside or outside \(C\). If SWITCHSIDES \((u, u)\) returns 0 , then the current bundle contains chords of \(L i\) and \(L o\) that conflict with \((u, u)\), so that the conflict graph is not bipartite. In this case \(G\) is non-planar. However, it is not particularly easy to find a \(T K 5\) or \(T K 3,3\) in \(G\) in this situation. If the algorithm succeeds in placing all the chords, then a rotation system can be found from the order in which the chords occur in \(L i\) and Lo.
So the three main components of this simplified version of the Hopcroft-Tarjan algorithm are:
1. Construct the linked lists Li and Lo in the right order.
2. Construct the connected components of the conflict graph as a stack of bundles.
3. Keep the bundles up to date. Each time that a chord is added to one of Li or Lo , the current bundle must be updated.
These steps can all be performed in linear time.
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Algorithm 12.15.2: SWITCHSIDES \((u, u)\)
comment: \((u, u)\) conflicts with the leading chord of \(L i\) and \(L o\) while (true)
let \(B\) denote the current bundle
let \(B^{\prime}\) denote the previous bundle on the stack if the leading chord of \(L_{o}\) is within \(B_{o}\)
(if the leading chord of \(L_{i}\) is within \(B_{i}\)
then return (0) "non-planar"
comment: \(\left\{\begin{array}{l}\text { otherwise the leading chord of } L_{i} \\ \text { is in a bundle previous to } B\end{array}\right.\)
if the leading chord of \(L_{i}\) is within \(B_{i}^{\prime}\)
do \(\left\{\begin{array}{l}\text { then }\left\{\begin{array}{l}\text { interchange the chords of } B_{i} \text { and } B_{o} \\ \text { merge } B \text { and } B^{\prime} \\ \text { find the new leading chords of } L_{i} \text { and } L_{o} \\ \text { if }(u, v) \text { does not conflict with } L_{o} \\ \text { then return }(-1)\end{array}\right. \\ \text { else merge } B \text { and } B^{\prime} \quad \text { " }(u, v) \text { now fits in } L_{o} \text { " }\end{array}\right.\)
else \(\left\{\begin{array}{l}\text { comment: } \begin{array}{l}\left\{\begin{array}{l}\text { the leading chord of } L_{i} \text { is within } B_{i} \\ \text { the leading chord of } L_{o} \text { is } \\ \text { in a bundle previous to } B\end{array}\right. \\ \text { if the leading chord of } L_{o} \text { is within } B_{o}^{\prime}\end{array} \\ \text { then }\left\{\begin{array}{l}\text { interchange the chords of } B_{i} \text { and } B_{o} \\ \text { merge } B \text { and } B^{\prime} \\ \text { find the new leading chords of } L_{i} \text { and } L_{o} \\ \text { if }(u, v) \text { does not conflict with } L_{i} \\ \text { then return }(1) \text { " }(u, v) \text { now fits in } L_{i} "\end{array}\right. \\ \text { else merge } B \text { and } B^{\prime}\end{array}\right.\)

\subsection*{12.15.3 The general Hopcroft-Tarjan algorithm}

Up to now, we have assumed that we are given a hamilton cycle \(C\) in a 2 -connected graph \(G\) which we are testing for planarity. If we are not given a starting hamilton cycle, the algorithm is the recursive extension of the hamilton cycle case. We give a brief sketch only. The first step is to perform a depth-first search in \(G\) starting from vertex 1, to assign a DF-numbering to the vertices, and to calculate the low-points of all vertices. The DFS will construct a DF-spanning tree \(T\) rooted at 1. Number the vertices of \(G\) according to their DF-numbers. It
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FI GURE 12.26
Subtrees Tuvand Tuw
can happen that the tree \(T\) is a hamilton path, in which case we have the proceeding situation exactly-the vertices are numbered consecutively along \(T\), and as the DFS returns from recursion, it visits the vertices in the order \(n, n-1, \ldots, 1\). If we sort the chords incident on each vertex in order of increasing DF-number of the other endpoint, the algorithm is identical.
If the tree \(T\) is not a hamilton path, consider the situation where the DFS is visiting vertex \(u\), and a recursive call
While visiting vertex \(u\), the recursive call DFS( \(u\) ) constructs a subtree Tuu, where \(u\) is adjacent to \(u\). Refer to Figure 12.26. When Tuv is constructed, LowPt \([u]\) is calculated. Since \(G\) is 2 -connected, this is a vertex somewhere above \(u\) in \(T\). The entire subtree Tuu behaves very much like a single chord ( \(u\), LowPt[u]). Therefore the vertices adjacent to \(u\) must be sorted according to LowP[[•] values, just as in the simplified algorithm (where \(u\) is used rather than LowPt[ \(u]\) ).
There are two kinds of possible subtree, and these are illustrated as Tuu and Tuw in Figure 12.26. Both Tuu and Tuw have the same LowPt, equal to \(x\). Notice that Tuu has a frond with endpoint \(y\) between \(u\) and \(x\), but that Tuw has no such

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frond. We will call a subtree like Tuw a type I subtree, and a subtree like Tuu a type II subtree. It is easy to distinguish type I and II subtrees. The DFS can compute the second low-point as well as the low-point. If the second low-point is between \(u\) and LowPt[U], then the subtree Tuv is of type II; otherwise it is of type I. Now a type I subtree Tuw behaves exactly like a chord ( \(u\), LowPt[w]). There can be any number of them, and they can be nested inside each other in any order. However, they cannot be nested inside a type II subtree. Therefore we must embed all type I subtrees at \(u\) with LowPt=x before any type II subtrees at \(u\) with LowPt \(=x\). This can be accomplished by ordering the adjacent vertices at \(u\) so that fronds \(u x\) and type I subtrees Tuw with LowPt[w] = \(x\) precede type II subtrees Tuv with LowPt[u]=x. Hopcroft and Tarjan assign a weight to all edges incident on \(u\). A frond \(u x\) has weight \(2 x\). A type I subtree Tuu has weight \(2 \cdot\) LowPt[u]. A type II subtree Tuu has weight \(2 \cdot\) LowPt[U]+1. The adjacent vertices are then sorted by weight, which can be done by a bucket sort in linear time, because the weights are all in the range \(1, \ldots, 2 n+1\).
The general algorithm takes place in two stages. The first stage is LOWPTDFS() which constructs a DF-tree, calculates the low-points and second low-points, and sorts the adjacency lists. The second stage is another DFS which we call EMBEDDINGDFS(). It is a DFS using the re-ordered adjacency lists and is given as Algorithm 12.15.3
Algorithm 12.15.3 constructs an embedding by placing the edges into two linked lists Li and Lo, as in the simplified algorithm. The list Li which originally corresponded to the chords placed inside \(C\), now corresponds to the edges placed to the left of the DF-tree, since the drawings were made with the interior of \(C\) to the left
of the path. Similarly, the list Lo now corresponds to edges placed to the right of the DF-tree, since the exterior of \(C\) was drawn to the right of the path. EMBEDDINGDFS() first descends the DF-tree. The first leaf it encounters will have a frond back to vertex 1 . This is a consequence of the ordering of the adjacency lists. This creates a cycle \(C\), which we draw to the left of the path. The remaining fronds and subtrees will be placed either in Li or Lo, exactly as in the simplified algorithm.
A subtree Tuu is similar to a chord ( \(u, \operatorname{LowPt[U])\cdot Tuu~fits~in~Li~if~and~only~if~a~chord~(~} u, L o w P t[u]\) ) does. In this case, we place a dummy chord ( \(u, L o w P t[U]\) ) in \(L i\) and a dummy chord \((u, u)\) in Lo. A dummy chord is a chord that has an associated flag indicating that it is a placeholder for a subtree. If the dummy chord ( \(u\), LowPt[U]) is assigned to Li, the subtreeTuu is to be embedded to the left of the path of the DF-tree containing \(u\). The algorithm then calls EMBEDDINGDFS(U) recursively. The fronds placed in Li by the recursion are placed to the left of the tree Tuu. The fronds placed in Lo are to the right of Tuu. The dummy chord (u, \(u\) ) in Lo has the purpose of ensuring that any fronds placed in Lo by the recursive call must have UpperPt \(\geq u\). page_323

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Algorithm 12.15.3: EMBEDDI NGDFS(u)
comment: extend the embedding DFS from \(u\)
for all \(u\) adjacent to \(u\)
(if \(u v\) is a frond and \(v\) is above \(u\)
\(\left\{\begin{array}{l}\text { if }(u, v) \text { fits in } L_{i} \\ \text { then place }(u, v) \text { in } L_{i} \\ \text { else if }(u, v) \text { fits in } L_{o} \\ \text { then place }(u, v) \text { in } L_{o}\end{array}\right.\)
then
else
\(m=\operatorname{SWITCHSIDES}(u, v)\)
if \(m=0\)
then \(\left\{\begin{array}{l}\text { NonPlanar }=\text { true } \\ \text { exit }\end{array}\right.\)
else if \(m=1\)
then place \((u, v)\) in \(L_{i}\)
else if \(m=-1\)
then place \((u, v)\) in \(L_{o}\)


The important area of graph minors was developed in a series of over 20 papers by Robertson and Seymour. Some of their early work on graph minors is surveyed in their paper ROBERTSON and SEYMOUR [103]. They have proved a theorem of far-reaching significance, that in any infinite collection of graphs, there are always two graphs such that one is a minor of the other; or in other words, any set of graphs in which no graph is a minor of another, is finite. The books by DIESTEL [35] and ZIEGLER [129] contain excellent chapters on graph minors.
Rotation systems were developed by HEFFTER [60] and EDMONDS [37].
Read's algorithm to draw a planar graph by reducing a triangulation is from READ [100]. It was modified to use a regular polygon as the outer face by KOCAY and PANTEL [79]. Tutte's method of using barycentric coordinates to construct convex drawings of graphs appeared in TUTTE [116].
Whitney's theorem appeared in WHITNEY [125].
Good source books for polytopes are the books by GRÜNBAUM [56] and ZIEGLER [129].
The original proof of the four-color theorem appeared in APPEL and HAKEN [4] and [5]. An excellent survey article of the Appel-Haken proof is WOODALL and WILSON [128]. There are a number of excellent books on the 4-color problem, including SAATY and KAINEN [106] and ORE [93]. A very readable history of the 4-color problem can be found in WILSON [127]. A shorter proof was accomplished by ROBERTSON, SANDERS, SEYMOUR, and THOMAS in [104]. Much of the development of graph theory arose out of attempts to solve the 4-color problem. AI GNER [2] develops the theory of graphs from this perspective.
Kuratowski's theorem is a famous theorem of graph theory. It originally appeared in KURATOWSKI [82]. The proof presented here is based on a proof of KLOTZ [73]. See also THOMASSEN [112].
The Hopcroft-Tarjan planarity algorithm is from HOPCROFT and TARJ AN [66]. See also WILLIAMSON [126]. It is usually presented as a "path-addition" algorithm; that is, an algorithm that embeds one path at a time across a cycle. It is presented here as an equivalent algorithm that recursively embeds a branch of the DFtree.

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\section*{Graphs and Surfaces}

\subsection*{13.1 I ntroduction}

The plane and the sphere are the simplest topological surfaces. The structure of planar graphs, and algorithms for embedding graphs on the plane are well understood. Much less is known about graph embeddings on other surfaces, and the structure of these graphs. We begin with the torus, the doughnutshaped surface shown in Figure 13.1. We imagine this surface made out of rubber, and using scissors, cut it along the two circumferences shown in the diagram. The surface of the torus then unfolds into a rectangle, which is indicated on the right. The opposite sides of the rectangle labeled a must be glued together with the arrows aligned, as must the sides labeled \(b\), in order to reconstruct the torus. We could glue the edges in the order \(a\), then \(b\); or else \(b\), then \(a\). Both represent the same torus.


\section*{FIGURE 13.1 The torus}

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When a graph is drawn on the rectangle representing the torus, we must remember that the two sides labeled a (and the two sides \(b\) ) are really the same, so that graph edges can "wrap around" the diagram.

Notice that the four corners of the rectangle all represent the same point. Figure 13.2 shows two embeddings of \(K 3,3\) on the torus.

(i)

(ii)

\section*{FI GURE 13.2}

\section*{Two embeddings of \(K 3,3\) on the torus}

These embeddings of \(K 3,3\) are very different from each other. Unlike the plane in which a 3 -connected graph has a unique embedding (up to orientation), some graphs have very many distinct embeddings in the torus, or other surfaces.
DEFINITION 13.1: An embedding of a graph \(G\) in a surface \(\Sigma\) is a function \(\psi\) that maps the vertices of \(G\) into points of \(\Sigma\), and the edges of \(G\) into continuous curves in \(\Sigma\), such that the curves representing two edges intersect only at a common endpoint. We write \(G \psi\) for the image of \(G\) under the embedding \(\psi\).
In the embeddings of Figure 13.2, we can assign a "coordinate system" to the rectangle representing the torus, and then construct \(\psi\) by assigning coordinates to the vertices, and then draw the edges as straight lines. This is how the diagram was constructed.
Definition 12.1 uses an intuitive notion of a surface, and an intuitive notion of continuous. Currently we have the plane, sphere, or torus in mind. We will later make the definition of a surface more precise. Because we have coordinate systems for the above surfaces, by "continuous" we mean continuous mappings of the coordinates. However, topological continuity does not require coordinates.
If we cut the torus along the edges of the embeddings of \(K 3,3\), the torus surface falls apart into several connected regions. As in the case of the plane, we call these regions the faces of the embedding. A facial cycle is an oriented cycle of the graph which bounds a face. The embedding of \(K 3,3\) in Figure 13.2(i) has page_328

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three faces, each bounded by a hexagon. The embedding in Figure 13.2(ii) also has three faces, two bounded by quadrilaterals, and one bounded by a 10-cycle.
An open disc in the plane is the region interior to a circle. We will use an intuitive understanding of the notion of homeomorphic subsets of surfaces. Two regions \(R 1\) and \(R 2\) are said to be homeomorphic if one can be transformed into the other by a one-to-one continuous deformation, whose inverse is also continuous. For example, an open disc is homeomorphic to a face bounded by a hexagon, or by any other polygon. A
homeomorphism of \(R 1\) and \(R 2\) is any continuous one-to-one mapping from \(R 1\) onto \(R 2\), whose inverse is also continuous.
DEFINITION 13.2: A 2-cell is a region homeomorphic to an open disc. An embedding \(G \psi\) is a 2 -cell embedding if all faces are 2-cells.
Corresponding to the idea of a 2 -cell are 0 -cells (a single point), 1-cells (an open line segment), and 3-cells (the interior of a sphere in 3-space), etc.
It is necessary to restrict embeddings of a graph \(G\) to 2 -cell embeddings. For example, we could draw a planar embedding of a planar graph \(G\), such as the cube, in the rectangle representing the torus. The result would not be a 2 -cell embedding of \(G\), for the outer face of the embedding would be homeomorphic to a torus with a hole in it, which is not a 2-cell. Embeddings which are not 2 -cell embeddings really belong in a different surface. For the cube, there are five distinct 2-cell embeddings on the torus.
DEFINITION 13.3: Two embeddings \(G \psi 1\) and \(G \psi 2\) in a surface \(\Sigma\) are homeomorphic if there is a homeomorphism of \(\Sigma\) which maps \(G \psi 1\) to \(G \psi 2\). Otherwise \(G \psi 1\) and \(G \psi 2\) are distinct embeddings.
It is clear that a homeomorphism of \(G \psi 1\) and \(G \psi 2\) induces an automorphism of \(G\), and that the faces of \(G \psi 1\) map to the faces of \(G \psi 2\). The embeddings of Figure 13.2 are then easily seen to be distinct, because their facial cycles have different lengths. In general, there is no easy method of determining all the embeddings of a graph on a given surface, or even to determine whether a graph is embeddable.

\subsection*{13.2 Surfaces}

We can use the method of representing the torus as a rectangle, as in Figure 13.1, to represent a cylinder,
and a Möbius band, shown in Figures 13.3 and 13.4.
The cylinder is glued along only one edge of the rectangle. We call it an open surface because it has a boundary (the edges which are not glued). If we project a cylinder onto a plane, one possible result is an annulus, that is, a disc with a hole in it. This immediately gives the following theorem:

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FI GURE 13.3
Three representations of the cylinder
THEOREM 13.1 A graph can be embedded on the cylinder if and only if it can be embedded on the plane.
PROOF Given an embedding of a graph \(G\) on the plane, choose any face and cut a hole in it. The result is an embedding on the cylinder, and vice versa.
Notice that an embedding of \(G\) on the cylinder corresponds to an embedding of \(G\) on the plane with two distinguished faces. They can be any two faces of a planar embedding of \(G\).
The Möbius band is constructed by giving one end of the rectangle a twist of 180 degrees before aligning and gluing the opposite edges. This is indicated by the opposite orientation of the arrows. Notice that if we follow the boundary of the Möbius band, it is a single closed curve, unlike the boundary of the cylinder.
The sphere, torus, and cylinder can all be considered as two-sided surfaces-they have an inside and outside. One way to define this is to imagine a small clockwise-oriented circle drawn in the surface. If we reflect this circle in the
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\section*{FI GURE 13.4}

\section*{The Möbius band}
surface, we obtain a circle of the opposite orientation. On the sphere, torus, and cylinder it is not possible to walk along the surface, taking the oriented circle with us, until it coincides with its opposite orientation (we are not allowed to walk over the boundary of the cylinder or Möbius band). On the Möbius band, it is possible to make these two circles coincide. We therefore say that the Möbius band is a one-sided surface. A twosided surface is said to be orientable. We can assign an orientation to the surface, by partitioning the set of oriented circles defined at every point of the surface. One orientation is called the inside, and the other the outside. A one-sided surface is non-orientable, as the set of oriented circles does not have this partition into two subsets. For surfaces like the sphere and torus which can be constructed in euclidean 3 -space, we could alternatively use a normal vector to the surface, and its reflexion in the surface in place of an oriented circle and its reflexion.
Aside from the cylinder and Möbius band, the surfaces we will be interested in are closed surfaces. A closed surface is a generalized notion of a polyhedron. A polyhedron is a
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three-dimensional object consisting of a set of polygons, of three or more sides each. Each polygon is bounded by a sequence of \(p\) straight-line segments connecting \(p\) vertices in cyclic order, for some \(p \geq 3\). The line segments are the edges of the polyhedron. Each edge is shared by exactly two polygons. Any two polygons may intersect only on a single common edge. There are at least three polygons meeting at each vertex, and the polygons meeting at any vertex form a single cycle.
This idea is generalized by allowing the polygons composing a polyhedron to be infinitely stretchable and interpenetrable. They are then called curvilinear polygons.
DEFINITION 13.4: A closed surface is a set of points homeomorphic to a polyhedron made of curvilinear polygons.
Thus a closed surface is an object that is capable of being represented as a collection of curvilinear polygons glued together along common edges. However, the surface is not any single one of these representations, since many different polygonal representations of the same closed surface are possible. For example, the surface of the sphere may be partitioned into many different polygonal forms. Similarly, the embeddings of K 3 , 3 of Figure 13.2 partition the torus into two different polyhedra.
When we identify opposite edges of a rectangle, we are identifying two edges of a single polygon. In order to conform to the definition of polyhedron, in which only distinct polygons share a common edge, we can subdivide the rectangle into two or more polygons, as necessary. There are three more surfaces that can be made from a rectangle by identifying its edges in pairs. They are the sphere, projective plane, and Klein bottle, illustrated in Figures 13.5, 13.6, and 13.8. The sphere can also be represented as a digon (a polygon of two sides), as shown, where \(c=a b\).


FI GURE 13.5
The sphere as a rectangle and as a digon
The projective plane is constructed from a rectangle by first using a twist to page_332

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create a Möbius band, and then by gluing the boundary of the Möbius band to itself, to create a closed surface. This is illustrated in Figure 13.6, where the edges labeled \(b\) have been glued. This cannot be done in euclidean space, for the polygon must cross itself without intersecting itself. But mathematically, we can consider the surface to be constructed in this way. The projective plane can also be viewed as a digon, as illustrated in Figure 13.7, by combining the \(a\) and \(b\) sides of the rectangle into a single side, labeled \(c\). Because of the orientation of the arrows, the two "corners" of the digon represent the same point. We can identify the corners, creating a "figure eight", and then identify the two lobes of the figure eight to complete the projective plane. If we then remove a disc from the projective plane by cutting along the circle containing the dotted line, the result is called a crosscap. It is a projective plane with a hole in it. In Exercise 13.2.1, it is proved that a crosscap is homeomorphic to a Möbius band.


\section*{FI GURE 13.6}

\section*{The projective plane as a rectangle, and as a Möbius band}

The Klein bottle can be constructed by gluing two edges of a rectangle to create a cylinder, and then by gluing the ends of the cylinder together, according to the orientation of the arrows. This also cannot be done in eudidean space, as the cylinder must cross through itself without intersecting itself.
These surfaces have been constructed from a rectangle by gluing edges together. By subdividing the rectangle into curvilinear polygons, closed surfaces represented as curvilinear polyhedra are obtained. Polygons with more sides than a rectangle could also be used. By the classification theorem of closed surfaces (Theorem 13.2), every closed surface can be constructed by gluing together edges of a single curvilinear polygon.
The rectangle representing the torus in Figure 13.1 can be written symbolically as \(a+b+a-b-\). This means that we choose a clockwise orientation of the rectangle and write a+ for the edge labeled a when the direction of the arrow is clockwise, and a- when the direction of the arrow is counterclockwise. The boundary of the rectangle is then determined by the above symbol. Similarly the rectangle


FI GURE 13.7
The projective plane as a digon, and as a crosscap
representing the sphere (Figure 13.5 ) can be characterized as \(a+b+b-a-\), that of the projective plane (Figure 13.6) as \(a+b+a+b+\), and that of the Klein bottle (Figure 13.8) as \(a+b+a+b-\).
In general, we have a polygon with an even number of sides, with labels a1, a2,.., ap, such that every label ai appears on exactly two sides. We place an arrow on each edge of the polygon in some direction, and choose a clockwise orientation of the polygon. This defines a symbolic representation in terms of the \(a_{i}^{+}\)and \(a_{i}^{-}\). Every closed surface can be represented symbolically by a normal form of this type, as shown by the following theorem:
THEOREM 13.2 (Dehn and Heegard-Normal forms for closed surfaces) Every closed surface can be represented symbolically by one of the following normal forms. Two closed surfaces are homeomorphic if and only if they have the same normal form.
1. \(a+a-\)
2. \(a_{1}^{+} b_{1}^{+} a_{1}^{-} b_{1}^{-} a_{2}^{+} b_{2}^{+} a_{2}^{-} b_{2}^{-} \ldots a_{p}^{+} b_{p}^{+} a_{p}^{-} b_{p}^{-}\)
3. \(a_{1}^{+} a_{1}^{+} a_{2}^{+} a_{2}^{+} \ldots a_{q}^{+} a_{q}^{+}\)

The proof of this theorem can be found in FRÉCHET and FAN [44] or STILLWELL [109]. It is too lengthy to include here. It involves cutting and pasting a curvilinear polygon until the normal form is achieved. This is done in stages. The torus is represented in Figure 13.1 by \(a+b+a-b-\), which is already in normal form. The sphere is represented by \(a+b+b-a-\). It is clear from the diagram that the adjacent \(b+b-\) corresponds to edges that can be glued, so that they cancel from the formula, leaving \(a+a-\) as the normal form for the sphere. The projective

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\section*{FI GURE 13.8}

The Klein bottle as a rectangle
plane is represented in Figure 13.6 as \(a+b+a+b+\), which is not in normal form. The proof of Theorem 13.2 would convert it to normal form by the following sequence of operations, illustrated in Figure 13.9.


FI GURE 13.9

\section*{Transforming the projective plane to normal form}

We make a diagonal cut across the rectangle, and label it \(c\). We thereby obtain two triangles which we glue together along the edge \(b\), which then disappears. The symbolic form is now \(c+a+a-c+\). The edges \(a+a-\) are then glued to produce the digon in Figure 13.9, giving the normal form \(c+c+\) for the projective plane.

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\subsection*{13.2.1 Handles and crosscaps}

Consider a surface with normal form \(a+b+a-b-c+d+c-d-\), shown in Figure 13.10. We have an octagon with edges that are to be identified in pairs. Both endpoints of the upper edge marked \(b\) represent the same point, as they are the same endpoint of the arrows marked \(a\). Therefore gluing the edges labeled \(a\) will make this \(b\)-edge the boundary of a hole in the surface. Consequently, the other edge labeled \(b\) also represents the boundary of a hole in the surface, and these two boundaries must be identified. One way to identify them is to attach the two ends of a cylindrical tube to each of these holes. The result is a handle attached to the surface.


FI GURE 13.10

\section*{A sphere with two handles}

Now the same can be done for the edges marked \(c\) in the diagram-we attach another handle. This same argument holds for any normal form of this type. This gives:
THEOREM 13.3 A surface with normal form
\[
a_{1}^{+} b_{1}^{+} a_{1}^{-} b_{1}^{-} a_{2}^{+} b_{2}^{+} a_{2}^{-} b_{2}^{-} \ldots a_{p}^{+} b_{p}^{+} a_{p}^{-} b_{p}^{-}
\]
is homeomorphic to a sphere with \(p\) handles.
Because a sphere is an orientable surface, so is a sphere with \(p\) handles.
Consider now a surface with normal form \(a+a+b+b+c+c+\), illustrated in Figure 13.11. We have a hexagon in which consecutive edges are to be identified in pairs. The vertex common to the two sides marked \(a\) is both the head and tail of the a-arrow. Therefore indentifying the endpoints of the edges marked \(a\)

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makes two holes in the surface, bounded by the a-edges. We must identify the boundaries of these two lobes.


\section*{FI GURE 13.11}

\section*{A sphere with three crosscaps}

This is nearly identical to the construction of the projective plane from a digon, illustrated in Figure 13.7. If we draw a circle around the two a-lobes in Figure 13.11, and cut along the dotted line, we see that what we have is in fact a crosscap glued into a hole cut in the surface. We can then do the same for the remaining pairs \(b+b+\) and \(c+c+\) to get a sphere with three crosscaps. In general, this gives:
THEOREM 13.4 A surface with normal form
\[
a_{1}^{+} a_{1}^{+} a_{2}^{+} a_{2}^{+} \ldots a_{q}^{+} a_{q}^{+}
\]
is homeomorphic to a sphere with \(q\) crosscaps.
Because a crosscap is an unorientable surface, so is a sphere with \(q\) crosscaps. Gluing a crosscap to a hole in a sphere is equivalent to gluing the boundary of a Möbius band to a hole in the sphere.

\subsection*{13.2.2 The Euler characteristic and genus of a surface}

We know from Chapter 12 that a connected planar graph with \(n\) vertices, \(\varepsilon\) edges, and \(f\) faces satisfies Euler's formula \(n-\varepsilon+f=2\). Furthermore, the skeleton of a polyhedron is a planar graph, so that any polyhedral division of the sphere also satisfies this formula. We say that the Euler characteristic of the sphere is 2.

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DEFINITION 13.5: Let \(\Sigma\) be a closed surface represented by a curvilinear polyhedron with \(n\) vertices, \(f\) polygons, and \(\varepsilon\) edges. The Euler characteristic of the surface is the value \(n-\varepsilon+f\). It is denoted \(\chi(\Sigma)\). It has been proved by Kerékjártó that the value \(n-\varepsilon+f\) for a surface is invariant, no matter what polygonal subdivision is used to represent it. This is difficult to prove because of the vast numbers of polygonal subdivisions that are possible. However, we can get an understanding of it as follows. If we add a diagonal across a face, \(n\) does not change, but \(\varepsilon\) and \(f\) both increase by one. Thus \(n-\varepsilon+f\) does not change. Similarly, if we subdivide an edge with a new vertex, \(n\) and \(\varepsilon\) increase by one, but \(f\) does not change. Modifying a polygonal subdivision by these operations does not change the value \(n-\varepsilon+f\). Suppose that we now add a handle or crosscap to a given polygonal division of a surface. Consider a polygonal division with \(n\) vertices, \(f\) polygons, and \(\varepsilon\) edges. Choose two polygons on the surface, cut a disc from the interior of each polygon, and attach a handle connecting them. Let the polygons be \(P 1\) and \(P 2\). Refer to Figure 13.12. We draw two curves along the handle connecting \(P 1\) to \(P 2\). The result is a polygonal division of the surface with an additional handle. Let it have \(n^{\prime}\) vertices, \(f^{\prime}\) polygons, and \(\varepsilon^{\prime}\) edges. The effect of drawing the curves connecting \(P 1\) to \(P 2\) is to add two vertices and two edges to each of \(P 1\) and \(P 2\), plus two additional edges represented by the curves. The number of polygons does not change. We therefore have \(n^{\prime}=n+4, f^{\prime}=f\), and \(\varepsilon^{\prime}=\varepsilon+6\), so that \(n^{\prime}-\varepsilon\) \('+f^{\prime}=(n-\varepsilon+f)-2\). Thus, when a handle is added to a surface, the Euler characteristic decreases by two. It does not matter to which faces the handle attaches. It follows that a sphere with \(p\) handles has Euler characteristic \(2-2 p\).

\title{
\(P_{1}\) \\ \(P_{2}\)
}

FI GURE 13.12
Attaching a handle to a surface

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Suppose that we now add a crosscap to a polygonal division of a surface (e.g., the sphere). We choose a polygon P1, cut a disc from its interior, and attach a crosscap. Let \(C\) be the boundary of the disc. Now a crosscap is a Möbius band, and the boundary of the disc becomes the boundary of the Möbius band. We draw a curve connecting \(P 1\) to \(C\), continue across the Möbius band to the opposite side of \(C\), and continue to an opposite point of \(P 1\). The result is a polygonal division of the surface with an additional crosscap. Let it have \(n\) ' vertices, \(f^{\prime}\) polygons, and \(\varepsilon^{\prime}\) edges. The effect of drawing the curve across \(P 1\) is to add two vertices and two edges to \(P 1\), plus an additional edge represented by the curve. When the Möbius band is cut in half, it remains connected. Therefore the number of polygons does not change. We therefore have \(n^{\prime}=n+2, f^{\prime}=f\), and \(\varepsilon^{\prime}=\varepsilon+3\), so that \(n^{\prime}-\varepsilon^{\prime}+f^{\prime}=(n-\varepsilon+f)-1\). Thus, when a crosscap is added to a surface, the Euler characteristic decreases by one. It does not matter to which face the crosscap attaches. It follows that a sphere with \(q\) crosscaps has Euler characteristic \(2-q\).
Consequently surfaces can be classified according to whether they are orientable or non-orientable, and their Euler characteristic. A related parameter of a surface is its genus.
DEFINITION 13.6: A Jordan curve in a surface \(\Sigma\) is contractible or null-homotopic if it can be continuously shrunk to a point within \(\Sigma\).
Cutting a surface along a contractible J ordan curve always separates it into two pieces.
DEFINITION 13. 7: The genus of a surface is the maximum number of J ordan curves that can be drawn on the surface such that cutting along the curves does not separate it into two or more pieces.
It is easy to see that the sphere has genus zero, and that the torus has genus one. In general, a sphere with \(p\) handles has genus \(p\), as exactly one J ordan curve can be drawn around each handle without separating the surface. Because a sphere with handles is an orientable surface, we say it has orientable genus \(p\).
The projective plane has genus one, as shown in Exercise 13.2.2. It then follows from Exercise 13.2.3 that a sphere with \(q\) crosscaps has genus \(q\). We say it has unorientable genus \(q\). Some texts use the term crosscap number in place of genus for a non-orientable surface. The relation with Euler characteristic can now be stated.
THEOREM 13.5 An orientable surface of genus p has Euler characteristic 2-2p. A non-orientable surface of genus \(q\) has Euler characteristic 2-q.
When a graph \(G\) is embedded in a surface \(\Sigma\), cycles of \(G\) map to Jordan curves in \(\Sigma\). We will be interested in cycles which are embedded as non-contractible curves.
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DEFINITION 13.8: Let \(G \psi\) be an embedding of a graph \(G\) in a surface \(\Sigma\). A cycle \(C\) in \(G\) is an essential cycle, or non-contractible cycle of the embedding if \(G \psi\) is not contractible.
For example, in the embedding of \(K 3,3\) on the left in Figure 13.2 , the cycles \((1,2,3,4)\) and \((2,3,4,5)\) are essential cycles, while ( \(1,2,3,6,5,4\) ) is not.
DEFINITION 13.9: The genus of a graph \(G\) is \(g(G)\), the smallest genus of an orientable surface \(\Sigma\) such that \(G\) has a 2-cell embedding in \(\Sigma\). The unorientable genus or crosscap number of \(G\) is \(\bar{g}(G)\), the smallest genus of a non-orientable surface \(\Sigma\) such that \(G\) has a 2 -cell embedding in \(\Sigma\).
Suppose that \(G \psi\) is a 2-cell embedding of \(G\) in an orientable surface \(\Sigma\) of genus \(p\). Let \(G \psi\) have n vertices, \(\varepsilon\) edges, and \(f\) faces. Then since the faces of \(G \psi\) determine a polygonal division of the surface, we have
\(n-\varepsilon+f=x(\Sigma)=2-2 p\). This is called the Euler-Poincaré formula. If \(G\) is embedded on an unorientable surface of genus \(q\), then \(n-\varepsilon+f=\chi(\Sigma)=2-q\). This gives the following relations for graphs embedded on the:
plane: \(n-\varepsilon+f=2\)
torus: \(n-\varepsilon+f=0\)
projective plane: \(n-\varepsilon+f=1\).
LEMMA 13.6 A triangulation of an orientable surface of genus \(p\), with \(n\) vertices satisfies \(\varepsilon=3 n+6(p-1)\). A triangulation of a non-orientable surface of genus \(q\) satisfies \(\varepsilon=3 n+3(q-2)\).
PROOF A triangulation satisfies \(3 f=2 \varepsilon\). Combining this with \(n-\varepsilon+f=\chi(\Sigma)\) gives the result.

\section*{Exercises}
13.2.1 Show that if a disc is removed from a projective plane, the result is a Möbius band.
13.2.2 Show that if a the projective plane is cut along a non-separating Jordan curve, the result is a disc.
13.2.3 Show that exactly one J ordan curve can be drawn on the Möbius band without separating it. What is the result of cutting the Möbius band along a non-separating J ordan curve?
13.2.4 Use cutting and pasting to convert the representation \(a+b+a+b-\) of the Klein bottle to normal form \(c+c+d+d+\).
13.2.5 A sphere with one handle and one crosscap can be represented symbolically by \(a+b+a-b-c+c+\). Find the normal form for this surface.

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13.2.6 Find the facial cycles of the embeddings of \(K 3,3\) on the torus shown in Figure 13.2.
13.2.7 Use Lemma 13.6 to obtain a lower bound on \(g(G)\) and \(\bar{g}(G)\) for an arbitrary graph \(G\).
13.2.8 Show that \(g(K n) \geq(n-3)(n-4) / 12\) and that \(\bar{g}\left(K_{n}\right) \geq(n-3)(n-4) / 6\).
13.2.9 Let \(G\) be a graph with no triangles embedded on a surface of genus \(g\). Find an upper bound on the number of edges of \(G\).

\subsection*{13.3 Graph embeddings, obstructions}

Three of the main algorithmic problems of graph embeddings are:
Problem 13.1: Graph Embeddability
Instance: a graph \(G\) and a surface \(\Sigma\).
Question: is \(G\) embeddable in \(\Sigma\) ?
Problem 13.2: Graph Embeddings
Instance: a graph \(G\) and a surface \(\Sigma\).
Question:find all embeddings of \(G\) in \(\Sigma\).
Problem 13.3: Graph Genus
I nstance: a graph \(G\) and an integer \(k\).
Question: is \(g(G) \leq k\) ?
It was proved by THOMASSEN [113] that Graph Genus is NP-complete.
The first two problems are solved for the plane, but only partially solved for other surfaces. Several efficient algorithms are known for Graph Embeddability on the projective plane. For the plane, Kuratowski's theorem tells us that \(G\) is embeddable if and only if it has no subgraph \(T K 5\) or \(T K 3,3\). We call these graphs obstructions to planarity.
DEFINITION 13.10: Given a surface \(\Sigma\), a topological obstruction for \(\Sigma\) is a minimal graph \(K\) with \(\delta(K) \geq 3\) such that any graph containing a subdivision \(T K\) cannot be embedded in \(\Sigma\).
There are two topological obstructions for the plane, \(K 5\) and \(K 3,3\). The definition requires that \(K\) be minimal, namely, that no proper subgraph of \(K\) has this
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property (otherwise \(K 6, K 7\), etc., would all be considered as topological obstructions).
DEFINITION 13.11: Given a surface \(\Sigma\), a minor-order obstruction (or minimal forbidden minor or excluded minor) for \(\Sigma\) is a graph \(K\), such that any graph having \(K\) as a minor cannot be embedded in \(\Sigma\), but no proper minor of \(K\) has this property.
Recall that the graph relation " \(H\) is a minor of \(G\) " forms a partial order on the set of all graphs. If we restrict the set to all graphs which are not embeddable in the surface \(\Sigma\), then the minimal graphs of this partial order are the minor-order obstructions. If \(K\) is a minor-order obstruction, then it is also a topological obstruction. \(K 5\) and \(K 3,3\) are both minor-order obstructions and topological obstructions for the plane, since any graph which has \(K 5\) or \(K 3\), 3 as a minor necessarily contains either a \(T K 5\) or \(T K 3\), 3 . For other surfaces, there is a distinction between the two concepts of topological obstruction and minor-order obstruction.
Robertson and Seymour have proved that there are a finite number of obstructions for any surface, as a consequence of the graph minor theorem, which we state without proof.
THEOREM 13.7 (Graph minor theorem) In any infinite collection of graphs, there are always two graphs such that one is a minor of the other.

It is known that there are 103 topological obstructions for the projective plane, of which 35 are minor-order obstructions, as found by ARCHDEACON [6]. These are often called Kuratowski subgraphs for the projective plane. A list of them can be found in MOHAR and THOMASSEN [88]. For the torus, the number of obstructions is in the hundreds of thousands, as shown by MYRVOLD [90]. From an algorithmic point of view, this is not an effective characterization, as there are too many obstructions.

\subsection*{13.4 Graphs on the torus}

Given a 2 -cell embedding \(\psi\) of a 2 -connected graph \(G\) on the torus, there must be an essential cycle \(C\) in \(G\). Cutting the torus along \(C \psi\) results in a cylinder. Since the cylinder is not a 2 -cell, but the embedding is a 2cell embedding, there must be another essential cycle \(C^{\prime}\) in \(G\), cutting the cylinder along an axis.
Consequently \(C\) and \(C^{\prime}\) must intersect, either in a path or a vertex.
DEFINITION 13.12: A theta-graph is a graph consisting of two vertices of degree three, connected by three paths of one or more edges each.

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A theta-graph is illustrated in Figure 13.13.


\section*{FI GURE 13.13}

\section*{A theta-graph in schematic form}

Thus, \(C U C^{\prime}\) must be either a theta-subraph of \(G\), or two cycles with a vertex in common, and \(\psi\) is a 2 -cell embedding of it. The simplest form of theta-subgraph is a multigraph consisting of two vertices connected by three parallel edges. A 2-cell embedding of it is shown in Figure 13.14. It is often necessary to consider embeddings of graphs with multiple edges and/or loops, as the duals of many graph embeddings (e.g., K4, \(K 5, K 3,3\) ) often have multiple edges, and sometimes loops. We shall always insist that in any embedding of a multigraph:
1. The cycle induced by any loop is an essential cycle (no face is a loop).
2. The cycle induced by any digon is an essential cycle (no face is a digon).


FI GURE 13.14

\section*{A 2-cell embedding of a theta-graph}

If we cut the torus along the edges of this theta-graph, we find there is one page_343

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face, a hexagon, shown in Figure 13.14 as \(c+d+e+c-d-e-\). Thus we see that the torus can also be represented as a hexagon. The hexagonal form corresponds to an embedding of a theta-subgraph. The rectangular form corresponds to an embedding of two cycles with a common vertex. Given an embedding \(G \psi\) on the torus, we choose an orientation of the torus, and walk around a vertex \(u \psi\) in a small clockwise circle in the surface, and construct the cyclic adjacency list, just as for embeddings in the plane (Section 12.5). This determines a rotation system for \(G\), exactly as in the planar case. We will denote a rotation system for a graph embedded on the torus by \(t\). The faces of the embedding are completely determined by \(t\), since Algorithm 12.5.1, \(\operatorname{FACI} \operatorname{ALCYCLE}()\), to find the facial cycles of a planar graph from its rotation system also applies to toroidal rotation systems, or to rotation systems for any orientable surface. Similarly, algorithm CONSTRUCTDUAL() applies equally to toroidal rotation systems as well as other orientable surfaces. Hence we denote a combinatorial toroidal embedding by \(G t\) and its dual by \(G t^{*}\). Now the rotation
system determines the faces of the embedding. Hence, we can determine from \(t\) whether or not \(G t\) has any faces which are digons or loops, but it cannot determine whether any digons or loops are embedded as essential cycles.
It is convenient to refer to a graph embedded on the torus as a torus map.
DEFINITION 13.13: A torus map is a combinatorial 2-cell embedding \(G t\), where \(G\) is a 2-connected graph. We begin by embedding planar graphs on the torus. Let \(G\) be a 2-connected planar graph that is not a cycle, with a planar rotation system \(p\). By Exercise 13.4.3, \(G\) has a theta-subgraph \(H\). Let \(u\) and \(u\) be the two vertices with degree three in \(H\), and let P1, P2, P3 be the three uu-paths of \(H\). Let w1, w2, and w3 be the first vertices of \(P 1, P 2, P 3\), respectively, adjacent to \(u\). Refer to Figure 13.15.
THEOREM 13.8 Let p be a planar rotation system for \(G\), and let \(H\) be a thetasubgraph of \(G\), as described above. Let \(t\) be a rotation system constructed from \(p\) by interchanging \(w 2\) and \(w 3\) in the cyclic adjacency list of \(u\), and leaving all other vertices the same. Then \(t\) is a toroidal rotation system for \(G\).
PROOF In the embedding \(G p\) in the plane, the three paths \(P 1, P 2, P 3\) divide the plane into three regions. Without loss of generality, let the paths occur in the order illustrated in Figure 13.15. Denote the subgraphs of G contained within the three regions as G12 (between paths P1 and P2), G23, and G31. Subgraph Gij may have edges connecting it only to Pi and \(P j\), and to \(u\) and \(u\). Construct the hexagonal representation of the torus with \(P 1, P 2\), and \(P 3\) on the boundary of the hexagon, and embed G12, G23, and G31 inside the hexagon as planar embeddings, as shown, resulting in a toroidal embedding of \(G\). It is easy to verify from the diagram that any vertex in G12, G23, and G31 has the same cyclic adjacency list
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in the toroidal embedding as in the planar embedding. Similarly any vertex other than \(u\) or \(u\) of any Pi has the same cyclic adjacencies in both embeddings. The same is also true for \(u\). The only vertex whose cyclic adjacencies differ is \(u\). The adjacency list of \(u\) has been arranged so that \(w_{1} \in P_{1}\) is followed by the edges to \(G 12\), followed by \(w_{3} \in P_{3}\), followed by the edges to \(G 23\), followed by \(w_{2} \in P_{2}\), followed by the edges to \(G 31\). The only difference to the planar adjacencies is that \(w 2\) and \(w 3\) have been interchanged for vertex \(u\). It is evident from Figure 13.15 that there are several other ways to convert the planar rotation system to a toroidal rotation system.


\section*{FI GURE 13.15}

\section*{Constructing a toroidal rotation system}

A planar graph also has non-2-cell embeddings on the torus. If a planar graph \(G\) is embedded in a disc on the surface of the torus, we will call this a disc embedding of \(G\). If a planar graph is embedded such that one of the faces is homeomorphic to a cylinder, we will call this a cylindrical embedding of G. A cylindrical embedding on the torus of any graph \(G\) determines an embedding on the cylinder, so that \(G\) must be planar. LEMMA 13.9 Every embedding of a non-planar graph on the torus is a 2-cell embedding.
PROOF A non-2-cell embedding would necessarily be either a disc or cylindrical embedding. But a non-planar graph has no disc or cylindrical embedding.
Currently, there is no satisfactory algorithm known to determine whether an arbitrary graph can be embedded on the torus, or to find all embeddings, or to
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characterize all possible embeddings. Whitney's theorem (12.20) on induced non-separating cycles does not apply to embeddings on the torus. There are several simple techniques that are useful in an exhaustive search to find the embeddings. Given two combinatorial embeddings with toroidal rotation systems \(t 1\) and \(t 2\), we need to distinguish whether \(G^{t_{1}}\) and \(G^{t_{2}}\) are equivalent embeddings. In general, two embeddings are considered equivalent if they have the same facial cycles, as the faces can be glued together along the facial
boundaries in a unique way to construct the torus. Since the facial cycles are completely determined by the rotation system, we define equivalence in terms of rotation systems. Definitions 12.16 and 12.17 of equivalent embeddings and graph orientability apply equally well to toroidal graphs as to planar graphs, using a toroidal rotation system \(t\) in place of a planar rotation system \(p\). We summarize the definitions as follows:
DEFINITION 13.14:
1. Embeddings \(G^{\psi_{1}}\) and \(G^{\psi_{2}}\) are homeomorphic embeddings if there is a homeomorphism of the torus mapping \(G^{\psi_{1}}\) to \(G^{\psi_{2}}\). Otherwise they are distinct.
2. Embeddings \(G^{t_{1}}\) and \(G^{t_{2}}\) are isomorphic if there is an automorphism of \(G\) which induces a mapping of \(t 1\) to \(t 2\).
3. Embeddings \(G^{t_{1}}\) and \(G^{t_{2}}\) are equivalent embeddings if there is an automorphism of \(G\) which induces a mapping of \(t 1\) to \(t 2\) or \(\bar{t}_{2}\), where \(\bar{t}_{2}\) is obtained by reversing the cycles of \(t 2\).
4. Embedding \(G t\) is a non-orientable embedding if there is an automorphism of \(G\) inducing a mapping of \(t\) to \(\bar{t}\). Otherwise it is an orientable embedding.
Now an embedding \(G^{\psi_{1}}\) determines a rotation system \(t 1\). Homeomorphic embeddings \(G^{\psi_{1}}\) and \(G^{\psi_{2}}\) determine equivalent combinatorial embeddings \(G^{t_{1}}\) and \(G^{t_{2}}\), since a homeomorphism can be either orientation preserving or orientation reversing. Conversely, if \(G^{t_{1}}\) and \(G^{t_{2}}\) are equivalent combinatorial embeddings of \(G\), then they have the same facial cycles (up to orientation). The facial cycles can be glued together to construct a curvilinear polyhedron representing the torus. Therefore, topological embeddings \(G^{\psi_{1}}\) and \(G^{\psi_{2}}\) can be constructed from \(G^{t_{1}}\) and \(G^{t_{2}}\), so that \(G^{\psi_{1}}\) and \(G^{\psi_{2}}\) are homeomorphic. This gives:
THEOREM 13.10 Topological embeddings \(G^{\psi_{1}}\) and \(G^{\psi_{2}}\) are homeomorphic if and only if the corresponding combinatorial embeddings \(G^{t_{1}}\) and \(G^{t_{2}}\) are equivalent.
Now the homeomorphism between \(G^{\psi_{1}}\) and \(G^{\psi_{2}}\) was constructed by gluing curvilinear polygons (the faces of the embeddings) together along common edges; that is, it involves cutting and pasting the torus. For example, the two embeddings
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\(G^{t_{1}}\) and \(G^{t_{2}}\) shown in Figure 13.16 are equivalent, which can easily be verified from the rotation systems. However, they are homeomorphic only by cutting the torus along a non-contractible cycle to create a cylinder, then twisting one end of the cylinder by 360 degrees, and then re-gluing the cylinder to create a torus. It is not possible to transform one into the other without cutting and pasting.



FI GURE 13.16

\section*{Two equivalent embeddings}

For graphs on a small number of vertices, it is possible to distinguish inequivalent embeddings by inspection. However, even for \(K 5\) and \(K 6\), it is reasonably difficult to determine the inequivalent embeddings by hand. One technique that helps is the dual graph-if \(G^{t_{1}}\) and \(G^{t_{2}}\) have non-isomorphic dual graphs, then the embeddings are distinct. More generally, we can use the medial digraph (Definition 12.19) to distinguish embeddings and orientations. It can also be used to determine the symmetries (automorphisms) of an embedding. The medial digraph was defined for planar rotation systems, but the definition is also valid for toroidal rotation systems. We use \(M(G t)\) to denote the medial digraph of \(G\) with a toroidal rotation system \(t\). The medial digraph was defined for multigraphs. If we want to allow for loops as well, then the definition must be modified slightly (Exercise 13.4.7). Usually graph isomorphism software is necessary to make effective use of the medial digraph.
Suppose that \(G\) is a 2-connected planar graph. Choose a theta-subgraph \(H\) of \(G\). We would like \(H\) to have as many edges as reasonably possible. We can do this by hand for small graphs. With larger graphs, a depth-
first search can be used to find a theta-subgraph with a large number of edges. Every embedding of \(G\) on the torus induces an embedding of \(H\). It will be either a 2 -cell embedding, a cylindrical embedding, or a disc embedding. We start with a 2-cell embedding of \(H\), and proceed as in Theorem 13.8 to find all ways of extending the embedding of \(H\) to \(G\). This usually gives a number of embeddings. We then proceed to the cylindrical and disc embeddings of \(H\). In each case, all possible ways of extending \(H\) to a 2-cell embedding must be exhaustively considered. For each
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embedding \(t\), we construct \(M(G t)\), and compare the medial digraphs found for isomorphism, using graph isomorphism software.
If \(G\) is a 3 -connected non-planar graph, we proceed recursively. We choose a vertex \(u\) and find all embeddings of \(G-u\). Let the adjacent vertices to \(u\) be \(u 1, u 2, \ldots, u k\). If \(G-u\) is non-planar, then every embedding of it is a 2 -cell embedding. If \(u 1, u 2, \ldots, u k\) are all on the same facial cycle in some embedding of it, we can add vertex \(u\) to get an embedding of \(G\), possibly in several ways. If \(G-u\) is planar, instead we first find a TK3, 3 in \(G\) with as many edges as reasonably possible (assuming a TK3, 3 exists). For each
embedding of \(T K 3,3\) in the torus, we exhaustively consider all possible ways of extending it to an embedding of \(G\), and then use medial digraphs to compare the results.
For example, consider the graph \(K 4\). If \(u u\) is an edge of \(K 4\), then \(K 4-u u\) is a theta-graph. We easily find that there are exactly two 2 -cell embeddings of \(K 4\) on the torus. We then consider \(K 5\), one of the few non-planar graphs that does not contain \(T K 3\), 3. If \(u\) is a vertex of \(K 5\), then \(K_{5}-v \cong K_{4}\). For each embedding of \(K 4\) on the torus, including cylindrical and disc embeddings, we find all ways of adding \(u\) to the embedding. The result is six embeddings of \(K 5\), of which three are orientable and three non-orientable. We proceed to \(K 6\) by looking for a face of \(K 5\) containing all five vertices. We find there are four embeddings of \(K 6\), of which two are orientable and two non-orientable. Exactly one of these has all six vertices on a common face. This gives one embedding of \(K 7\), shown in Figure 13.17. It is an orientable embedding. Its dual is also shown. The dual is known as the Heawood graph.


FI GURE 13.17
K7 and its dual, the Heawood graph, on the torus
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\section*{Exercises}
13.4.1 Use cutting and pasting to convert the representation \(c+d+e+c-d-e-\) of the torus to normal form.
13.4.2 Construct the duals of the embeddings of \(K 3,3\) on the torus of Figure 13.2.
13.4.3 Show that every 2-connected graph that is not a cycle has a theta-subgraph.
13.4.4 Describe \(O(\varepsilon)\) depth-first and breadth-first search algorithms to find a thetasubgraph of a 2-connected graph \(G\).
13.4.5 Construct the two distinct embeddings of \(K 3,3\) on the hexagonal form of the torus.
13.4.6 Verify that the two embeddings shown in Figure 13.16 are equivalent, and that the torus must be cut and pasted to construct a homeomorphism.
13.4.7 Show how to modify the definition of a medial digraph to allow for embeddings with loops (subdivide a loop with two vertices), and prove that it works.
13.4.8 Construct all distinct embeddings of \(K 4, K 5\), and \(K 6\) on the torus.
13.4.9 Construct all distinct embeddings of the 3 -prism on the torus. Begin with a thetagraph containing all six vertices.
13.4.10 Construct all distinct embeddings of the Petersen graph on the torus. Begin with a theta graph containing all 10 vertices.
13.4.11 Determine which graphs are shown in the toroidal embeddings of Figure 13.18. Determine the dual
graphs. (Note: None of the graph edges follow the boundary of the rectangles.)
13.4.12 Verify that the graphs in Figure 13.19 are distinct embeddings of the same graph. Do you recognize this graph? Are these embeddings orientable? Find the duals of both embeddings, and determine what graphs they are. (Note: None of the graph edges follow the boundary of the rectangles.)


FI GURE 13.18

\section*{Two torus maps}

\subsection*{13.4.1 Platonic maps on the torus}

The embedding of \(K 7\) in Figure 13.17 shows a triangulation of the torus in which each vertex has degree six. By translating it repeatedly horizontally and vertically,

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\section*{FI GURE 13.19}

\section*{Two torus maps}
we obtain a symmetric tiling of the plane by triangles. Its dual gives a symmetric hexagonal cover of the
plane in which three hexagons meet at each vertex. The graphs in Figure 13.19 also give hexagonal coverings
of the plane. Their duals will be 6-regular triangulations. The graphs of Figure 13.18 give symmetric tilings by parallelograms. These embeddings all belong to families of graphs with these properties. We will call them Platonic maps.
Let \(G\) be a \(k\)-regular torus map on \(n\) vertices whose dual map is \(\ell\)-regular. For the plane, such graphs are the graphs of the Platonic solids. Then \(n k=2 \varepsilon=\ell f\). Using Euler's formula for the torus \(n+f-\varepsilon=0\), we obtain
\[
\frac{1}{k}+\frac{1}{\ell}=\frac{1}{2}
\]

The only integral solutions are \((k, \ell)=(4,4),(3,6)\), and \((6,3)\). Clearly the last two are duals of each other. The graphs of Figure 13.18 are examples of the (4,4)-pattern.
Consider a torus map \(G t\) in which each vertex has even degree. Choose any edge \(u u\). The incident edges at \(u\) are cyclically ordered by \(t\). Let \(D E G(u)=2 i\). The diagonally opposite edge to \(u u\) is \(u w\), the ith edge following \(u u\) in \(t(u)\). Given a vertex \(u 0\), with adjacent vertex \(u 1\), we construct a diagonal path \(u 0, u 1, u 2, \ldots\) by always choosing vivi+1 as the diagonally opposite edge to \(u i-1 v i\). Eventually a vertex must repeat, creating a cycle. Let the cycle be \(C=(u 0, u 1, U 2, \ldots, u m) . C\) is a diagonal cycle if every edge is the diagonally opposite edge to its previous edge.
In the torus maps of \(K 7\) (Figure 13.17) and Figure 13.18 there are many diagonal cycles. They are drawn as straight lines diagonally across the rectangle. As can be seen in \(K 7\), a single diagonal cycle may wind around the torus several times.
Suppose now that Gt is a Platonic graph on the torus, with parameters \((6,3)\)
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or (4,4). Consider a cycle \(C=(u 0, u 1, u 2, \ldots, u m)\) constructed by following a diagonal path. LEMMA 13.11 If \(C=(u 0, u 1, u 2, \ldots, u m)\) is a diagonal cycle in \(G t\), then \(C\) is an essential cycle. PROOF Suppose first that \(G\) has parameters (4,4), and suppose that \(C\) is contractible, with interior INT(C). For each \(v_{i} \in C\), there is one adjacent vertex in INT(C). This is illustrated in Figure 13.20. Since each face of \(G t\) has degree four, the interior adjacent vertices to \(u 0, u 1, U 2, \ldots, u m\) form another diagonal cycle \(C^{\prime}\) in INT(C). Interior to \(C^{\prime}\) is another diagonal" cycle, etc., leading to an infinite sequence of diagonal cycles, a contradiction. Therefore \(C\) must be an essential cycle. If \(G\) has parameters \((6,3)\), the argument is nearly identical, except that each \(v_{i} \in C\) has two adjacent vertices in INT(C). Since \(G t\) is a triangulation, we again find \(C^{\prime}\) in INT(C), and so forth.


\section*{FI GURE 13.20}

\section*{A diagonal cycle with \((k, \ell)=(4,4)\)}

The proof of Lemma 13.11 also shows how to draw \(G t\). Given a diagonal cycle CO, a sequence of "parallel" adjacent diagonal cycles is determined, \(C 0, C 1, C 2, \ldots\). For any \(v_{0} \in C_{0}\), an edge not on \(C 0\) can then be selected, and a diagonal cycle containing it can be constructed. We find that the edges of \(G\) can be partitioned into "orthogonal" diagonal cycles \(C_{0}^{\prime}, C_{1}^{\prime}, \ldots\). Each Ci winds page_351

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around the torus one or more times, intersecting each \(C_{j}^{\prime}\) in a regular pattern, as can be seen from Figures 13.17 and 13.18.

If \(C=(u 0, u 1, u 2, \ldots, u m)\) is any cycle constructed by following a diagonal path in a Platonic map, then the argument of Lemma 13.11 can be used to show that \(C\) must be a diagonal cycle. The only way in which \(C\) may fail to be a diagonal cycle is if one pair of edges, say umu0 and \(u 0 u 1\) are not diagonal edges. Suppose that \(G\) has parameters \((4,4)\). We then find that \(u 0\) has either 0 or 2 adjacent vertices to the right of \(C\). Since every face has degree four, the parallel cycle \(C^{\prime}\) is either shorter than \(C\) or longer than \(C\), by two edges. If it is longer than \(C\), then the parallel cycle \(C^{\prime \prime}\) is again longer than \(C^{\prime}\) by two edges, and so forth. As this leads to a contradiction, suppose that \(C^{\prime}\) is two edges shorter than \(C\). Then \(C^{\prime \prime}\) is again two edges shorter than \(C^{\prime}\), etc. Eventually we find a cycle of length four or five for which no parallel cycle can exist. If \(G\) has parameters \((6,3)\), the argument is similar.

\subsection*{13.4.2 Drawing torus maps, triangulations}

Read's algorithm for drawing a planar graph, given a rotation system, can be extended to torus maps. Let Gt be a 2-connected combinatorial embedding, with no vertices of degree two. Suppose that \(G\) has no loops, and that if there are multiple edges, no face is a digon. If Gt is not a triangulation, we can triangulate it by adding diagonal edges across non-triangular faces, so that no loops or digon faces are created. The smallest possible triangulation of the torus is shown in Figure 13.21. We denote it by T3. It can be constructed from the thetagraph shown in Figure 13.14 by adding one vertex \(w\), adjacent to each vertex \(u\) and \(u\) three times. Notice that \(T 3\) is a 6-regular graph, whose faces are all triangles. It is the unique triangulation of the torus on three vertices (Exercise 13.4.4).
There is a triangulation on four points, denoted \(T 4\), which can be constructed from the rectangular form of the torus. A 3-vertex graph consisting of two digon cycles \((u, u)\) and \((u, w)\) with a common vertex \(u\) is 2-cell embedded on the torus. There is a single face, of degree eight. A fourth vertex \(x\) is placed in this face, and joined to each vertex on the boundary of the face. It is illustrated in Figure 13.22. In this diagram, the sides a and \(b\) of the rectangle are also graph edges. Notice that \(T 4\) has two vertices ( \(u\) and \(x\) ) of degree eight, and
two ( \(u\) and \(w\) ) of degree four.
Suppose that \(G\) has \(n\) vertices, with \(n 3\) of degree three, \(n 4\) of degree four, etc. Then since \(G t\) is a triangulation, we have \(3 f=2 \varepsilon\). Euler's formula then gives \(\varepsilon=3 n\), and:

\section*{\(3 n 3+2 n 4+n 5=n 7+2 n 8+3 n 9+\ldots\)}

LEMMA 13.12 Either there is a vertex of degree three, four, or five, or else all vertices have degree six.

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\section*{FI GURE 13.21}

\section*{The triangulation \(T 3\)}

Now any triangulation in which all vertices have degree six is a Platonic map of type (6,3), and we know how to draw it as a tiling of the plane. Otherwise, there is a vertex of degree three, four, or five. We can use a modification of the algorithm REDUCEGRAPH() of Section 13.4.2 to reduce the triangulation \(G t\) on \(n\) vertices to a triangulation on \(n-1\) vertices, until either \(T 3\) or \(T 4\) results, or a 6 -regular triangulation results. We must ensure that the reduction does not create any loops or digon faces.
Suppose that vertex \(u\) of \(G\) has degree three, four, or five, and suppose that \(n \geq 4\). If \(\operatorname{DEG}(u)=3\), then \(G t-u\) is a triangulation of the torus on \(n-1\) vertices. If \(\operatorname{DEG}(u)=4\), suppose that \(u\) is adjacent to \(u, w, x, y\), in cyclic order. If at least three of these vertices are distinct, then at least one of the diagonals of the 4 -cycle ( \(u, w, x\), \(y\) ) has distinct endpoints. Suppose it is \(u x\). Then \(G t-u+u x\) will be a triangulation on \(n-1\) vertices, without loops or digon faces. Otherwise there are only two distinct vertices in the 4-cycle, which is then ( \(U, w, U, w\) ); that is, there are four parallel edges connecting \(u\) and \(w\). Three of these parallel edges form a theta-graph, whose only embedding has a single face, a hexagon, shown in Figure 13.23. The fourth parallel edge cuts the hexagon into two quadrilaterals, one of which contains \(u\).
The remaining vertices of \(G\) are located in the other quadrilateral. If \(n=4\), then the map can only be the triangulation T4, with \(u\) and \(w\) as the two vertices of degree eight. If \(n \geq 5\), there are at least two other vertices in the other quadrilateral. This quadrilateral and the vertices it contains determine a planar graph, which must have several vertices of degree three, four, or five. We choose one of these, and delete it instead of \(u\).
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\section*{FI GURE 13.22}

\section*{The triangulation T4}

If \(\operatorname{DEG}(u)=5\), let \(u\) be adjacent to \(u, w, x, y, z\), in cyclic order. If \(u, x\), and \(y\) are distinct, we proceed as in the planar case, deleting \(u\) and adding two diagonals across the pentagon. Otherwise, we can assume that \(u=x\),
since \(G\) has no loops.
If \(w, y\), and \(z\) are distinct, then we can add the diagonals \(w y\) and \(w z\) to get a triangulation. Otherwise we can assume that \(w=y\). But then \(z, w\), and \(x\) are distinct, so that we can add the diagonals \(z w\) and \(z x\) to obtain a triangulation with no loops or digon faces. There are always at least three distinct vertices on the boundary of the pentagon. This gives the following theorem:
THEOREM 13.13 Let Gt be a torus map on \(n \geq 4\) vertices which is not 6-regular, with no loops or digon faces. Then Gt can be reduced to one of the following:
1. The triangulation T3
2. The triangulation T4
3. A 6-regular triangulation

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FI GURE 13.23
Reducing a triangulation, \(\operatorname{DEG}(u)=4\)


FI GURE 13.24
Reducing a triangulation, \(\operatorname{DEG}(u)=5\)
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Algorithm 13.4.1: REDUCETORUSMAP(G,t)
if \(G=T 3\) or \(G=T 4\) or \(G\) is 6-regular return (null)
if there is a vertex \(u\) with \(\operatorname{DEG}(u)=3\)
then \(\left\{\begin{array}{l}\text { let } t(u)=(u v, u w, u x) \\ G^{\prime} \leftarrow G-u \\ \text { return }\left(G^{\prime}\right)\end{array}\right.\)
if there is a vertex \(u\) with \(\operatorname{DEG}(u)=4\)
then \(\left\{\begin{array}{l}\text { let } t(u)=(u v, u w, u x, u y) \\ \text { if } v=x \text { and } w=y \\ \text { then pick a new } u \text { from } V(G)-\{u, v, w\} \text { of degree } 4 \text { or } 5\end{array}\right.\) if \(\operatorname{DEG}(u)=4\)
```

    let \(t(u)=(u v, u w, u x, u y)\)
    if \(v \neq x\)
    then $\left\{\begin{array}{l}\text { then }\left\{\begin{array}{l}G^{\prime} \leftarrow G^{\prime}-u+v x \\ v x \text { replaces } v u \text { in } t(v) \text { and } x v \text { replaces } x u \text { in } t(x)\end{array}\right. \\ \text { else }\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+w y \\ w y \text { replaces } w u \text { in } t(w) \text { and } y w \text { replaces } y u \text { in } t(y)\end{array}\right. \\ \text { return }\left(G^{\prime}\right)\end{array}\right.$

```
    pick \(u\) of degree 5, let \(t(u)=(u u, u w, u x, u y, u z)\)
                            if \(u, x\) and \(y\) are distinct
    then \(\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+v x+v y \\ v x, v y \text { replace } v u \text { in } t(v) \\ x v \text { replaces } x u \text { in } t(x) \text { and } y v \text { replaces } y u \text { in } t(y)\end{array}\right.\)
else if \(w, y\) and \(z\) are distinct
    then \(\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+w y+w z \\ w y, w z \text { replace } w u \text { in } t(w) \\ y z \text { replaces } y u \text { in } t(y) \text { and } z w \text { replaces } z u \text { in } t(z)\end{array}\right.\)
else if \(x, z\) and \(u\) are distinct
    then \(\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+x z+x v \\ x z, x v \text { replace } x u \text { in } t(x) \\ z v \text { replaces } z u \text { in } t(z) \text { and }\end{array}\right.\)
else if \(y, u\) and \(w\) are distinct
then \(\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+y v+y w \\ y v, y w \text { replace } y u \text { in } t(y) \\ v w \text { replaces } v u \text { in } t(v) \text { and }\end{array}\right.\)
else if \(z, w\) and \(x\) are distinct
then \(\left\{\begin{array}{l}G^{\prime} \leftarrow G-u+z w+z x \\ z w, z x \text { replace } z u \text { in } t(z) \\ w x \text { replaces } w u \text { in } t(w) \text { and } x z \text { replaces } x u \text { in } t(x)\end{array}\right.\) return ( \(G^{\prime}\) )

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Algorithm 13.4.1 is input a triangulation of the torus, \(G\), on \(n \geq 4\) vertices with rotation system \(t\), with no loops or digon faces. It constructs a triangulation \(G^{\prime}\) on \(n-1\) vertices, whenever possible. It can be used to successively reduce \(G t\) to one of \(T 3, T 4\), or a 6 -regular triangulation. Drawings of \(T 3\) and \(T 4\) on the torus are shown in Figures 13.21 and 13.22. We can use these coordinatizations in a rectangle as topological embeddings. If a 6-regular triangulation is obtained, we can use diagonal cycles to obtain a coordinatization of it. These embeddings have no loops, and no digon faces. Every digon is embedded as an essential cycle. We then replace the deleted vertices in reverse order, exactly as in the planar case of READSALGORITHM(), using the visible region to assign coordinates to the deleted vertex. The result is a straight-line drawing of Gt on the torus; that is, a topological embedding \(G \psi\).
We summarize this as follows:
THEOREM 13.14 Every torus map has a straight-line embedding in the rectangle and hexagon models of the torus.

\section*{Exercises}
13.4.1 Find the duals of the embeddings of the triangulations \(T 3\) and \(T 4\) shown in Figures 13.21 and 13.22. What graphs are they?
13.4.2 Find the two tilings of the plane determined by a Platonic map of \(K 4,4\).
13.4.3 Find the tiling of the plane determined by a Platonic map of C3×C3.
13.4.4 Show that \(T 3\) is the unique triangulation of the torus on three vertices.
13.4.5 Show that there are two distinct triangulations of the torus on four vertices.
13.4.6 Show that there are five distinct embeddings of the cube on the torus.

\subsection*{13.5 Graphs on the projective plane}

The projective plane is most conveniently represented as a disc with antipodal points identified. This is equivalent to the digon form \(c+c+\) of the projective plane shown in Figures 13.7 and 13.9. An embedding of \(K 6\) and its dual are shown in Figure 13.25. It is easy to verify that the dual of \(K 6\) on the projective plane is the Petersen graph. As before, we shall only be concerned with 2-cell embeddings of 2-connected graphs. Now it can be a little tricky to visualize the faces of an embedding on the projective plane, because the projective plane is non-orientable. Each point on the circumference of the disc is identified with its antipodally opposite point. When
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an edge of the graph meets the disc boundary, it continues from the antipodal point. But the region immediately to the right of the edge as it meets the boundary is identified with the region immediately to the left of the antipodal point. A consequence is that rotation systems must be defined somewhat differently for non-orientable surfaces, and the algorithm FACIALCYCLE() which constructs the faces of an embedding must be modified.


\section*{FI GURE 13.25}

\section*{K6and its dual, the Petersen graph, on the projective plane}

Let \(G \psi\) be an embedding of a 2-connected graph \(G\) in the projective plane, and let \(u\) be a vertex of \(G\). If we walk around \(u \psi\) in a small clockwise circle, we encounter the incident edges in a certain cyclic order, say uul, \(u U 2, \ldots\), uuk. If we walk along the edges of \(G \psi\), always staying within the disc, then the embedding appears exactly like a planar embedding. If we traverse an edge that crosses the boundary of the disc, and continue on until we reach \(u \psi\) again, we find that the cyclic order of the incident edges at \(u \psi\) has been reversed.
Consequently a rotation system must be defined differently for a non-orientable surface. The projective plane is represented as a disc. We choose an orientation for this disc. Then given any vertex \(u\), we walk around \(u \psi\) in a small clockwise circle, and construct a cyclic list of incident edges. We assign a signature to each edge \(u u\), denoted \(\operatorname{SGN}(u v)\). If an edge (uui) \(\psi\) crosses the boundary of the disc, then \(\operatorname{SGN}(u u i)=-1\). Otherwise it is +1 . The signature does not depend on the direction in which the edge is traversed. In this way, the embedding \(\psi\) determines a signed rotation system.
We will use \(n\) to denote a rotation system for an embedding on the projective plane. For each vertex \(u, n(u)\) denotes a cyclic list of signed incident edges. Two embeddings of \(К 3,3\) are shown in Figure 13.26. The rotation systems corresponding to them are shown in Figure 13.27. Although the rotation systems are


FI GURE 13.26
Two equivalent embeddings of \(K 3\), 3on the projective plane
\(\pi(1)=(12,16,-14)\)
\(\pi(2)=(21,-25,23)\)
\(\pi(3)=(32,-36,34)\)
\(\pi(4)=(43,-41,45)\)
\(\pi(5)=(54,-52,56)\)
\(\pi(6)=(61,65,-63)\)
\[
\begin{aligned}
& \pi(1)=(14,16,-12) \\
& \pi(2)=(23,-21,25) \\
& \pi(3)=(34,32,36) \\
& \pi(4)=(41,-45,43) \\
& \pi(5)=(52,-54,56) \\
& \pi(6)=(61,63,65)
\end{aligned}
\]

FI GURE 13.27

\section*{Two rotation systems for \(K 3,3\) on the projective plane}

Given a topological embedding \(G \psi\), this method of defining a rotation system for \(G\) is not unique, as it depends on the disc chosen to represent the projective plane. With an orientable surface, this situation does not arise. We must show that a signed rotation system uniquely determines the faces of an embedding, and that all rotation systems corresponding to \(\psi\) determine the same faces. In the embedding on the right of Figure 13.27, we can cut the projective plane along the non-contractible J ordan curve indicated by the dotted line, call it \(C\). We then flip one-half of the disc over, and glue the two pieces along the common boundary.

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We obtain another disc representing the projective plane, with antipodal points identified. This gives another rotation system \(n^{\prime}\) for \(G\). In Figure 13.27, the edges crossed by \(C\) will now have a signature of -1 ; edges which previously had a signature of -1 will now have +1 . It is easy to see that with respect to the new disc, the embedding will have a face which is a 6 -cycle ( \(1,2,3,4,5,6\) ) with "spokes" to the boundary of the disc, exactly like the embedding on the left. Thus the embeddings are equivalent.
If \(n\) is a signed rotation system determined by an embedding \(G \psi\) and a disc representation of the projective plane, we call \(G \sqcap\) a combinatorial embedding. As we shall see, the faces of \(G \psi\) are completely determined by Gп.
DEFINITION 13.15: A projective map is a combinatorial embedding \(G\) n of a 2-connected graph \(G\) on the projective plane, where \(\pi\) is a signed rotation system corresponding to a disc representation of the projective plane.
In order to show that all signed rotation systems arising from \(G \psi\) have the same facial cycles, we begin by rewriting the algorithm FAClALCYCLE() for a non-orientable surface. Notice that when traversing the facial cycles of a graph embedded on a disc representing the projective plane, the clockwise cyclic order of the incident edges viewed from above the disc appears counterclockwise when viewed from below, and vice versa. The algorithm uses a boolean variable onTop to indicate whether it is currently viewing the disc from above. Initially onTop has the value true. Each time an edge uu with \(\operatorname{SGN}(u u)=-1\) is encountered, it reverses the value of onTop. Any vertices visited while onTop is false will see a counterclockwise orientation for their incident edges. Those with onTop true will see a clockwise orientation.
When a graph is embedded on an orientable surface, the facial cycles are oriented cycles. We can assign a clockwise orientation to one cycle and the orientation of all adjacent cycles is determined, and so forth, so that an orientation can be assigned to the entire embedding. Reversing the cycles gives an equivalent, but reversed, embedding.
Graphs embedded on the projective plane do not have this property. If we try to assign an orientation to the facial cycles of an embedding \(G \psi\), we can then choose different signed rotation systems corresponding to \(G \psi\),
and different orientations of the cycles will be obtained. However, if we are given the (unoriented) facial cycles, they can be glued together uniquely along their common edges to construct a polygonal representation of the projective plane. Therefore if we are given the facial cycles, they will determine a topological embedding \(G \psi\).
Algorithm 13.5.1 when given a signed rotation system \(п\) of an embedding \(G \psi\) on a non-orientable surface and a vertex \(u\) with an incident edge \(e\), will find the facial cycle containing \(e\).

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Algorithm 13.5.1: \(\operatorname{FAClALCYCLESGN(Gп,~u,~e)~}\) onTop \(\leftarrow\) true "initially view the disc from above"

\section*{\(e^{\prime} \leftarrow e\) \\ repeat}
comment: \(e^{\prime}\) currently equals \(u v\), for some \(v\)
if \(\operatorname{SGN}\left(e^{\prime}\right)=-1\) then onTop \(\leftarrow\) not onTop
\(v \leftarrow\) other end of \(e^{\prime}\)
\(e^{\prime \prime} \leftarrow\) edge of \(\pi(v)\) corresponding to \(e^{\prime}\)
comment: \(e^{\prime \prime}\) currently equals \(v u\)
if onTop
then \(e^{\prime} \leftarrow\) edge preceding \(e^{\prime \prime}\) in \(\pi(v)\)
else \(e^{\prime} \leftarrow\) edge following \(e^{\prime \prime}\) in \(\pi(v)\)
\(u \leftarrow v\)
until \(e^{\prime}=e\) and onTop
If this algorithm is applied to the combinatorial embeddings of \(K 3,3\) given in Figure 13.27, identical faces will be constructed. Since the projective plane is constructed by gluing together the faces, it follows that the topological embeddings they represent are equivalent.
In Definition 13.14 combinatorial embeddings \(G^{t_{1}}\) and \(G^{t_{1}}\) on the torus were defined to be equivalent if there exists an autmorphism of \(G\) that induces a mapping of \(t 1\) to \(t 2\) or \(\overline{t_{2}}\). This is inappropriate for signed rotation systems, as it cannot take the signatures into consideration. Therefore we define equivalence of signed rotation systems in terms of facial cycles.
DEFINITION 13.16: Projective maps \(G^{\pi_{1}}\) and \(G^{\pi_{2}}\) are equivalent if they have the same facial cycles. In general, let \(G \psi\) be a topological embedding, and consider two different representations of the projective plane as digons, \(a+a+\) and \(b+b+\). Let \(n a\) be the signed rotation system corresponding to \(a+a+a n d\) let \(n b\) correspond to \(b+b+\). The boundary of the disc given by \(b+b+\) is a non-contractible Jordan curve \(C\) in the \(a+a+\) representation. It will intersect the graph \(G \psi\) in one or more points. Refer to Figure 13.28. When the disc is cut along the curve \(C\), it may cut some edges of \(G \psi\) more than once. If so, we subdivide those edges which are cut more than once. Hence we can assume that every edge of \(G \psi\) is cut at most once by C. If the curve cuts through a vertex, we can move it slightly so that it misses the vertex. We can do this because the graph has a finite number of vertices.
Suppose first that \(C\) cuts the boundary of the \(a+a+\) disc in exactly two points (which are antipodal points), as in Figure 13.28. To transform the \(a+a+\) representation into the \(b+b+\) representation, we cut the disc along the dotted curve,

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\section*{FIGURE 13.28}

\section*{Representations \(a+a+\) and \(b+b+\) of the projective plane}
then flip one-half of the disc over, and glue the two equivalent parts of the a+a+ boundary together. We can flip either half over. Denote the two parts of the disc obtained as \(D 1\) and \(D 2\), where \(D 2\) is the part that is flipped over. The result is a disc whose boundary is \(b+b+\), shown as the disc on the right in Figure 13.28. We obtain \(п b\) from па by reversing the cyclic adjacencies for all vertices in \(D 2\), and by assigning -1 to those edges that are cut by \(C\). We now compare the facial cycles of \(G^{\pi_{a}}\) and \(G^{\pi_{b}}\) constructed by the algorithm FACIALCYCLESGN(). Let ( \(u 1, u 2, \ldots, u k\) ) be a facial cycle constructed for \(G^{\pi_{a}}\). Without loss of generality, suppose that FACIALCYCLESGN \(\left(G^{\pi_{b}}\right)\) begins to trace out this face from \(u 1\) which is in D1. If the entire face is within \(D 1\), the result will be the same as the face obtained using па. If the facial cycle crosses from \(D 1\) to \(D 2\) via the \(a+a+\) boundary, then since \(D 2\) was flipped upside down, and the cyclic adjacencies of na were reversed to obtain \(n b\), the same facial boundary will be constructed using \(n b\) or \(n\) a. If the facial cycle crosses from \(D 1\) to \(D 2\) via \(C\), then since \(\pi b\) attaches a signature of -1 to these edges, the cyclic adjacencies will be reversed by FACI ALCYCLESGN(). But the cyclic adjacencies of пa in \(D 2\) were also reversed in \(п b\). The net effect is that the same facial boundary is constructed using \(n b\) or \(п a\) at each step of the algorithm. It follows that the two embeddings \(G^{\pi_{a}}\) and \(G^{\pi_{b}}\) have the same 2 -cells as faces. Now we may have subdivided some edges of \(G\) before cutting and pasting the disc. Vertices of degree two do not affect the faces, and a cyclic order of two edges is invariant when reversed. Therefore, when traversing a facial cycle along a path created by subdividing an edge, the important factor is the number of -1 's encountered. Hence we can contract any subdivided edges and assign a signature of -1 if the number of edges in the subdivided path was odd. The result will be the same facial cycle. We conclude that the faces of \(G^{\pi_{a}}\) and \(G^{\pi_{b}}\) are identical, so that the embeddings are equivalent.
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Suppose now that \(C\) cuts the boundary of the \(a+a+\) disc in more than one pair of antipodal points, where \(C\) is the boundary of the \(b+b+\) disc. There are an infinite (in fact, uncountable) number of possible non-
contractible Jordan curves \(C\). But there are only a finite number of possible signed rotation systems for \(G \psi\), since \(G\) is finite. Therefore we will only consider non-contractible J ordan curves which meet the boundary \(a+a+\) in a finite number of points. A more complete treatment can be found in the book of MOHAR and THOMASSEN [88].
If \(C\) cuts the boundary of the \(a+a+\) disc in more than one pair of antipodal points, we proceed by induction. We are given a graph \(G\) embedded on the disc, with a rotation system пa. As before we assume that \(C\) does not cut any vertices of \(G \psi\). Choose two consecutive points \(P\) and \(Q\) on \(C\) at which the disc boundary is cut such that one of the intervals \([P, Q]\) and \([Q, P]\) on the boundary is not cut by any other points of \(C\). This is illustrated in Figure 13.29, where \(C\) is the dotted curve. Let the antipodal points of \(P\) and \(Q\) be \(P^{\prime}\) and \(Q^{\prime}\). Let \(C[P, Q]\) denote the portion of \(C\) from \(P\) to \(Q\). Make a cut in the disc very close to \(C[P, Q]\), cutting off a portion \(D 1\) of the disc, so that \(C[P, Q]\) is completely contained within \(D 1\), but so that the only part of \(G\) that is affected by the cut are the edges of \(G\) that cross \(C[P, Q]\). This is possible because the graph is finite. Let the remainder of the disc be \(D 2\). We now flip \(D 1\) over, and glue the matching boundaries of \(D 1\) and \(D 2\) near \(P^{\prime}\) and \(Q^{\prime}\). The result is a disc representation of the projective plane such that \(C\) cuts its boundary in four fewer points. Let the boundary of the new disc be \(c+c+\), and let the signed rotation system corresponding to it be пс.


FI GURE 13.29
Transforming a disc representation of the projective plane
Consider the faces of \(G^{\pi_{a}}\) and \(G^{\pi_{c}}\). The rotation systems \(\pi a\) and \(\pi c\) differ only in edges which are within \(D 1\), or which cross from \(D 1\) to \(D 2\). Vertices within \(D 1\) have adjacency lists of opposite orientation in па and \(п c\). FACIALCYCLESGN() will construct the same faces for both \(п \boldsymbol{a}\) and \(п с\). With respect to the \(п с\) disc,

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C has fewer intersections with the boundary. We use induction to conclude that FACIALCYCLESGN() will construct the same faces for both \(п c\) and \(n b\). It follows that \(G^{\pi_{a}}\) and \(G^{\pi_{b}}\) always have the same faces. Because the projective plane is constructed by gluing the faces together, this gives:
THEOREM 13.15 Let \(G \psi\) be an embedding on the projective plane. Given any two signed rotation systems па and \(n b\) for \(G \psi\), corresponding to different disc representations of the projective plane, \(G^{\pi_{a}}\) and \(G^{\pi_{b}}\) are equivalent embeddings.
THEOREM 13.16 Let \(G^{\psi_{1}}\) and \(G^{\psi_{2}}\) be topological embeddings of \(G\) on the projective plane with corresponding combinatorial embeddings \(G^{\pi_{a}}\) and \(G^{\pi_{b}}\), with respect to two disc representations of the projective plane. Then \(G^{\psi_{1}}\) and \(G^{\psi_{2}}\) are homeomorphic if and only \(G^{\pi_{a}}\) and \(G^{\pi_{b}}\) are equivalent.
PROOF If \(G^{\pi_{a}}\) and \(G^{\pi_{b}}\) are equivalent, they have the same facial cycles. The faces are homeomorphic to 2cells bounded by the facial cycles. It follows that the faces of \(G^{\pi_{a}}\) and \(G^{\pi_{b}}\) determine a homeomorphism of the embeddings \(G^{\psi_{1}}\) and \(G^{\psi_{2}}\). Therefore \(G^{\psi_{1}}\) and \(G^{\psi_{2}}\) are homeomorphic if and only if \(G^{\pi_{a}}\) and \(G^{\pi_{b}}\) are equivalent.
If \(G p\) is a planar embedding of a 2 -connected graph, it is very easy to convert \(p\) to a projective rotation system \(п\). When \(G\) is drawn in the plane, one face is always the outer face. We draw \(G\) in a disc representing the projective plane. We then choose any edge e on the outer face, and reroute it so that it crosses the boundary of the disc. The result is a projective map. The two faces on either side of e in the planar map become one face in the projective map. The cyclic order of adjacent vertices is unchanged, for all vertices of \(G\). Thus, \(p\) can be converted to a projective rotation system, by assigning a signature of -1 to any one edge of \(G\). However, the embeddings constructed in this way are somewhat unsatisfactory, as there is a noncontractible J ordan curve in the surface which cuts the embedding in only one point.

\subsection*{13.5.1 The facewidth}

The projective plane has unorientable genus one. The torus has orientable genus one. Although they both have genus one, these surfaces behave very differently. \(K 7\) can be embedded on the torus. Yet it is easy to see that it cannot be embedded on the projective plane, as the unique embedding of \(K 6\) shown in Figure 13.25 cannot be extended to K7. Alternatively, Euler's formula can be used to show that \(K 7\) has too many edges to embed on the projective plane. However, there are infinite families of graphs that can be embedded on the projective plane, but not on the torus.
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We begin with two families of graphs called the Möbius ladder and Möbius lattice, which can be embedded on both the projective plane and torus.
DEFINITION 13.17: The Möbius ladder L2n is the graph with \(2 n\) vertices \(\{u 1, u 2, \ldots, u 2 n\}\) such that \(v_{i} \longrightarrow v_{i+1}\), and \(v_{i} \longrightarrow v_{i+n}\), where subscripts larger than \(2 n\) are reduced modulo \(2 n\).
The Möbius ladder L6 is just K3, 3, shown in Figure 13.26. L8 is shown in Figure 13.30. Notice that \(L 2 n\) is always a 3-regular graph.
DEFINITION 13.18: The Möbius lattice \(L 2 n-1\) is the graph with \(2 n-1\) vertices \(\{u 1, u 2, \ldots, u 2 n-1\}\) such that \(v_{i} \longrightarrow v_{i+1}, v_{i} \longrightarrow v_{i+n-1}\), and \(v_{i} \longrightarrow v_{i+n}\) where subscripts larger than \(2 n-1\) are reduced modulo \(2 n-1\). The Möbius lattice \(L 5\) is just \(K 5\). \(L 7\) is shown in Fiqure 13.30 . Notice that \(L 2 n-1\) is always a 4-regular graph.


FI GURE 13.30
The Möbius ladder L8and Möbius lattice L7
There is a clever trick that can be used to convert these projective embeddings of \(L 2 n\) and \(L 2 n-1\) to toroidal embeddings. Draw an essential cycle C across the disc as shown by the dotted line in Figure 13.30, dividing it
into two parts, \(D 1\) and \(D 2\). Notice that \(C\) intersects the graph embedding in only two points, representing the vertices 1 and 5 . Vertices 1 and 5 are both joined to several vertices located in \(D 1\) and \(D 2\). Now cut the disc along \(C\), flip \(D 2\) over, and glue \(D 1\) and \(D 2\) along the common boundary to get a new disc representation of the projective plane, as shown in Figure 13.31. Vertices 1 and 5 are on the boundary of the new disc. These antipodal points are the only points of the embedding \(G \psi\) on the disc boundary. Therefore we can convert the disc into the hexagonal form of the torus, obtaining an embedding on the torus.

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\section*{\(L_{7}\)}


\section*{FI GURE 13.31}

\section*{Converting a projective embedding to a toroidal embedding}

DEFINITION 13.19: Let \(G \psi\) be a graph embedding in a surface \(\Sigma\). Let \(C\) be a non-contractible J ordan curve in \(\Sigma\). The facewidth of \(C\) is \(f w(C)\), the number of points of \(G \psi\) common to \(C\). The facewidth of \(G \psi\) is fw \((G \psi)\), the minimum fw \((C)\), where \(C\) is any non-contractible Jordan curve in \(\Sigma\).
The facewidth is sometimes known as the representativity of an embedding. Let \(C\) be a non-contractible J ordan curve of minimum possible facewidth, for an embedding \(G \psi\). If \(C\) intersects a face \(F\), then it also intersects the boundary of \(F\). If an intersection point is not at a vertex, then it is at an interior point of an edge \(e\). Then \(C\) also intersects the face on the other side of \(e\). In such a case, we can alter \(C\) slightly so that it passes through e at an endpoint of the edge. The result is another non-contractible Jordan curve, also of minimum facewidth. This gives the following:
LEMMA 13.17 Given any embedding \(G \psi\), there is a non-contractible Jordan curve \(C\) of facewidth \(f w(G \psi)\) such that C intersects \(G \psi\) only at images of vertices.
Now the faces of an embedding \(G \psi\) are determined completely by its rotation system. If C intersects \(G \psi\) only at images of vertices, then \(C\) determines a cyclic sequence of vertices ( \(u 1, u 2, \ldots, u k\) ), where \(k=f w(C)\), such that consecutive vertices are on the same facial boundary. It follows that fw \((G \psi)\) depends only on the rotation system, so that we can also write fw(Gn). We show that the method used above to convert a projective embedding of \(L 7\) to a toroidal embedding works in general, whenever fw(Gn) is at most three.
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THEOREM 13.18 Let Gn be a projective embedding with facewidth at most three. Then \(G\) can be embedded on the torus.
PROOF The proof proceeds as in the previous example. Let \(C\) be a non-contractible J ordan curve with \(\mathrm{fw}(C) \leq 3\). Use the hexagonal form of the torus, \(a+b+c+a-b-c-\), as in Figure 13.32. We cut the disc of the projective plane along \(C\) obtaining \(D 1\) and \(D 2\), which we glue together to obtain a new disc. Without loss of generality, assume that \(\mathrm{fw}(C)=3\), so that there are three pairs of antipodal points of \(G n\) on the boundary of the new disc, call them \(p, q\), and \(r\), and suppose that they occur in this order along the boundary of the disc. We convert the disc into the hexagonal form of the torus, as in Figure 13.32. We place \(p\) on the side of the hexagon labeled \(a, q\) on the side labeled \(b\), and \(r\) on the side labeled \(c\). In the hexagon, the identified pairs of sides are not antipodally reversed as in the disc model of the projective plane. However, there is only one point \(p, q\), or \(r\) on each side, so that the sequence of points on the boundary is the same. The result is a toroidal embedding of \(G\).


FI GURE 13.32
Converting a projective embedding to a toroidal embedding
This transformation will only work when \(f w(C) \leq 3\). It has been shown by FIEDLER, HUNEKE, RICHTER, and ROBERTSON [41] that the converse of this theorem is also true; so that if a toroidal graph can be embedded on the projective plane, then the facewidth on the projective plane is at most three. We can use this theorem to construct projective graphs which are not toroidal. The construction uses Möbius ladders or lattices. Suppose that we start with a Möbius ladder \(L 2 n\) with vertices \(\{u 1, u 2, \ldots, u 2 n\}\). Add \(2 n\) more vertices \(u 1, u 2\), \(\ldots, u 2 n\) forming a cycle in which \(u_{i} \longrightarrow u_{i+1}\), and add the edges \(u_{i} \longrightarrow v_{i}\), for all \(i\). The result is a graph as shown in Figure 13.33. The facewidth of this graph is four, if \(n \geq 4\),
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so that it cannot be embedded in the torus. The facewidth can be made arbitrarily high by increasing \(n\) and adding more cycles in this fashion.


FI GURE 13.33

\section*{An embedding with facewidth four}

\subsection*{13.5.2 Double covers}

The projective plane can be viewed as exactly one-half of a sphere, by considering the disc representing the projective plane as the upper hemisphere of a sphere. If we choose a great circle as the equator of a sphere, and identify antipodal points on the sphere, the result is a two-to-one mapping of the sphere onto the projective plane. Open discs on the sphere are mapped to open discs on the projective plane. The equator becomes the boundary of the disc representing the projective plane. Thus, we say that the sphere is a double cover of the projective plane.
It is also possible for one graph to be a double cover of another graph.
DEFINITION 13.20: Let \(G\) and \(H\) be simple graphs such that there is a two-to-one mapping \(\theta: V(H) \rightarrow V(G)\) with the property that \(\theta\) induces a two-to-one mapping of edges, \(\theta: E(H) \rightarrow E(G)\). Then \(H\) is said to be a double cover of \(G\).
Since the sphere is a double cover of the projective plane, given any graph \(G\) embedded on the projective plane, there is a corresponding planar graph \(H\) embedded on the sphere, such that \(H\) is a double cover of \(G\). We begin with an embedding \(G n\) with respect to a disc representation of the projective plane. We make a copy of the disc, and transform it by reflecting each of its points in the center of the disc (an antipodal reflection). We then glue the two discs together

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\section*{FI GURE 13.34}

\section*{The cube is a double cover of \(K 4\)}
along their common boundary. The common boundary of the discs becomes the equator of the sphere. The result is an embedding on the sphere of a double cover \(H\) of \(G\). The edges of \(H\) which cross the equator are those with signature -1 in \(n\). We denote this double cover by \(D C(G n)\). An example is illustrated in Figure 13.34, showing that the graph of the cube is a double cover of \(K 4\). We also find that the dodecahedron is a double cover of the Petersen graph, as can be seen from Figure 13.25.
The medial digraph of an embedding as expressed in Definition 12.19 is inappropriate for unorientable surfaces. Although it encapsulates the cyclic adjacencies of each vertex, it does not take into account the signature of the edges. In order to distinguish inequivalent projective embeddings of a graph, and to determine the symmetries (automorphisms) of an embedding, we can use the double cover. THEOREM 13.19 Let \(G^{\pi_{1}}\) and \(G^{\pi_{2}}\) be projective embeddings with corresponding spherical embeddings \(D C\left(G^{\pi_{1}}\right)^{p_{1}}\) and \(D C\left(G^{\pi_{2}}\right)^{p_{2}}\). Then \(G^{\pi_{1}}\) and \(G^{\pi_{2}}\) are equivalent if and only if \(D C\left(G^{\pi_{1}}\right)^{p_{1}}\) and \(D C\left(G^{\pi_{2}}\right)^{p_{2}}\) are equivalent.
PROOF Consider a projective embedding \(G n\). Let \(H=D C(G \square)\), and let \(H p\) be the spherical embedding determined by \(G \sqcap\). Let \(V(G)=\{u 1, u 2, \ldots, u n\}, V(H)=\{u 1, u 2, \ldots, u n, w 1, w 2, \ldots, w n\}\), and let \(\theta: V(H) \rightarrow V(G)\) be the natural double cover mapping; namely, \(\theta(u i)=\theta(w i)=u i\). The mapping \(\theta\) maps the cyclic adjacency lists of \(u i\) and wi to the cyclic adjacency list of ui, but the orientations are opposite to each other. Notice that \(H\) has an antipodal automorphism; namely, the unique permutation of \(V(H)\) which interchanges ui
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and wi is an automorphism of \(H\). Corresponding to each face of \(G n\), there are two faces of \(H p\), and these faces have opposite orientation. Each face of \(H p\) corresponds uniquely to a face of \(G \pi\). We see that \(H\) is an unorientable planar graph.
Suppose that \(G^{\pi_{1}}\) and \(G^{\pi_{2}}\) are equivalent embeddings. As the topological embeddings they represent are homeomorphic, they correspond to different disc representations of a projective embedding of \(G\). Let \(H=D C\left(G^{\pi_{1}}\right)\), with corresponding spherical embedding \(H p\). The boundary of the disc corresponding to \(G^{\pi_{2}}\) is a non-contractible J ordan curve \(C\) in the projective plane. There is a unique antipodally symmetric Jordan curve \(C^{\prime}\) on the sphere corresponding to \(C\). As in Figures 13.28 and 13.29, we cut the disc of \(G^{\pi_{1}}\) along \(C\) to get \(D 1\) and \(D 2\), flip \(D 2\) over, and glue the two parts along the common boundary to obtain the disc of \(G^{\pi_{2}}\). The corresponding transformations on the sphere are as follows. The equator of the sphere corresponds to the boundary of the \(n 1\)-disc. The curve \(C^{\prime}\) corresponds to the boundary of the \(n 2\)-disc. The equator and \(C^{\prime}\) intersect in a pair of antipodal points, dividing the surface of the sphere into four crescents, which correspond alternately to \(D 1\) and \(D 2\). The transformation of the \(n 1\)-disc into the \(n 2\)-disc is equivalent to taking \(C^{\prime}\) as the new equator, leaving \(H\) unchanged. It follows that \(D C\left(G^{\pi_{2}}\right)^{p_{2}}\) and \(D C\left(G^{\pi_{1}}\right)^{p_{1}}\) are equivalent. Conversely, suppose that \(D C\left(G^{\pi_{1}}\right)^{p_{1}}\) and \(D C\left(G^{\pi_{2}}\right)^{p_{2}}\) are equivalent. Since a double cover is an unorientable planar graph, the embeddings of \(D C\left(G^{\pi_{1}}\right)\) and \(D C\left(G^{\pi_{2}}\right)\) on the sphere can only differ in their equators. It follows that \(G^{\pi_{1}}\) and \(G^{\pi_{2}}\) are equivalent disc representations of homeomorphic projective embeddings.
Notice that if \(D C\left(G^{\pi_{1}}\right)\) is a 3-connected graph, by Whitney's theorem, it has a unique embedding on the sphere. Consequently \(G^{\pi_{1}}\) and \(G^{\pi_{2}}\) will be equivalent if and only if \(D C\left(G^{\pi_{1}}\right) \cong D C\left(G^{\pi_{2}}\right)\) in this case.

\section*{Exercises}
13.5.1 Find all embeddings of \(K 4\) and \(K 5\) on the projective plane.
13.5.2 Find a projective embedding of \(K 3,4\) and find its projective dual. What graph is it?
13.5.3 Let \(G \psi\) be a projective embedding of \(G\), and let \(n\) be an associated rotation system. Let \(C\) be any cycle of \(G\). Show that \(C\) is an essential cycle of \(G \psi\) if and only if the number of edges of \(C\) with signature -1 is congruent to \(2(\bmod 4)\).
13.5.4 Find the dual maps of the embeddings shown in Figure 13.30.
13.5.5 Show that the Möbius ladder L2n contains a topological subgraph \(T L 2 n-2\), when \(n \geq 3\).
13.5.6 Show that the Möbius lattice \(L 2 n-1\) is a minor of \(L 2 n+1\), if \(n \geq 3\).
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13.5.7 Show that the Möbius ladder \(L 2 n\) has an embedding on the torus in which all faces are hexagons. Construct the embedding for \(n=4,5\), and find the dual map. (Hint: In the rectangular representation of the torus, draw a cycle of length \(2 n\) which wraps around the torus twice. Show how to complete this to a hexagonal embedding of \(\mathrm{L2}\) n.)
13.5.8 Show that the Möbius lattice \(L 2 n-1\) has an embedding on the torus in which all faces are quadrilaterals. Construct the embedding for \(n=4,5\), and find the dual map. (Hint: In the rectangular representation of the torus, draw a cycle of length \(2 n-1\) which wraps around the torus twice. Show how to complete this to a quadrilateral embedding of \(L 2 n-1\).)
13.5.9 Show that there is a unique triangulation of the projective plane with three vertices and six edges.
13.5.10 Show that Read's algorithm for drawing a planar graph can be adapted to the projective plane. Show that there is a unique triangulation of the projective plane on three vertices, and that any triangulation can be reduced to it by deleting vertices of degrees three, four, or five, and adding diagonals to the faces obtained. Conclude that every projective graph has a straight-line drawing in the disc model of the projective plane.
13.5.11 Show that the graph of the \(2 n\)-prism is a double cover of the Möbius ladder L2n.
13.5.12 The cube is a double cover of \(K 4\). Find another double cover of \(K 4\).
13.5.13 Find a double cover of K6, as illustrated in Figure 13.25.
13.5.14 Find a projective embedding of the graph of the cube, and find its double cover.
13.5.15 The Desargues graph is shown in Figure 13.35. The Desargues graph is non-planar, non-projective, and non-toroidal. Show that it is a double cover of the Petersen graph.


FI GURE 13.35
The Desargues graph

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\subsection*{13.6 Embedding algorithms}

In this section, we outline an algorithm to determine whether a 2-connected graph \(G\) can be embedded on the projective plane, and to find an embedding Gn. It is modeled on algorithms of Gagarin, Mohar, and Myrvold and Roth. If \(G\) is planar, we know how to convert a planar embedding to a projective embedding. Hence we can assume that \(G\) is non-planar, so that it contains a Kuratowski subgraph.


\section*{FI GURE 13.36}

\section*{The embeddings of \(K 5\) on the projective plane}

There is exactly one embedding of \(K 3,3\) on the projective plane, shown in Figure 13.26, and two embeddings of \(K 5\), shown in Figure 13.36. These embeddings all have the property that there are no repeated vertices on any facial cycle. In Figure 13.26, the hamilton cycle ( \(1,2,5,6,3,4\) ) is an essential cycle. Since \(K 3,3\) has a unique embedding on the projective plane, this gives:
LEMMA 13.20 In any embedding of \(K 3,3\) in the projective plane, some hamilton cycle is an essential cycle. If we now cut the projective plane along this cycle, the result is a disc in which each vertex \(1,2, \ldots, 6\) appears twice on the boundary. The resulting diagram, shown in Figure 13.37, is a very convenient representation of the projective plane.
Consider a subgraph \(T K 3,3\) in \(G\). We want to determine whether \(G\) can be embedded in the projective plane. We will begin by embedding the subgraph \(T K 3,3\). There are six hamilton cycles of \(K 3,3\). Each corresponds to a cycle of \(T K 3\), 3. One of them must be essential. Exercise 13.6.2 describes an easy way to enumerate the hamilton cycles of \(K 3,3\). We take each of the six cycles in turn, and construct an embedding of \(T K 3,3\), as in Figure 13.37, and try to extend it

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FI GURE 13.37
A representation of the projective plane
to an embedding of \(G\). If any one succeeds, then \(G\) is projective. Otherwise we conclude that \(G\) is nonprojective.
The embedding of \(K 3,3\) divides the projective plane into four faces-a hexagon, and four quadragons. The remaining vertices and edges of \(G\) must be placed in one of these faces. If we delete \(V(T K 3,3)\) from \(G\), the result is a subgraph consisting of a number of connected components. If \(H\) is such a connected component, then since \(G\) is 2-connected, there must be at least two edges with one endpoint in \(H\) and the other in TKB, 3. DEFINITION 13.21: A bridge of \(G\) with respect to \(T K 3,3\) is either:
1. An edge \(u u\), where \(u, v \in V\left(T K_{3,3}\right)\) but \(u v \notin E\left(T K_{3,3}\right)\), or
2.A connected component \(H\) of \(G-V(T K 3,3)\) together with all edges connecting \(H\) to \(T K 3,3\) If \(B\) is a bridge, then a vertex of attachment of \(B\) is any vertex \(u\) of \(B\) such that \(u \in V\left(T K_{3,3}\right)\).

We can use a breadth-first (BFS) or depth-first search (DFS) to find the bridges of \(G\) with respect to an embedding of \(T K 3,3\). Each bridge has at least two vertices of attachment. Since each face is a 2 -cell, and each bridge must be embedded in a face of TK3, 3, each bridge must be planar. Into which faces of TK3, 3 can the bridges be placed?
The embedding of \(K 3,3\) in Figure 13.37 determines a classification of the vertices and edges of \(K 3\), 3 . Edges on the boundary of the hexagon are called hexagon
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edges. Edges which are on the boundary of a quadragon, but not on the hexagon are called quadragon edges. Hexagon edges like \(\{1,2\}\) and \(\{4,5\}\) are called opposite edges. Vertices like 1 and 4 are called diagonally opposite vertices, because they are diagonally opposite on the hexagon. By a path of TK3, 3 we mean a path connecting two corner vertices. The paths of TK 3,3 corresponding to hexagon edges of \(K 3,3\) are called hexagon paths, those corresponding to opposite edges of the hexagon are called opposite paths, and so forth. In general, a path of \(T K 3,3\) will be either a hexagon or a quadragon path of TK3, 3 . The following lemmas on bridges can be proved by considering all possibilities of placing a bridge in Figure 13.37. LEMMA 13.21 A bridge B can be placed in three faces of TK3, 3 if and only if \(B\) has exactly two vertices of attachment, which are diagonally opposite vertices.
LEMMA 13.22 A bridge B can be placed in two faces of TK3, 3 if and only if all vertices of attachment of B are on the same path, or on opposite paths of TK3, 3.
It follows from these lemmas that a bridge \(B\) can be placed in at most three faces, and that bridges for which these lemmas do not apply, either cannot be placed in any face, or can be placed in at most one face. A bridge is compatible with a face if it can be placed in the face. A bridge \(B\) is a \(k\)-face bridge if it can be placed in exactly \(k\) faces of \(T K 3,3\). Thus, we have 3 -face, 2 -face, 1 -face, and 0 -face bridges with respect to an embedding of \(T K 3,3\). We can determine which faces a bridge may embed in by using its vertices of attachment and the previous lemmas.
Two bridges \(B 1\) and \(B 2\) conflict in face \(F\) if they can both be placed in face \(F\), but cannot be simultaneously placed in face \(F\). Suppose that \(B 1\) can be embedded in face \(F\) and that it has \(k\) vertices of attachment \(u 1, u 2\), \(\ldots, u k\), where \(k \geq 2\), and where the vertices occur in that order on the facial cycle of \(F\). The vertices divide the facial cycle into \(k\) intervals [u1, U2], [ \(u 2, u 3\) ],..., [ \(u k-1, u k],[u k, u 1\) ], where each interval is a path from ui to \(u i+1\). If \(B 2\) is another bridge that can also be embedded in face \(F\), then \(B 1\) and \(B 2\) do not conflict if and only if all vertices of attachment of \(B 2\) lie in one interval of \(B 1\), and vice versa.
Suppose that \(B\) is a 3-face bridge, with vertices of attachment \(u\) and \(u\). All 3-face bridges with these vertices of attachment can be combined into one bridge, as they can always all be embedded in the same face if any embedding is possible. Thus, we can assume that there are at most three 3-face bridges, one for each pair of diagonally opposite vertices. Furthermore, any two distinct 3-face bridges conflict in the hexagon, so that at most one 3 -face bridge can be embedded in the hexagon. The algorithm looks for embeddings with no 3 -face bridges in the hexagon, or with exactly one 3-face bridge in the hexagon. Thus there are four subproblems to consider.
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If we choose a DFS to find the bridges, it can be organized as follows. The procedure uses a variable \(n B r i d g e s\) to count the number of bridges found so far. It is initially zero. We take each vertex \(u \in V\left(T K_{3,3}\right)\) in turn, and consider all incident edges \(u v \notin E\left(T K_{3,3}\right)\). Edge uu belongs to exactly one bridge. We store a value \(B(u u)\) for each edge, indicating the bridge to which \(u u\) belongs. If \(B(u u)=0\), then \(B(u u)\) has not yet been assigned. If \(v \in V\left(T K_{3,3}\right)\), then edge \(u u\) is a bridge, and we assign \(B(u u)\). Otherwise we call a procedure BRIDGEDFS \((u)\) to build the bridge \(B\). Because of the nature of a DFS, it will visit all vertices of \(B\) before returning, and will explore only edges of \(B\). For each edge \(x y \in B, B(x y)\) is assigned to be the current value of nBridges. Each time it encounters a vertex of \(T K 3,3\), it has found a vertex of attachment. For each bridge, a list of vertices of attachment is saved. If two bridges \(B\) and \(B^{\prime}\) are both found to have exactly two vertices of attachment, and they are the same two vertices, then \(B\) and \(B^{\prime}\) are combined into a single bridge. The vertices of attachment are later used to determine the compatible faces, using the previous lemmas, and to sort the adjacency list of each \(u\).

Algorithm 13.6.1: CONSTRUCTBRIDGES ( G,TK3, 3)
comment: Construct all bridges of \(G\) with respect to \(T K 3,3\)
\(n\) Bridges \(\leftarrow 0\)
for each \(u \in V\left(T K_{3,3}\right)\)
( for each edge \(u v\) such that \(u v \notin E\left(T K_{3,3}\right)\)
do \(\left\{\begin{array}{l}\text { if } B(u v)=0 \text { then } \\ \left\{\begin{array}{l}\text { comment: the bridge of } u v \text { has not been visited } \\ n B r i d g e s ~ \leftarrow n \text { Bridges }+1 \\ \text { if } v \in V\left(T K_{3,3}\right) \\ \text { then } B(u v) \leftarrow n B r i d g e s \\ \text { else BRIDGEDFS }(v)\end{array}\right.\end{array}\right.\)
comment: all bridges incident on \(u\) have now been constructed
We can now present an algorithm for determining the conflicts of bridges in a face \(F\). Let ( \(u 1, u 2, \ldots, u k\) ) denote the facial cycle of \(F\). We assign a numbering to the facial cycle, such that ui is numbered \(i\). This defines an ordering \(u 1<u 2<\ldots<u k\). For each bridge \(B\) that is compatible with \(F\), we sort the vertices of attachment according to this ordering. The purpose of this is to determine whether the vertices of attachment of each bridge lie completely within an interval of all other bridges. Let bmin denote the smallest vertex of attachment of bridge \(B\), and let bmax denote the largest. The algorithm then walks along the facial cycle from \(u k\) to \(u l\) and sorts the adjacency list of each ui. The edges incident on ui can be divided into the following three classes:
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1. Edges uiu belonging to a bridge \(B\) such that \(u i=b m i n\)
2. Edges uiu belonging to a bridge \(B\) such that \(b\) min \(<u i<b m a x\)
3. Edges uiu belonging to a bridge \(B\) such that \(u i=b\) max


FI GURE 13.38
Determining whether bridges conflict
Refer to Figure 13.38 . The adjacency list is ordered so that edges in the first class precede those in the second class, which precede those in the third class, and so that edges of each bridge are contiguous in each of the three classes. The edges in the first class are further sorted so that if uiu and uiu' belong to bridges \(B\) and \(B^{\prime}\), respectively, where \(b_{\max }<b_{\max }^{\prime}\), then uiu precedes uiu'. If \(b_{\max }=b_{\max }^{\prime}\), then uiu precedes uiu' if \(B\) has more vertices of attachment. The edges in the third class are further sorted so that if uiu and uiu' belong to bridges \(B\) and \(B^{\prime}\), respectively, where \(b_{\min }<b_{\min }^{\prime}\), then uiu precedes uiu'. If \(b_{\min }=b_{\min }^{\prime}\), then uiu precedes uiu' if \(B\) has fewer vertices of attachment.

In Figure 13.38 the \(u l u k\)-path of the facial cycle is drawn approximately vertically, and the bridges are placed to the left of the path. With this representation, the ordering of the adjacency lists appears as a clockwise circle drawn at each
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\(u i\). If \(u i=b\) min for some bridge \(B\), there can be several edges uiu belonging to bridge \(B\). The last such edge is saved as Bmin. Similarly, if \(u i=b\) max for some bridge \(B\), the first edge uiu of \(B\) is saved as Bmax.
The algorithm then walks along the facial cycle from \(u k\) to \(u l\). Every edge uiu such that \(B\) (uiu) is compatible with \(F\) is placed on a stack. When \(B m i n\), the last edge of bridge \(B\) is encountered, all edges of \(B\) are removed from the stack, and conflicts with \(B\) are determined. The algorithm stores a linked list of conflicting bridges for each bridge \(B\).
comment: \(\left\{\begin{array}{l}\text { Determine the conflicts among bridges incident on face } F \\ \text { The facial cycle of } F \text { is }\left(u_{1}, u_{2}, \ldots, u_{k}\right) \\ \text { The adjacency list of each } u_{i} \text { has been sorted }\end{array}\right.\)
for \(u i \leftarrow u k\) downto \(u l\)
\begin{tabular}{|c|}
\hline  \\
\hline
\end{tabular}

We prove that the algorithm works. Suppose that \(B\) is a bridge with \(b m a x=u i\) and \(b m i n=u j\), with extreme edges \(B\) max=uiu and \(B m i n=u j w\). All vertices of attachment of \(B\) lie between ui and uj. If no other bridge has an attachment \(u \ell\) here, such that \(u i \neq u \ell \neq u j\), then \(B\) does not conflict with other bridges. Consider the point in the algorithm when \(B\) min is reached. The algorithm will have stacked each edge of \(B\) incident on the facial cycle, including Bmin. It then removes all these edges. If there is no \(u \ell\) between \(u i\) and \(u j\), no conflicts are discovered. But if \(B^{\prime}\) is a bridge with a vertex of attachment \(u \ell\) in this range, then an edge \(u \ell x\) of \(B^{\prime}\) will have been stacked after Bmax and before Bmin. Since the
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edges of \(B^{\prime}\) are still on the stack while edges of \(B\) are being removed, \(b_{\min }^{\prime} \leq b_{\min }\). If \(b_{\min }^{\prime}<b_{\min }\), then \(B\) and \(B^{\prime}\) are in conflict, and this is discovered. If \(b_{\min }^{\prime}=b_{\min }\), then since \(B\) min precedes \(B_{\min }^{\prime}\) in the adjacency lists, we know that \(b_{\max }^{\prime} \geq b_{\max }\) If \(b_{\max }^{\prime}>b_{\max }\), then \(B\) and \(B^{\prime}\) are in conflict, and this is discovered.
Otherwise \(b_{\max }^{\prime}=b_{\max }\) and \(b_{\min }^{\prime}=b_{\min }\), and \(u \ell\) is strictly between these limits. The ordering of the adjacency list tells us that \(B\) has at least as many vertices of attachment as \(B^{\prime}\). Therefore \(B\) and \(B^{\prime}\) both have a vertex of attachment strictly between ui and \(u j\). We conclude that the bridges conflict.
Once all bridges have been constructed and all conflicts have been determined, we construct an instance of the 2-Sat problem to represent this embedding problem. The 2-Sat problem will have boolean variables corresponding to the placement of bridges, and boolean expressions to characterize the conflicts of bridges. Let the bridges be \(B 1, B 2, \ldots, B m\). If there are any 0 -face bridges, the embedding of \(T K 3,3\) cannot be extended. If \(B i\) is a 1-face bridge, embeddable in face \(F\), create a boolean variable xi for it. We require
\(x i=\) true, and consider this to mean that \(B i\) is assigned to face \(F\). If \(B i\) is a 2 -face bridge embeddable in faces \(F\) and \(F^{\prime}\), create boolean variables \(x i\) and \(y i\) for \(i\). We consider \(x i=\) true to mean that \(B i\) is assigned to \(F\) and \(y i=\) true to mean that \(B i\) is assigned to \(F^{\prime}\). Since we do not want \(x i\) and \(y i\) both to be true, or both to be false, we construct the clauses
\[
\left(x_{i}+y_{i}\right)\left(\bar{x}_{i}+\bar{y}_{i}\right)
\]

This ensures that exactly one of \(x i\) and \(y i\) will be true.
If \(B i\) is a 3 -face bridge, create boolean variables \(x i\), \(y i\), and \(z i\) for it as above, where \(z i=\) true means that \(B i\) is embedded in the hexagon. The 3-face bridges require special treatment. We take \(z i=\) false and \(z i=t r u e ~ a s ~\) separate cases. If \(z i=\) false, \(B i\) becomes a 2 -face bridge, and we construct the clauses
\[
\left(x_{i}+y_{i}\right)\left(\bar{x}_{i}+\bar{y}_{i}\right)
\]
to represent this bridge. If \(z i=\) true, \(B i\) becomes a 1-face bridge, and we require \(x i=y i=\) false.
If \(B i\) and \(B j\) are bridges that conflict in a face \(F\), suppose without loss of generality that xi represents the assignment of \(B i\) to face \(F\), and that wj represents the assignment of \(B j\) to face \(F\), where wj is one of \(x j, y j\), or \(z j\). We then construct the clause
\[
\left(\bar{x}_{i}+\bar{w}_{j}\right)
\]
to ensure that at most one of \(B i\) and \(B j\) can be placed in face \(F\).
When a variable is required to have a certain value (e.q., \(x i=\) true), we construct clauses
\[
\begin{gathered}
\left(x_{i}+x_{0}\right)\left(x_{i}+\bar{x}_{0}\right) \\
\text { page_378 }
\end{gathered}
\]

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where \(x 0\) is an additional boolean variable. Notice that this can only be satisfied if \(x i=\) true. If \(x i=\) false is required, we construct the clauses
\[
\left(\bar{x}_{i}+x_{0}\right)\left(\bar{x}_{i}+\bar{x}_{0}\right)
\]
which can only be satisfied if \(x i=\) false.
Thus, we can represent all conflicts and all placements of bridges by instances of 2-Sat. Suppose that an algorithm for 2-Sat finds a solution satisfying the constraints. If \(B 1, B 2, \ldots, B K\) are the bridges assigned to face \(F\), they are all mutually non-conflicting bridges. Consequently, all vertices of attachment of any Bi lie in an interval determined by two vertices of attachment of every \(B j\). If each \(B i\) has a planar embedding in \(F\), then they all have a common planar embedding in \(F\), and conversely. If there is no common planar embedding of the bridges in F, then some bridge Bi has no planar embedding in F. It then follows that Bi cannot be embedded in any other face \(F^{\prime}\). Thus, we can complete the projective embedding of \(G\) by determining whether there is a planar embedding of the bridges in \(F\). We construct a graph \(G(F)\) which is the union of the facial cycle of \(F\), and all bridges \(B 1, B 2, \ldots, B k\) assigned by 2 -Sat to \(F\). We add one additional vertex \(u 0\), joined to every vertex of the facial cycle, in order to distinguish an "inside" and "outside" for \(F\). We have: LEMMA 13.23 \(G(F)\) is planar if and only if bridges \(B 1, B 2, \ldots, B k\) have a common embedding in \(F\). We can now present the algorithm for projective planarity.

\section*{Algorithm 13.6.3: PROJ ECTIVEPLANARITY( \(G\),TK3,3)}
comment: \(\left\{\begin{array}{l}\text { Given a graph } G \text { with a subgraph } T K_{3,3}, \\ \text { determine whether } G \text { is projective. }\end{array}\right.\)
let \(n\) and \(\varepsilon\) denote the number of vertices and edges of \(G\)
\[
\text { if } \varepsilon>3 n-3
\]
then return (NonProjective)
construct the bridges \(B 1, B 2, \ldots, B m\) with respect to \(T K 3,3\)
for each embedding of \(T K 3,3\)
classify the bridges as 0-face, 1-face, 2-face, or 3-face if there is a 0 -face bridge go to \(L 1\)
determine all conflicts of bridges
construct the clauses representing all conflicts of bridges
wlog, assume that \(B_{1}, B_{2}\) and \(B_{3}\) are 3-face bridges
\(z_{1}, z_{2}, z_{3} \leftarrow\) false
for \(i \leftarrow 0\) to 3
\(\left\{\begin{array}{l}\text { construct the clauses representing } B_{1}, B_{2} \text { and } B_{3} \\ \text { solve the resulting 2-Sat problem } \\ \text { if there is no solution go to } L 2 \\ \text { for each face } F \text { of } T K_{3,3} \text { do } \\ \left\{\begin{array}{l}\text { take all bridges assigned to } F, \text { construct graph } G(F) \\ \text { if } G(F) \text { is non-planar go to } L 2\end{array}\right. \\ \text { comment: we now have a projective embedding } \\ \text { return (Projective) }\end{array}\right.\)

L2 :
\(z_{i} \leftarrow\) false
\(z_{i+1} \leftarrow\) true
\(L 1\) :
return (NonProjective)
The bridges \(B 1, B 2, \ldots, B m\) are constructed using a BFS or DFS. This takes \(O(n)\) steps, since \(\varepsilon \leq 3 n-3\). There are six embeddings of \(T K 3,3\) that are considered. For each embedding, the bridges are classified according to the faces they can be assigned to. The conflicts between bridges are then calculated. As the number of bridges \(m\) is bounded by \(n\), the number of conflicts is at most \(O(n 2)\). An instance of 2-Sat is then constructed with at most \(3 m+1\) variables and at most \(4 m+m(m-1)\) clauses. This can be solved in \(O(n 2)\) time. If a solution is found, a planar embedding algorithm must be used for each face to find the actual embedding. If a linear or quadratic planarity algorithm is used, the result is at most \(O(n 2)\) steps to complete the projective embedding. The result is a \(O(n 2)\)
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algorithm for finding a projective embedding of \(G\), if one exists, when we are given a \(T K 3,3 i n G\).
Now it would be possible to consider the embeddings of \(T K 5\) in a similar way, and construct the bridges with respect to each embedding of \(T K 5\), etc. There are 27 embeddings of \(T K 5\) in the projective plane. However, there is an easier way. In Section 12.3 we found that most graphs containing a subgraph TK5 also contain \(T K 3,3\) and that a simple breadth-first search algorithm can find a \(T K 3,3\), given a \(T K 5\). Thus, if we are given a TK5 in \(G\), we first try to find a TK3,3 in its place, and then use Algorithm PROJ ECTIVEPLANARITY() to extend it to \(G\).
If \(T K 5\) cannot be extended to \(T K 3,3\), then the structure of \(G\) is limited. Let \(\{u 1, u 2, u 3, u 4, u 5\}\) be the corners of \(T K 5\). Then \(G-\{\cup 1, \cup 2, \cup 3, \cup 4, u 5\}\) is a disconnected graph. Each component is adjacent to exactly two of the corner vertices of TK5. Let Gij denote the subgraph induced by all components adjacent to ui and uj, together with all edges connecting them to \(u i\) or \(u j\). Gij is called a K5-component of \(G\). Notice that ui and uj are vertices of \(G i j\), and that \(E(G)=\cup_{i, j} E\left(G_{i j}\right)\). An augmented \(K 5\)-component is the graph \(G_{i j}^{+}\)with the additional edge uiuj; namely, \(G_{i j}^{+}=G_{i j}+v_{i} v_{j}\). We have the following theorem:
THEOREM 13.24 Suppose that \(G\) has a subgraph TK5 which cannot be extended to TK3,3. Then \(G\) is projective if and only if all augmented K5-components \(G_{i j}^{+}\)are planar.
The proof of this theorem is left as an exercise. A consequence of it is that algorithms for projective planarity can focus on TK3,3 subgraphs, which have fewer embeddings. A similar, but more complicated result holds for toroidal graphs containing a \(T K 5\) which cannot be extended to \(T K 3,3\).

\section*{Exercises}
13.6.1 Show that a graph can be embedded on the projective plane if and only if it can be embedded on the Möbius band.
13.6.2 Show that \(K 3,3\) has six hamilton cycles. If \(C=(1,2,5,6,3,4)\) is a hamilton cycle, show that all hamilton
cycles can be obtained by successively applying the permutation \((1)(3,5)(2,4,6)\) to \(C\).
13.6.3 Show how to find a projective rotation system for a graph \(G\) containing \(T K 3,3\), When the algorithm PROJ ECTIVEPLANARITY() determines that \(G\) is projective. Hint: Use the projective embedding of \(K 3,3\) in Figure 13.39.
13.6.4 Prove Theorem 13.24.

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FI GURE 13.39

\section*{A projective embedding of \(K 3,3\)}

\subsection*{13.7 Heawood's map coloring theorem}

We conclude this chapter with a discussion of Heawood's map coloring theorem. The 4-color theorem states that \(\chi(G) \leq 4\), for graphs of genus zero. Heawood's map coloring theorem gives an analogous result for graphs of genus one or more.
LEMMA 13.25 Let \(n \geq 3\). Then \(g\left(K_{n}\right) \geq\lceil(n-3)(n-4) / 12\rceil\).
PROOF By Lemma 13.6, \(\varepsilon(K n)=n(n-1) / 2 \leq 3 n+6(g-1)\). Solving for \(g\) gives the result.
THEOREM 13.26 (Heawood's theorem) Let \(G\) be a graph on \(n\) vertices with genus \(g \geq 1\). Then \(\chi(G) \leq\left\lfloor\frac{1}{2}(7+\sqrt{1+48 g})\right\rfloor\).
PROOF Let \(\chi(G)=k\). If \(G\) is not a critical graph, then it contains a critical subgraph. Since a \(k\)-critical graph has minimum degree at least \(k-1\), we conclude that the sum of degrees of \(G\) is at least \((k-1) n\), so that \(\varepsilon \geq(k-1) n / 2\). Lemma 13.6 gives \(\varepsilon \leq 3 n+6(g-1)\). These two inequalities together give
\[
k \leq 7+\frac{12(g-1)}{n}
\]
with equality only if both inequalities above are equalities. Now \(g \geq 1\) so that for fixed \(g\), this is a nonincreasing function of \(n\), so that \(\chi(G)\) will be bounded.

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This arises because the number of edges in a \(k\)-critical graph increases as \(k n / 2\), whereas the maximum number of edges in a graph embedded in a surface of genus \(g\) increases as \(3 n\). We also know that \(k \leq n\), an increasing function. Therefore the largest possible value of \(k\) is when \(k=n=7+12(g-1) / n\). The equation then becomes \(n 2-7 n-12(g-1)=0\), which gives the solution \(n=\frac{1}{2}(7+\sqrt{1+48 g})\). Since \(k \leq n\) and \(k\) must be an integer, the result follows.
If \(\frac{1}{2}(7+\sqrt{1+48 g})\) is an integer, the inequalities used in the proof of Heawood's theorem must be equalities. This requires that \(\varepsilon=(n-1) n / 2\), so that \(G=K n\). The quantity \(\frac{1}{2}(7+\sqrt{1+48 g})\) represents the largest number of vertices that a graph can have, and still satisfy \(n(n-1) / 2 \leq 3 n+6(q-1)\). If it is not an integer, this means that \(n(n-1) / 2<3 n+6(g-1)\), so that a complete graph on \(n=\left\lfloor\frac{1}{2}(7+\sqrt{1+48 g})\right\rfloor\) vertices will not be a triangulation. In general, we have:
THEOREM 13.27 Let \(G\) be a graph on \(n\) vertices with genus \(g \geq 1\) and chromatic number
\(\chi(G)=\left\lfloor\frac{1}{2}(7+\sqrt{1+48 g})\right\rfloor\).Then \(G\) contains a spanning complete graph.
PROOF Let \(h=\left\lfloor\frac{1}{2}(7+\sqrt{1+48 g})\right\rfloor\). Let \(G\) have \(n\) vertices. Since \(\chi(G)=h\), we know that \(n \geq h\), and \(\varepsilon(G) \geq n(h-1) / 2 \geq h(h-1) / 2\). But \(h\) is the largest integer such that \(h(h-1) / 2 \leq 3 h+6(g-1)\). Therefore \(n=h\) and \(G\) contains a spanning complete graph. Note that a complete graph may not triangulate the surface, so that the number of edges in a triangulation, \(3 h+6(g-1)\), may be larger than \(h(h-1) / 2\).
We conclude that the extreme value of \(X\) will be achieved only if a complete graph with this many vertices can be embedded in the surface.
This theorem gives \(x(G) \leq 7\) for the torus. An embedding of \(K 7\) in the torus is shown in Figure 13.17, so that seven colors are necessary for the torus. We say that the chromatic number of the torus is seven, because all toroidal graphs can be colored in at most seven colors, and seven colors are necessary for some graphs. The dual of the embedding of \(K 7\) on the torus is the Heawood graph. Heawood was coloring the faces of an embedding rather than the vertices of a graph, and discovered this graph. The 4 -color theorem tells us that the chromatic number of the plane is four. The formula of Heawood's theorem gives the bound \(x(G) \leq 4\) for the plane. However, the proof is invalid when \(q=0\), and there are many planar graphs other than \(K 4\) which require four colors. Lemma 13.25 gives \(g\left(K_{n}\right) \geq\lceil(n-3)(n-4) / 12\rceil\). A proof that \(g(K n)\) equals this bounds would mean that \(K n\) can always be embedded in a surface of this genus. Since \(x(K n)=n\), the inequality of Heawood's theorem could then be replaced by an equality. Calculating the genus of Kn was accomplished by a number of people
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(Heffter, Gustin, Ringel, Youngs, and Mayer) over many years. We state the result, without proof, as the following theorem.
THEOREM 13.28 (Ringel-Youngs) Let \(n \geq 3\). Then \(g\left(K_{n}\right)=\lceil(n-3)(n-4) / 12\rceil\).
A complete proof of this result can be found in the survey paper of WHITE [124]. A consequence is the following:
THEOREM 13.29 (Heawood map coloring theorem) The chromatic number of an orientable surface of genus \(g \geq 1\) is \(\left\lfloor\frac{1}{2}(7+\sqrt{1+48 g})\right\rfloor\).
The corresponding results for non-orientable surfaces of genus \(\bar{g} \geq 1\) are as follows. Corresponding to Lemma 13.25 is the bound \(\bar{g}\left(K_{n}\right) \geq\lceil(n-3)(n-4) / 6\rceil\). Corresponding to Heawood's theorem is the bound \(\chi(G) \leq\left\lfloor\frac{1}{2}(7+\sqrt{1+24 \bar{g}})\right\rfloor\), which is proved in an analogous way. Again, the graphs which meet the bound are the complete graphs. The non-orientable version of the Ringel-Youngs theorem is the following:
THEOREM 13.30 Let \(n \geq 5\). Then \(\bar{g}\left(K_{n}\right)=\lceil(n-3)(n-4) / 6\rceil\), except that \(\bar{g}\left(K_{7}\right)=3\).
The formula \((n-3)(n-4) / 6\) gives \(\bar{g}\left(K_{7}\right) \geq 2\). However, \(\bar{g}\left(K_{7}\right)=3\), as \(K 7\) does not embed on the Klein bottle. The map coloring theorem for non-orientable surfaces is then:
THEOREM 13.31 The chromatic number of a non-orientable surface of genus \(\bar{g} \geq 1_{\text {is }}\left\lfloor\frac{1}{2}(7+\sqrt{1+24 \bar{g}})\right\rfloor\), except that the chromatic number of the Klein bottle is 6.

\section*{Exercises}
13.7.1 Let \(t\) be a positive integer, and let \(t K 3\) denote the graph with three vertices, and \(t\) parallel edges connecting each pair of vertices, so that \(\varepsilon(t K 3)=3 t\). Consider embeddings of \(t K 3\) in which there are no digon faces. Show that \(g(t K 3) \geq(t-1) / 2\) and that \(\bar{g}\left(t K_{3}\right) \geq t-1\).
13.7.2 Show that \(g(t K 3)=(t-1) / 2\) and \(\bar{g}\left(t K_{3}\right)=t-1\) by constructing embeddings of \(t K 3\) on the appropriate surfaces.

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13.7.3 Let \(G\) be a graph with \(n\) vertices and genus \(g\), and let \(n k\) denote the number of vertices of degree \(k\). Suppose that \(n 1=n 2=0\). Construct an inequality satisfied by \(n 3, n 4, \ldots\) in terms of \(g\), using the number of edges in a triangulation. Do the same for \(\bar{g}(G)\).
13.7.4 The maximum genus of a graph \(G\) is the largest value \(g\) such that \(G\) has a 2-cell embedding on a surface of genus \(g\). If \(g^{\prime}\) is the maximum genus of a graph \(G\) on \(n\) vertices, use the Euler-Poincaré formula to show that \(g^{\prime} \leq(\varepsilon-n+1) / 2\). Find the maximum orientable genus of \(K 4\).
13.7.5 Show that \(K 7\) does not embed on the Klein bottle.

\subsection*{13.8 Notes}

An excellent source book related to topology and geometry is HILBERT and COHN-VOSSEN [61]. It is perhaps one of the best and most readable mathematics books ever written. Proofs of the Dehn-Heegard theorem can be found in FRECHET and FAN [44] and in STILLWELL [109]. Both contain very readable accounts of combinatorial topology. Fréchet and Fan call the Euler-Poincaré formula Descartes' formula.
There are excellent chapters in DIESTEL [35] and ZIEGLER [129] on the graph minor theorem. The minor order obstructions for the projective plane were found by ARCHDEACON [6]. MYRVOLD [90] has found over 200,000 topological obstructions for the torus, and the list may not be complete.
An excellent source for graphs and surfaces is the book by MOHAR and THOMASSEN [88], or the book on topological graph theory by GROSS and TUCKER [54].
THOMASSEN [113] has proved that Graph Genus is NP-complete.
The algorithm for drawing graphs on the torus, given a toroidal rotation system is from KOCAY, NEI LSON, and SZYPOWSKI [76]. It is adapted from Read's algorithm READ [100] for planar graphs.
Theorem 13.18 relating projective embeddings to toroidal embeddings is from FIEDLER, HUNEKE, RICHTER, and ROBERTSON [41].
The algorithm for embedding a graph on the projective plane is based on algorithms by GAGARIN [48], MOHAR [87], and MYRVOLD and ROTH [91]. Theorem 13.24 is from GAGARIN and KOCAY [49].
The survey article by WHITE [124] contains a complete proof of the RingelYoungs theorem and the Heawood map coloring theorem.

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14
Linear Programming

\subsection*{14.1 Introduction}

Problems which seek to find a "best" configuration to achieve a certain goal are called optimization problems.
Programming problems deal with determining optimal allocations of limited resources to meet given
objectives. They deal with situations in which a number of resources such as manpower, materials, and land are available to be combined to yield one or more products. There are, however, restrictions imposed, such as the total number of resources available and the quantity of each product to be made. Linear programming deals with the class of programming problems in which these restrictions and any other relations among the variables are linear. In particular, the constraints imposed when searching graphs and networks for say shortest paths, matchings, and maximum flow can be expressed as linear equations.

\subsection*{14.1.1 A simple example}

Let us consider a furniture shop that produces tables, chairs, cabinets, and stools. There are three type of machines used: table saw, band saw, and drill press. We assume that the production is continuous and each product first uses the table saw, then the band saw, and finally the drill press. We also assume that setup time for each machine is negligible. Table 14.1 shows
1. The hours required on each machine for each product
2. The profit realized on the sale of each product

We wish to determine the weekly output for each product in order to maximize profit. Let \(x 1, x 2, x 3\), and \(x 4\) denote the number of tables, chairs, cabinets, and stools produced per week, respectively. We want to find the values of \(x 1, x 2, x 3, x 4\) which maximizes the profit. The available machine time is lim-

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\section*{TABLE 14.1 \\ Data for the simple example}

Machine type
\(\begin{array}{rr}\text { Table } & \text { Chair } \\ 1.5 & 1 \\ 1 & 5 \\ 1.5 & 3\end{array}\)
Cabinet
Stool
Time available table saw band saw drill press profit \(\quad 5.24\)
7.30

1
3.5
3.5 2000 8000
1
5000
\(8.34 \quad 4.18\)
ited so we cannot arbitrarily increase the output of any one product. Thus we must allocate machine hours among the products without exceeding the maximum number of machine hours available.
Consider the restriction imposed by the table saw. According to Table 14.1 it will be used a total of
\(1.5 \times 1+x 2+2.4 \times 3+x 4\)
hours per week, but can only be used for at most 2000 hours. This yields the linear inequality \(1.5 \times 1+x 2+2.4 \times 3+x 4 \leq 2000\).
Similarly, the band saw and drill press yield the following restrictions:
\(x 1+5 \times 2+x 3+3.5 x 4 \leq 8000\)
\(1.5 \times 1+3 \times 2+3.5 \times 3+x 4 \leq 5000\)
Furthermore, we can't produce a negative amount of a product, so we also have \(x 1 \geq 0, x 2 \geq 0, x 3 \geq 0\), and \(x 4 \geq 0\).
Now we have all the restrictions. The profit is
\[
z=5.24 \times 1+7.30 \times 2+8.34 \times 3+3.4 .18 \times 4 .
\]

Thus our goal is to solve the following linear program.
Maximize: \(z=5.24 \times 1+7.30 \times 2+8.34 \times 3+3.4 .18 \times 4\)
Subject to: \(1.5 \times 1+x 2+2.4 \times 3+x 4 \leq 2000\)
\(x 1+5 \times 2+x 3+3.5 \times 4 \leq 8000\)
\(1.5 \times 1+3 \times 2+3.5 \times 3+x 4 \leq 5000\)
\(x 1 \geq 0, x 2 \geq 0, x 3 \geq 0, x 4 \geq 0\)
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\subsection*{14.1.2 Simple graphical example}

We can graphically solve linear programs that involve only two variables. For example, consider the linear program
1. Solving \(x 1=0\) and \(3 x 1+5 \times 2=18\) gives the vertex \((0,3.6)\).
2. Solving \(3 \times 1+5 \times 2=18\) and \(5 \times 1+2 \times 2=11\) gives the vertex \((1,3)\).
3. Solving \(5 x 1+2 \times 2=11\) and \(x 2=0\) gives the vertex \((2.2,0)\).
4. Solving \(x 1=0\) and \(x 2=0\) gives the vertex \((0,0)\).

The set of feasible solutions is depicted in Figure 14.1. We have also drawn the objective function \(z=5 \times 1+3 \times 2\), when \(z=3.5,11\), and 38.5.
Consider any line segment joining points \(P=(p 1, p 2)\) and \(Q=(q 1, q 2)\). Let
\(z 0=5 p 1+3 p 2\)
be the value of the objective function at \(P\) and let
\[
z 1=5 q 1+3 q 2
\]
be the value of the objective function at \(Q\). Assume without loss that \(z 0 \leq z 1\). The coordinates of any point on the line segment between \(P\) and \(Q\) is given by
\[
((1-t) p 1+t q 1,(1-t) p 2+t q 2
\]
for \(0 \leq t \leq 1\). The value of the objective function at this point is
\[
\begin{gathered}
z t-5((1-t) p 1+t q 1)+3((1-t) p 2+t q 2) \\
=(1-t)(5 p 1+3 p 2)+t(5 q 1+3 q 2) \\
=(1-t) z 0+t 21 \\
\text { page_389 }
\end{gathered}
\]


FI GURE 14.1 Graphical example
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Observe that
and
\[
z 0=(1-t) z 0+t z 0 \leq(1-t) z t+t z 1=z t
\]

Thus \(z 0 \leq z t \leq z 1\) for any \(0 \leq t \leq 1\). Therefore the maximum value of the the objective function among the points on any line segment occurs at one of the endpoints. It follows that the maximum value of any (linear) objective function among the points of a compact region occurs at some point on the boundary. Also, if the boundary is a polygon, then the maximum value of the objective will occur at one of the vertices. Hence for our example the maximum value of the objective function
\[
z=5 \times 1+3 \times 2
\]
occurs at least one of \((0,3.6),(1,3),(2.2,0)\), or \((0,0)\). The value at these points is \(10.8,14,11\), and 0 . Thus the maximum occurs at \(x 1=1, x 2=3\).
If the region of feasible solutions is unbounded, then there may be no point in the region for which the objective function achieves its absolute maximum. Of course there is also no point for which the objective function achieves its maximum, if there are no feasible solutions.
If the objective function achieves its maximum at two adjacent vertices \(P\) and \(Q\), and the region of feasible solutions is connected and polygonally bounded, then it will achieve its maximum at infinitely many points: namely, those lying on the line segment joining \(P\) and \(Q\).
In a general linear program the region of feasible solutions is the intersection of the half hyper-planes determined by the linear constrains. Thus the region of feasible solutions to the general linear program is a compact convex region, bounded by facets (hyper-plane segments). That is, it is a polyhedron. 1 Consequently
it follows that:
THEOREM 14.1 In a general linear program the objective function achieves its maximum either at exactly one point, at infinitely many points, or at no point. Furthermore if it does achieve a maximum, it does so on one of the vertices of the polyhedral boundary of the feasible solutions.

\subsection*{14.1.3 Slack and surplus variables}

It is easier to work with equalities, than with inequalities because we can take advantage of linear algebra. We can convert linear inequalities into equalities by introducing surplus and slack variables.
10.K. it is only a polyhedron if it is bounded. It could have some open sides, but the closed sides are bounded by hyper-planes.
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For example, consider an inequality of the form
\[
\begin{equation*}
a 1 x 1+a 2 x 2+\ldots+a t x t \leq b \tag{14.1}
\end{equation*}
\]

Given fixed assignments to the variables \(x 1, x 2, \ldots, x t\) that satisfy this inequality, there will be slack or "room for improvement" amounting to
\[
x j=b-a 1 x 1+a 2 x 2+\ldots+a t x t \geq 0 .
\]

Thus introducing a new variable \(x j\) and requiring \(x j \geq 0\) we obtain the equality
\[
a 1 \times 1+a 2 \times 2+\ldots+a t x t+x j=b
\]
which is equivalent to Inequality 14.1.
The variable \(x j\) is called a slack variable. Similarly an inequality of the form
\[
\begin{equation*}
a 1 \times 1+a 2 \times 2+\ldots+a t x t \geq b \tag{14.2}
\end{equation*}
\]
represents a surplus of
\[
s^{\prime}=a 1 x 1+a 2 \times 2+\ldots+a t x t-b
\]
for a fixed assignment to the variables \(x 1, x 2, \ldots, x t\). Thus introducing \(x j^{\prime}\) as a new variable and requiring \(x j\) \(\geq 0\) we obtain the equality
\[
a 1 x 1+a 2 \times 2+\ldots+a t x t-x j^{\prime}=b
\]
which is equivalent to Inequality 14.2. The variable \(x j^{\prime}\) is called a surplus variable. Adding slack and surplus variables in this way will reduce the the system of inequalities to a system of equalities and variables xi that are either unbounded or satisfy \(x i \geq 0\). If \(x i\) is unbounded, we find an inequality that \(x i\) satisfies, solve for \(x i\), and substitute to eliminate xi from the set of equations. A linear program in which all the variables are required to be non-negative and the remaining constraints are equality constraints is said to be in standard form or a standard linear program.
For example, to convert the linear program
\[
\begin{gathered}
\text { Maximize: } z=5 \times 1+3 \times 2+3 \times 3+x 4 \\
\text { Subject to: } 2 \times 2+x 4=2 \\
x 1+x 2+x 4 \leq 3 \\
-x 1-2 \times 2+x 3 \geq 1 \\
x 1 \leq 0, \times 2, x 3 \geq 0 \\
\text { page_392 }
\end{gathered}
\]

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into standard form we introduce slack and surplus variables \(x 5\), \(x 6\), and \(x 7\) obtaining
\[
\text { Maximize: } z=5 \times 1+3 \times 2+3 \times 3+x 4
\]
\[
\begin{gathered}
\text { Subject to: } 2 \times 2+x 4=2 \\
x 1+x 2+x 4+x 5=3 \\
-x 1-2 \times 2+x 3-x 6=1 \\
x 1+x 7=0
\end{gathered}
\]
\(x 2, x 3, x 5, x 6, x 7 \geq 0\)
Now variables \(x 1\) and \(x 4\) are unbounded, so we solve for them
\[
\begin{gathered}
x 1=-x 7 \\
x 4=2-2 \times 2
\end{gathered}
\]

Substituting, we obtain
Maximize: \(z=x 2+3 \times 3-35 \times 7+2\)
Subject to: \(-x 2+x 5-x 7=1\)
\(-2 \times 2+x 3-x 6=x 7=1\)
\(x 2, x 3, x 5, x 6, x 7 \geq 0\).
Finally, to convert to a problem of minimization, we set
\[
Z=2-z
\]
and we have
a linear program in standard form. In matrix form we set
\[
\begin{gathered}
X=[x 2, x 3, x 5, x 6, x 7] T \\
C=[-1,-3,0,0,35] T \\
b=[1,1] T \\
A=\left[\begin{array}{rrrr}
-1 & 0 & 1 & 0 \\
-2 & 1 & 0 & -1 \\
-1
\end{array}\right]
\end{gathered}
\]
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and we see that the original linear program is equivalent to
\[
\begin{aligned}
& \text { Minimize: } Z=c T X \\
& \text { Subject to: } A X=b \\
& X \geq 0
\end{aligned}
\]

\section*{Exercises}
14.1.1 Solve the following graphically and shade the region representing the feasible solutions:

Maximize: \(z=x 1+1.5 \times 2\)
Subject to: \(2 \times 1+3 \times 2 \leq 6\)
\(x 1+4 \times 2<4\)
\(x 1, x 2 \geq 0\)
14.1.2 Solve the following graphically and shade the region representing the feasible solutions:

Minimize: \(Z=6 \times 1+4 \times 2\)
Subject to: \(2 \times 1+x 2 \geq 1\)
\(3 x 1+4 \times 2 \geq 1.5\)
\(x 1, x 2 \geq 0\)
14.1.3 Carefully examine Exercises 14.1 .1 and 14.1.2. How are the solutions related? They form what is called a pair of dual problems. Note that they involve the same constants, but in a rearranged order.
14.1.4 Put the following linear program into standard form:

Maximize: \(z=2 \times 1+3 \times 2+5 \times 3\)
Subject to: \(3 \times 1+10 \times 2+5 \times 3 \leq 15\)
\(33 \times 1-10 \times 2+9 \times 3<33\)
\(x 1+2 \times 2+x 3 \geq 4\)
\(x 1, x 2 \geq 0\)
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\subsection*{14.2 The simplex algorithm \\ \subsection*{14.2.1 Overview}}

The simplex algorithm can be used to solve linear programs of the form Minimize: \(Z=c T X\)
Subject to: \(A X=b\)
\(X \geq 0\)
There are three phases to the algorithm.
Phase 0: Find a basis solution or show that the linear program is infeasible.
Phase 1: Find a basis feasible solution or show that the linear program is infeasible.
Phase 2: Improve the basis feasible solution until
1. it's optimal, or
2. the linear program can be shown to be unbounded.

\subsection*{14.2.2 Some notation}

In order to study the linear program in Equation 14.3 we first introduce some notation. Denote the columns of the \(m\) by \(n\) matrix \(A\) by
\[
A=[A 1, A 2, A 3, \ldots, A n] .
\]

Without loss, assume that the rank of \(A\) is \(m\). Suppose
\[
\mathcal{B}=\left\{A_{j_{1}}, A_{j_{2}}, A_{j_{3}}, \ldots A_{j_{m}}\right\}
\]
is a set of \(m\) linearly independent columns of \(A\). Then the \(m\) by \(m\) matrix
\[
B=\left[A_{j_{1}}, A_{j_{2}}, A_{j_{3}}, \ldots A_{j_{m}}\right]
\]
is nonsingular. Thus \(B-1\) exists. Consequently, \(B-1 A\) contains the identity matrix on columns \(j 1, j 2, j 3, \ldots, j m\) of \(A\). Let \(X B=B-1 b\). Then \(X\) given by
\[
X[j]=\left\{\begin{array}{cl}
X_{B}[i] & \text { if } j=j_{i} \\
0 & \text { if not }
\end{array}\right.
\]
satisfies
\[
A X=b
\]
and is called the basis solution given by \(B\). If \(X B \geq 0\), then \(X\) also satisfies \(x \geq 0\)
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with the remaining constraints of the linear program given in Equation 14.3. Thus in this case we call X a basis feasible solution. Also corresponding to \(B\) we define \(c B\) by
\[
C B[i]=c[j i] \text {, for } i=1,2, \ldots, m
\]

The value \(Z\) of the objective function at \(X\) is
\[
\begin{gathered}
Z=c^{T} X=c_{B}^{T} X_{B} \\
z_{j}=c_{B}^{T} Y_{j} ;
\end{gathered}
\]
where \(Y j=B-1 A j\). Using this notation we can safely identify \(X B\) with \(X\) and refer to \(X B\) as the basis solution.

\subsection*{14.2.3 Phase 0: finding a basis solution}

A basis solution if one exists can be found by pivoting. Pivoting is also known as Gaussian elimination or row reduction. To pivot on non-zero entry aij of the \(m\) by \(n\) matrix \(A\) we replace row \(k\) by
\[
a_{i j}\left[a_{k 1}, a_{k 2}, \ldots, a_{k n}\right]-a_{k j}\left[a_{i 1}, a_{i 2}, \ldots, a_{i n}\right]
\]
for \(k \neq i\) and we replace row \(i\) by
\[
\frac{1}{a_{i j}}\left[a_{i 1}, a_{i 2}, \ldots, a_{i n}\right] .
\]

The result is that column \(j\) becomes
\[
[0,0, \ldots, 0, \underbrace{1}_{j \text { th }}, 0,0, \ldots, 0]^{T} .
\]

We record this procedure as Algorithm 14.2.1.
Algorithm 14.2.1: \(\operatorname{PIVOT}(i, j)\) for \(k \leftarrow 1\) to \(m\)
```

if $k \neq i$
do
then $\left\{\begin{array}{l}\text { for } h \leftarrow 0 \text { to } n \\ \text { do } A[k, h] \leftarrow A[i, j] \cdot A[k, h]-A[k, j] \cdot A[i, h]\end{array}\right.$
else $\left\{\begin{array}{l}\text { for } h \leftarrow 0 \text { to } n \\ \text { do } A[k, h] \leftarrow A[k, h] / A[i, j]\end{array}\right.$

```
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Thus to find a basis solution we iteratively select a non-zero entry \(a_{i j_{i}}\) in row \(i\) and pivot on \(a_{i j_{i}}\) for \(i=1,2\), ..., \(m\). That is, we have selected the linearly independent columns
\[
\left\{A_{j_{1}}, A_{j_{2}}, A_{j_{3}}, \ldots, A_{j_{m}}\right\}
\]
of \(A\) and have determined the basis solution \(X B\); where
\[
B=\left[A_{j_{1}}, A_{j_{2}}, A_{j_{3}}, \ldots A_{j_{m}}\right] .
\]

\subsection*{14.2.4 Obtaining a basis feasible solution}

Suppose that \(A=[A 1, A 2, \ldots, A n]\) has rank \(m\) and that \(X\) is a feasible solution to the linear program in Equation 14.3. Let \(p \leq n\) be the number of positive variables in \(X\). Without loss we may assume that the first \(p\) variables are positive. Then
\[
X=[x_{1}, x_{2}, \ldots, x_{p}, \underbrace{0,0, \ldots, 0}_{n-p}]^{T}
\]
and so
\[
\begin{equation*}
\sum_{j=1}^{p} x_{j} A_{j}=b \tag{14.4}
\end{equation*}
\]

If the \(p\) columns \(A 1, A 2, \ldots, A p\) are linearly independent, then \(p \leq m\), and there exists an additional \(m-p\) columns of \(A\) whose inclusion with the first \(p\) are a linear independent set. Thus we can form a basis solution with \(m-p\) zeros.
If the \(p\) columns are linearly dependent, then there exists aj not all zero such that
\[
\sum_{j=1}^{p} \alpha_{j} A_{j}=0
\]

Let \(\operatorname{Ar}\) be any column for which \(a r \neq 0, i=1,2, \ldots, p\). Then
\[
\begin{equation*}
A_{r}=-\sum_{\substack{j=1 \\ j \neq r}}^{p} \frac{\alpha_{j}}{\alpha_{r}} A_{j} \tag{14.5}
\end{equation*}
\]

Substituting Equation 14.5 into Equation 14.5, we obtain
\[
\sum_{\substack{j=1 \\ j \neq r}}^{p}\left(x_{j}-x_{r} \frac{\alpha_{j}}{\alpha_{r}}\right) A_{j}=b
\]
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Thus we have a solution with at most \(p-1\) non-zero entries. However, we are not certain that they are nonnegative. We need
\[
x_{j}-x_{r} \frac{\alpha_{j}}{\alpha_{r}} \geq 0
\]
for \(j=1, \ldots, p, j \neq r\). If \(a j=0\), then this is automatically satisfied. If \(a j \neq 0\), then dividing by \(a j\) gives
\[
\frac{x_{j}}{\alpha_{j}}-\frac{x_{r}}{\alpha_{r}} \geq 0 \quad \text { if } \alpha_{j}>0
\]
and
\[
\frac{x_{j}}{\alpha_{j}}-\frac{x_{r}}{\alpha_{r}} \leq 0 \quad \text { if } \alpha_{j}<0
\]

Thus we may select Ar such that
\[
\frac{x_{r}}{\alpha_{r}}=\operatorname{MIN}_{j}\left\{\frac{x_{j}}{\alpha_{j}}: \alpha_{j}>0\right\}
\]
or such that
\[
\frac{x_{r}}{\alpha_{r}}=\operatorname{MAx}_{j}\left\{\frac{x_{j}}{\alpha_{j}}: \alpha_{j}<0\right\}
\]

Then each entry in Equation 14.6 will be non-negative and a feasible solution with no more than \(p-1\) nonzero variables has been found. We can continue this process of selecting columns \(A j\) until \(p \leq m\) in which case a basis feasible solution has been found.

\subsection*{14.2.5 The tableau}

The initial tableau for the linear program in Equation 14.3 is the array
\[
\left[\begin{array}{c|c}
A & b  \tag{14.7}\\
\hline c^{T} & 0
\end{array}\right] .
\]

Note that other authors use a more elaborate tableau, but this is sufficient. Suppose \(B\) is the submatrix of \(A\) corresponding to the basis \(B\) as in Section 14.2.2. For the purpose of explanation assume \(B\) is the first \(m\) columns of \(A\); then the tableau has the form
\[
\left[\begin{array}{r|lll|l}
B & \cdots & A_{j} & \cdots & b \\
\hline c_{B}^{T} & \cdots & c_{j} & \cdots & 0
\end{array}\right] ;
\]
where \(A j\) is the \(j\) th column of \(A\). Multiplying the tableau with
\(\left[\begin{array}{c|c}B^{-1} & 0 \\ \hline-c_{B}^{T} B^{-1} & 1 \\ \text { page } & 398\end{array}\right]\)

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we obtain the new tableau
\[
\left[\left.\begin{array}{c|c|c|c}
I & \cdots & Y_{j} & \cdots  \tag{14.8}\\
X_{B} \\
\hline 0 & \cdots & c_{j}-z_{j} & \cdots
\end{array} \right\rvert\,-z .\right.
\]

This is because \(Y j=B-1 A j, X B=B-1 b\), and \(z_{j}=c_{B}^{T} Y_{j}\) as defined in Section 14.2.2. Thus selecting a new basis solution is equivalent to pivoting on entries of the first \(m\) rows of the tableau 14.7 to obtain a tableau similar to the tableau 14.8. (The identity matrix need not be among the first columns.) Thus in the process of selection of a basis the values of \(c j-z j, z\), and \(X B\) are also easily determined.
An algorithm for the Phase 0 portion of the simplex algorithm that takes as input a tableau for a linear program whose adjacency matrix has \(m\) rows and \(n\) columns is provided in Algorithm 14.2.2.

Algorithm 14.2.2: PHASEO (Tableau, \(m, n\) )
Infeasible-false
for \(r \leftarrow 1\) to \(m\)
\((c \leftarrow 1\)
while Tableau \([r, c]=0\) do \(c \leftarrow c+1\)
if \(c>n\)
do
if Tableau \([n+1] \neq 0\)
then \(\left\{\begin{array}{l}\text { comment: the linear program is infeasible } \\ \text { Infeasible } \leftarrow \text { true } \\ \text { exit }\end{array}\right.\)
then
comment: \(\left\{\begin{array}{l}\text { the linear program has } \\ \text { rank }<M \text {-shift up the rows }\end{array}\right.\)
for \(i \leftarrow r+1\) to \(m\)
else
do for \(j \leftarrow 1\) to \(n+1\)
do Tableau \([i-1, j] \leftarrow\) Tableau \([i, j]\)
\(m \leftarrow m-1\)
\(r \leftarrow r-1\)
else \(\left\{\begin{array}{l}\operatorname{PIVOT}(r, c) \\ \text { pivots }[r] \leftarrow c\end{array}\right.\)
14.2.6 Phase 2: improving a basis feasible solution

Let \(X B=B-1 b\) be a basis feasible solution of the linear program given in Equation 14.3. The value of the objective function at \(X B\) is
\[
Z=c_{B}^{T} X_{B}
\]
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Some questions naturally arise:
Can we find another basis feasible solution with better Z?
Furthermore, can we do this by changing only one column of B?
Can we remove one column Brof \(B\) and replace it with a column \(A j\) of \(A\) and get a smaller value \(Z\) of the objective function \(Z=c T X\) ?
Column \(A j\) is a linear combination of the columns in \(B\) because \(B\) is a nonsingular \(m\) by \(m\) submatrix of \(A\). Thus
\[
A_{j}=\sum_{i=1}^{m} y_{i j} B_{i} .
\]
\[
y r j \neq 0 .
\]

Then \(A j\) can replace any \(B r\) for which because the new set of vectors
\[
\{B 1, B 2, \ldots, B r-1, A j, B r+1, B r+2, \ldots, B m\}
\]
will be linearly independent. Let
\[
B^{\star}=\left[B_{1}, B_{2}, \ldots, B_{r-1}, A_{j}, B_{r+1}, B_{r+2}, \ldots, B_{m}\right] .
\]

Then \(X_{B^{*}}\) is a basis solution, but it may not be feasible. Observe from Equation 14.9 that
\[
B_{r}=\frac{1}{y_{r j}} A_{j}-\sum_{\substack{i=1 \\ i \neq r}}^{m} \frac{y_{i j}}{y_{r j}} B_{i}
\]

Also
\[
b=B X_{B}=\sum_{i=1}^{m} X_{B}[i] B_{i}=\sum_{\substack{i=1 \\ i \neq r}}^{m} X_{B}[i] B_{i}+X_{B}[r] B_{r}
\]

So, substituting we have:
\[
b=\sum_{\substack{i=1 \\ i \neq r}}^{m} X_{B}[i] B_{i}+X_{B}[r]\left(\frac{1}{y_{r j}} A_{j}-\sum_{\substack{i=1 \\ i \neq r}}^{m} \frac{y_{i j}}{y_{r j}} B_{i}\right)
\]

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\[
=\sum_{\substack{i=1 \\ i \neq r}}^{m}\left(X_{B}[i]-X_{B}[r] \frac{y_{i j}}{y_{r j}}\right) B_{i}+\frac{X_{B}[r]}{y_{r j}} A_{j}
\]

Thus
\[
X_{B}[i]=\left\{\begin{array}{cc}
X_{B}[i]-X_{B}[r] \frac{y_{i j}}{y_{r j}} & \text { if } i \neq r  \tag{14.10}\\
\frac{X_{B}}{y_{r j}} & \text { if } i=r
\end{array}\right.
\]
is feasible if and only if
\[
\begin{equation*}
X_{B}[i]-X_{B}[r] \frac{y_{i j}}{y_{r j}} \geq 0 \tag{14.11}
\end{equation*}
\]
and
\[
\begin{equation*}
\frac{X_{B}[r]}{y_{r j}} \geq 0 . \tag{14.12}
\end{equation*}
\]

Thus if \(X B[r] \neq 0\), we see from Equation 14.12 that we must have
If \(y i j \leq 0\), then Equation 14.11 automatically holds. So, we need to only be concerned with coordinates \(i\) for which \(y i j>0\). When \(y i j>0\), the condition given by Equation 14.11 can be rewritten as
\[
\frac{X_{B}[r]}{y_{r j}} \leq \frac{X_{B}[i]}{y_{i j}} .
\]

Thus we need to choose that column \(r\) of \(B\) such that
\[
\begin{equation*}
\frac{X_{B}[r]}{y_{r j}}=\operatorname{MiN}_{i}\left\{\frac{X_{B}[i]}{y_{i j}}: y_{i j}>0\right\}=\theta \tag{14.13}
\end{equation*}
\]

To summarize:
We began with a nonsingular submatrix \(B\) of \(A\)
\[
B=\left[A_{i_{1}}, A_{i_{2}}, A_{i_{3}}, \ldots A_{i_{m}}\right]
\]

If \(X B=B-1 b\) is a basis feasible solution to the linear program given in Equation 14.3, then we selected an arbitrary column \(A j\) of \(A\), not in \(B\) and wrote
\[
A j=B Y j,
\]
a linear combination of the columns of \(B\), where
\[
Y j=[y 1 j, y 2 j, \ldots, y m j] .
\]

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If some \(y i j>0\), there is a column \(B r\) of \(B\), which we can replace with \(A j\) to get a basis feasible solution \(X_{B^{*}}\). Equation 14.13 shows how to select \(B r\) so that this is possible.
Now what about the value of the objective function-is it better? Let
\[
B^{\star}=\left[B_{1}^{\star}, B_{2}^{\star}, \ldots, B_{m}^{\star}\right]
\]
be the new matrix obtained by replacing \(B r\) with \(A j\). That is \(B_{i}^{\star}=B_{i}\), for \(i \neq r\) and \(B_{r}^{*}=A_{j}\). The new basis feasible solution is \(X_{B^{\star}}\) given in Equation 14.10. The objective function corresponding to \(B^{\star}\) is
\[
Z^{\star}=c_{B^{\star}}^{T} X_{B^{\star}}
\]
where
\[
c_{B^{*}}[i]=\left\{\begin{array}{cl}
c_{B}[i] & \text { if } i \neq r \\
c_{j} & \text { if } i=r
\end{array}\right.
\]

Therefore
\[
\begin{aligned}
Z^{\star}= & \sum_{i=1}^{m}\left(X_{B}[i]-X_{B}[r] \frac{y_{i j}}{y_{r j}}\right) c_{B}[i]+\frac{X_{B}[r]}{y_{r j}} c_{j} \\
= & \sum_{i=1}^{m}\left(X_{B}[i]-X_{B}[r] \frac{y_{i j}}{y_{r j}}\right) c_{B}[i]+\frac{X_{B}[r]}{y_{r j}} c_{j}, \\
& \quad \text { because } X_{B}[i]-X_{B}[r] \frac{y_{i j}}{y_{r j}}=0 \text { when } i=r \\
= & Z-\frac{X_{B}[r]}{y_{r j}} \sum_{i=0}^{m} y_{i j} c_{B}[i]+\frac{X_{B}[r]}{y_{r j}} c_{j} \\
= & Z+\frac{X_{B}[r]}{y_{r j}}\left(c_{j}-c_{B}^{T} Y_{j}\right)
\end{aligned}
\]

Setting \(z_{j}=c_{B}^{T} Y_{j}=c_{B}^{T} B^{-1} A_{j}\) we have
\[
Z^{\star}=Z+\frac{X_{B}[r]}{y_{r j}}\left(c_{j}-z_{j}\right)=Z+\theta\left(c_{j}-z_{j}\right)
\]

Therefore if we can find \(A j\) such that \(c j-z j<0\)
and at least one \(y i j>0\), then it is possible to replace one of the columns of the columns of \(B\) by \(A j\) and obtain a new value \(Z^{\star}\) of the objective function satisfying
\[
Z^{\star} \leq Z
\]

If the given basis solution is not degenerate, then
\[
Z^{\star}<Z .
\]

In terms of the tableau given in Equation 14.8, this means we can find a column \(j\) with a negative entry in last row and a positive entry yij in row \(i\). For each positive entry yij compute the ratio
\[
\theta_{i}=\frac{X_{B}[i]}{y_{i j}}
\]
and chose \(i\) so that \(\theta i\) is smallest. Recall that
\[
[X B[1], X B[2], . ., X B[m],-Z] T
\]
is the last column of the tableau. Then pivoting on yij produces a new tableau with smaller \(Z\). Note that \(-Z\) is the entry in the last row and column of the tableau.

\subsection*{14.2.7 Unbounded solutions}

In Section 14.2.6, given a basis feasible solution \(X B=B-1 b\) we found a column \(A j\) that had at least one yij>0, \(i=1,2, \ldots, m\) where
\[
Y j=B-1 A j .
\]

For this column, we found a column \(B r\) in \(B\) which when replaced by \(A j\), resulted in a basis feasible solution \(X_{B^{*}}\). The value of the objective function for \(X_{B^{*}}\) was
\[
Z^{\star}=Z+\theta\left(c_{j}-z_{j}\right)
\]

Let us consider a column \(A j\) for which \(y i j \leq 0\) for each \(i=1,2, \ldots, m\). We have
\[
b=B X_{B}=\sum_{i=1}^{m} X_{B}[i] B_{i}
\]
with value of the objective function equal to \(Z=c_{B}^{T} X_{B}\).
Adding and subtracting \(\theta A j\) for any theta yields
\[
\begin{aligned}
b & =B X_{B}-\theta A_{j}+\theta A_{j} \\
& =\sum_{i=1}^{m} X_{B}[i] B_{i}-\theta A_{j}+\theta A_{j}
\end{aligned}
\]
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but
\[
A_{j}=B Y_{j}=\sum_{i=1}^{m} y_{i j} B_{j}
\]

So substituting we obtain:
\[
\begin{equation*}
b=\sum_{i=1}^{m}\left(X_{B}[i]-\theta y_{i j}\right) B_{i}+\theta A_{j} \tag{14.14}
\end{equation*}
\]

When \(\theta>0\), then \((X B[i j B i-\theta y i j) \geq 0\) because we have assumed that \(y i j \leq 0\), for \(i=1,2, \ldots, m\). Thus Equation 14.14 is feasible. The value of the objective function is again
\[
\begin{aligned}
Z^{\star} & =\sum_{i=1}^{m} c_{B}[i]\left(X_{B}[i]-\theta y_{i j}\right)+c_{j} \theta \\
& =Z+\theta\left(c_{j}-z_{j}\right)
\end{aligned}
\]

Thus choosing \(\theta\) arbitrarily large we can make \(Z^{\star}\) arbitrarily small if
\[
c j-z j<0
\]
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\section*{To summarize:}

Given any basis feasible solution to a linear program, if there is a column \(A j\) not in the basis for which \(c j-z j<0\)
and
\[
Y j=B-1 A j \leq 0,
\]
then the linear program in Equation 14.3 has an unbounded solution.

In terms of the tableau given in Equation 14.8, this means that if we can find a column \(j\) with every entry less than or equal to zero with a negative entry in last row, then the linear program in Equation 14.3 has an unbounded solution.

\subsection*{14.2.8 Conditions for optimality}

Assume that \(X B=B-1 b\) is a basis feasible solution of the linear program given in Equation 14.3 and that the value of the objective function at \(X B\) is
\[
\begin{gathered}
Z_{0}=c_{B}^{T} X_{B} . \\
c j-z j \geq 0
\end{gathered}
\]

In addition, suppose that
for every column \(A j\) of \(A\) not in \(B\). Thus the value of the objective function cannot be improved by replacing a column of \(B\) with a column of \(A\). We will show that \(Z 0\) is the minimal value of the linear program and hence that \(X B\) is an optimal solution. Set \(\vec{z}=\left[z_{1}, z_{2}, \ldots, z_{n}\right]^{T}\). So,
\[
\vec{z} \leq c
\]

Let \(X\) be any feasible solution of linear program given in Equation 14.3. Then
\[
\begin{equation*}
x 1 A 1+x 2 A 2+\ldots+x n A n=b \tag{14.15}
\end{equation*}
\]
and \(X \geq 0\). Let \(Z^{\star}=C^{T} X\) be the value of the objective function at \(X\).
Every column \(A j\) of \(A\) can be written as a linear combination of the columns of \(B\) :
\[
\begin{gathered}
\begin{array}{c}
A j=B Y j \\
\mathrm{dd} \\
\vec{z} \\
B X B=c_{B}^{T} Y . \text { Then } \\
B X X=B Y X \\
\text { page_405 }
\end{array} .
\end{gathered}
\]

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Therefore,
\[
X B=Y X
\]
because \(B\) is nonsingular. Hence
\[
Z_{0}=c_{B}^{T} X_{B}=c_{B}^{T} Y X=\vec{z} X \leq c^{T} X=Z^{\star}
\]
which proves that \(Z O\) is the optimal value of the objective function and hence \(X B\) is an optimal solution to the linear program in Equation 14.3.
In terms of the tableau given in Equation 14.8, this means if there is no negative entry in the last row, then we have found an optimal solution. The optimal value is \(Z\) the negative of the entry in the last row and column. To find the optimal solution, first discover the columns \(\{j 1, j 2, \ldots, j m\}\) that contain the identity matrix on the first \(m\) rows of the tableau. That is, column \(j i\) is
\[
[0,0, \ldots, 0, \underbrace{1}_{i \text { th }}, 0,0, \ldots, 0]^{T}
\]

Then the optimal solution is \(X\) where
\[
X[j]= \begin{cases}X_{B}[i] & \text { if } j=j_{i} \\ 0 & \text { if not }\end{cases}
\]

The Phase 2 portion of the simplex algorithm is given in Algorithm 14.2.3.
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Algorithm 14.2.3: PHASE2 (Tableau m, n)
\(C \leftarrow 1\)
while \(c<n\)
        if Tableau \([m+1, c]<0\)
\(\left\{\begin{array}{l}\left\{\begin{array}{l}\text { if }(\text { Tableau }[i, c]>0) \text { and } \frac{\text { Tableau }[i, n+1]}{\text { Tableau }[i, c]}<M \\ \text { then }\left\{\begin{array}{l}M \leftarrow \frac{\text { Tableau }[i, n+1]}{\text { Tableau }[i, c]} \\ r \leftarrow i\end{array}\right. \\ i \leftarrow i+1\end{array}\right. \\ \begin{array}{l}\text { PIVOT }(r, c) \\ \text { pivots }[r] \leftarrow c \\ c \leftarrow 1\end{array} \\ c \leftarrow c+1\end{array}\right.\)
```


### 14.2.9 Phase 1: initial basis feasible solution

Suppose that after Phase $\mathbf{0}$ we have obtained a basis solution $X B$ to the linear program Minimize: $Z=c T X$
Subject to: $A X=b$ $X \geq 0$
where $A$ is a $m$ by $n$ matrix of rank $m$. Then $B=[B 1, B 2, . ., B m]$
is a nonsingular submatrix of $A$ and $X B=B-1 b$. The result of pivoting on the columns in $B$ is the tableau

$$
\left[\begin{array}{ccc|c}
\cdots & Y_{j} & \cdots & X_{B} \\
\hline \cdots & c_{j}-z_{j} & \cdots & -z
\end{array}\right] .
$$

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If $X B[i]<0$ for some row $i$, then the basis solution is infeasible and we cannot proceed to Phase 2. Let E be the m-dimensional vector defined by

$$
E[i]=\left\{\begin{array}{r}
-1, \text { if } X_{B}[i]<0 \\
0, \text { if } X_{B}[i] \geq 0
\end{array}\right.
$$

Let $x 0$ be a new artificial variable and define the Phase 1 linear program as follows:

$$
\text { Minimize: } w=[1,0,0, \ldots, 0]\left[\begin{array}{c}
x_{0}  \tag{14.17}\\
X
\end{array}\right]
$$

Subject to: $[E, Y]\left[\begin{array}{l}x_{0} \\ X\end{array}\right]=X_{B}$

$$
x_{0}, X \geq 0
$$

The Phase 1 tableau is:
$\left[\begin{array}{c|ccc|r}E & \cdots & Y_{j} & \cdots & X_{B} \\ \hline 0 & \cdots & c_{j}-z_{j} & \cdots & -Z \\ \hline 1 & 0,0, \cdots & 0 & \cdots & 0,0\end{array}\right]$.

It has columns $0,1, \ldots, n$. The second to last row corresponds to the cost equation to the linear program 14.16 and can almost be ignored for this discussion. Let $w(x 0, X)$ denote the value of the objective function for the Phase 1 linear program at a given $x 0$ and $X$.
LEMMA 14.2 The Phase 1 linear program is always feasible and has a non-negative optimal objective value.

PROOF Let $i$ be the row for which $X B[i]$ is smallest (i.e., most negative), and pivot on row $i$ column 0 . The result is a tableau with a feasible solution to the Phase 1 linear program. We can apply the techniques of Section 14.2.6 to obtain an optimal feasible solution or show that it is unbounded below. It cannot be unbounded below, however, because

$$
w=w\left(x_{0}, X\right)=[1, \overrightarrow{0}]\left[\begin{array}{c}
x_{0} \\
X
\end{array}\right]=x_{0} \geq 0
$$

Hence $w$ is non-negative.
THEOREM 14.3 A linear program is feasible if and only if the artificial objective function has minimum value $w m i n=0$. Moreover, if $w=0$ for $x 0 \geq 0$ and $X \geq 0$, then $X$ is a feasible solution of the original linear program.

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PROOF By Lemma 14.2, the Phase 1 linear program has an optimal solution for which the minimum value $w m i n \geq 0$. First suppose $w m i n=0$. Then there exist $X$ and $x 0$ such that $w(x 0, X) \geq 0, x 0=0, X \geq 0$, and

$$
[E, Y]\left[\begin{array}{c}
x_{0} \\
X
\end{array}\right]=x_{0} E+Y X=X_{B}
$$

Then $B-1 A X=Y X=B-1 b=X B$ and hence $A X=b$. Thus the linear program 14.16 is feasible.
Conversely, suppose there exists a feasible solution $\hat{X}$ of $A X=b$. Then $Y \hat{X}=B^{-1} A \hat{X}=B^{-1} b=\hat{X}_{B}$ and thus

$$
[E, Y]\left[\begin{array}{c}
x_{0} \\
X
\end{array}\right]=X_{B}
$$

is solvable in non-negative variables, by choosing $x 0=0$, and $X=\hat{X}$. For these values of $x 0$ and $X$, $w(0, \hat{X})=0$. Because $w \geq 0$ must hold for every feasible solution of the Phase $\mathbf{1}$ linear program, we have wmin $=0$.
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Algorithm 14.2.4: PHASE1 (Tableau, $m$, $n$ ) Unbounded-false Infeasibleヶfalse
comment: Use column 0 and row $m+2$ to create the Phase 1 tableau

$$
\begin{aligned}
& \text { for } i \leftarrow 1 \text { to } m \text { do }\left\{\begin{array}{l}
\text { if Tableau }[i, n+1]<0 \\
\text { then Tableau }[i, 0] \leftarrow-1 \\
\text { else Tableau }[i, 0] \leftarrow 0
\end{array}\right. \\
& \text { Tableau }[\mathrm{m}+2,0] \leftarrow 1 \\
& \text { for } j \leftarrow 1 \text { to } n \text { do Tableau }[m+2, j] \leftarrow 0 \\
& \text { Tableau }[m+2, n+1] \leftarrow 0 \\
& \text { comment: find the Phase } 1 \text { first pivot } \\
& \begin{array}{l}
r=0 \\
M \leftarrow 0
\end{array} \\
& \text { for } i \leftarrow 1 \text { to } m \text { do }\left\{\begin{array}{l}
\text { if Tableau }[i, n+1]<M \\
\text { then }\left\{\begin{array}{l}
M \leftarrow \text { Tableau }[i, n-1] \\
r \leftarrow i
\end{array}\right.
\end{array}\right. \\
& \text { if } r=0 \\
& \text { then }\left\{\begin{array}{l}
\text { comment: no phase } 1 \text { is necessary } \\
\text { return }
\end{array}\right. \\
& \text { PIVOT }(r, 0) \\
& c \leftarrow 1 \\
& \text { while } c \leq n
\end{aligned}
$$

```
(if Tableau \([m+2, c]<0\)
do \(\left\{\begin{array}{l}\text { then }\left\{\begin{array}{l}i \leftarrow 1 \\ \text { while Tableau }[i, c] \leq 0 \text { do } i \leftarrow i+1 \\ M \leftarrow \frac{\text { Tableau }[i, n+1]}{\text { Tableau }[i, c]} \\ r \leftarrow i \\ i \leftarrow i+1 \\ \text { while } i \leq m \text { do }\end{array}\right. \\ \left\{\begin{array}{l}\text { if }(\text { Tableau }[i, c]>0) \text { and } \frac{\text { Tableau }[i, n+1]}{\text { Tableau }[i, c]}<M \\ \text { then }\left\{\begin{array}{l}M \leftarrow \frac{\text { Tableau }[i, n+1]}{\text { Tableau }[i, c]} \\ r \leftarrow i+1\end{array}\right. \\ i \leftarrow i+1\end{array}\right.\end{array}\right.\)
    PIVOT \((r, c)\)
    pivots \([r] \leftarrow c\)
\(c \leftarrow c+1\)
if Tableau \([m+2, n+1] \neq 0\)
then Infeasibleヶtrue
```

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We can now give the complete simplex algorithm.
Algorithm 14.2.5: SIMPLEX (Tableau, $m$, $n$ )
Unbounded -false
Infeasible false
PHASE0 (Tableau, $m, n$ )
if Infeasible
then $\left\{\begin{array}{l}\text { output ('The Linear Program is infeasible.") } \\ \text { return }\end{array}\right.$
PHASE 1 (Tableau, $m, n$ )
if Infeasible
then $\left\{\begin{array}{l}\text { output ("The Linear Program is infeasible.") } \\ \text { return }\end{array}\right.$
PHASE2 (Tableu m, n) if Unbounded
then $\left\{\begin{array}{l}\text { output ("The Linear Program is unbounded.") } \\ \text { return }\end{array}\right.$
$Z \leftarrow-$ Tableau[ $m+1, n+1]$
for $j \leftarrow 1$ to $n$ do $X[j] \leftarrow 0$
for $\mathbf{i} \leftarrow 1$ to $m$ do $X$ [pivots $[r]]=T a b l e a u[i, n+1]$
return $(X, Z)$
14.2.10 An example

Minimize: $Z=7 \times 1+2 \times 2$
Subject to: $-x 1+2 \times 2+x 3=4$
$4 x 1+3 \times 2+x 3+x 4=24$
$-2 \times 1-2 \times 2+x 5=-7$ $x 1, x 2, x 3, x 4 \geq 0$
The initial tableau is

$$
\left[\begin{array}{rrrrr|r}
-1 & 2 & 1 & 0 & 0 & 4 \\
4 & 3 & 1 & 1 & 0 & 24 \\
-2 & -2 & 0 & 0 & 1 & -7 \\
\hline 7 & 2 & 0 & 0 & 0 & 0
\end{array}\right]
$$

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and we start Phase 0. Pivoting on the $[1,3]$ entry obtains

$$
\left[\begin{array}{rrrrr|r}
-1 & 2 & 1 & 0 & 0 & 4 \\
5 & 1 & 0 & 1 & 0 & 20 \\
-2 & -2 & 0 & 0 & 1 & -7 \\
\hline 7 & 2 & 0 & 0 & 0 & 0
\end{array}\right]
$$

Thus we have the basis solution

$$
X=[0,0,4,20,-7] T
$$

corresponding to basis consisting of columns 3,4 and 5 . This ends Phase $\mathbf{0}$. The basis solution found was not feasible so we start Phase 1. In this phase we have columns 0, 1, 2, 3, 4, and 5.
$\left[\begin{array}{r|rrrrr|r}0 & -1 & 2 & 1 & 0 & 0 & 4 \\ 0 & 5 & 1 & 0 & 1 & 0 & 20 \\ -1 & -2 & -2 & 0 & 0 & 1 & -7 \\ \hline 0 & 7 & 2 & 0 & 0 & 0 & 0 \\ \hline 1 & 0 & 0 & 0 & 0 & 0 & 0\end{array}\right]$

First we price out column 0 .
$\left[\begin{array}{r|rrrrr|r}0 & -1 & 2 & 1 & 0 & 0 & 4 \\ 0 & 5 & 1 & 0 & 1 & 0 & 20 \\ 1 & 2 & 2 & 0 & 0 & -1 & 7 \\ \hline 0 & 7 & 2 & 0 & 0 & 0 & 0 \\ \hline 0 & -2 & -2 & 0 & 0 & 1 & -7\end{array}\right]$

We now have a feasible solution, but the Phase 1 objective value is 7 . We proceed to reduce this value to 0 . In the last row we find a negative entry in column 1. The smallest ratio of the last column entries with positive entries in column 1 is $\theta=7 / 2$, so we pivot on the [ 3,1 ] entry to obtain the tableau below.

$$
\left[\begin{array}{r|rrrrr|r}
\frac{1}{2} & 0 & 3 & 1 & 0 & -\frac{1}{2} & 7 \frac{1}{2} \\
-2 \frac{1}{2} & 0 & -4 & 0 & 1 & 2 \frac{1}{2} & 2 \frac{1}{2} \\
\frac{1}{2} & 1 & 1 & 0 & 0 & -\frac{1}{2} & 3 \frac{1}{2} \\
\hline-3 \frac{1}{2} & 0 & -5 & 0 & 0 & 3 \frac{1}{2} & -24 \frac{1}{2} \\
\hline 1 & 0 & 0 & 0 & 0 & 0 & 0
\end{array}\right]
$$

The Phase 1 objective value is now zero and we have a feasible solution. We proceed to Phase $\mathbf{2}$ by dropping column 0 and the last row.
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$$
\left[\begin{array}{rrrrr|r}
0 & 3 & 1 & 0 & -\frac{1}{2} & 7 \frac{1}{2} \\
0 & -4 & 0 & 1 & 2 \frac{1}{2} & 2 \frac{1}{2} \\
1 & 1 & 0 & 0 & -\frac{1}{2} & 3 \frac{1}{2} \\
\hline 0 & -5 & 0 & 0 & 3 \frac{1}{2} & -24 \frac{1}{2}
\end{array}\right]
$$

There is a negative entry in the last row in column 2 , so the objective value can be reduced. The smallest ratio is $\theta=\left(7 \frac{1}{2}\right) / 3$, so we pivot on the [1,2]-entry.

$$
\left[\begin{array}{rrrrr|r}
0 & 1 & \frac{1}{3} & 0 & -\frac{1}{6} & 2 \frac{1}{2} \\
0 & 0 & 1 \frac{1}{3} & 1 & 1 \frac{5}{6} & 12 \frac{1}{2} \\
1 & 0 & -\frac{1}{3} & 0 & -\frac{1}{3} & 1 \\
\hline 0 & 0 & 1 \frac{2}{3} & 0 & 2 \frac{2}{3} & -12
\end{array}\right]
$$

There are no negative entries left in the last row so an optimal solution has been found. This solution is

$$
X=\left[1,2 \frac{1}{2}, 0,12 \frac{1}{2}, 0\right]^{T}
$$

and has objective value $Z=12$.

### 14.3 Cycling

It is prudent to ask whether it is possible for the simplex algorithm to go through an endless sequence of iterations without terminating. Consider the following linear program:

Minimize: $Z=-10 \times 1+57 \times 2+9 \times 3+24 \times 4$
Subject to: $0.5 \times 1-5.5 \times 2-2.5 \times 3+9 \times 4+\times 5=0$
$0.5 \times 1-1.5 \times 2-0.5 \times 3+\times 4+\times 6=0$
$x 1+x 7=1$
$x 1, x 2, x 3, x 4, x 5, x 6, x 7 \geq=0$
The initial tableau is

$$
\left[\begin{array}{rrrrrrr|r}
0.5 & -5.5 & -2.5 & 9 & 1 & 0 & 0 & 0 \\
0.5 & -1.5 & -0.5 & 1 & 0 & 1 & 0 & 0 \\
1 & 0 & 0 & 0 & 0 & 0 & 1 & 1 \\
\hline-10 & 57 & 9 & 24 & 0 & 0 & 0 & 0
\end{array}\right] .
$$

If we adopt the following rule:

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Always pivot on the column with the smallest (most negative) entry in the bottom row, choosing the first row that achieves the smallest ratio. If there are two such columns always choose the first one, (i.e., the one with smallest index).
Then the sequence of iterations is:

1. Pivot on (1, 1).

$$
\left[\begin{array}{rrrrrrr|r}
1 & -11 & -5 & 18 & 2 & 0 & 0 & 0 \\
0 & 4 & 2 & -8 & -1 & 1 & 0 & 0 \\
0 & 11 & 5 & -18 & -2 & 0 & 1 & 1 \\
\hline 0 & -53 & -41 & 204 & 20 & 0 & 0 & 0
\end{array}\right]
$$

2. Pivot on (2, 2).

$$
\left[\begin{array}{rrrrrrr|r}
1 & 0 & 0.5 & -4 & -0.75 & 2.75 & 0 & 0 \\
0 & 1 & 0.5 & -2 & -0.25 & 0.25 & 0 & 0 \\
0 & 0 & -0.5 & 4 & 0.75 & -2.75 & 1 & 1 \\
\hline 0 & 0 & -14.5 & 98 & 6.75 & 13.25 & 0 & 0
\end{array}\right]
$$

3. Pivot on (1, 3).

$$
\left[\begin{array}{rrrrrrr|r}
2 & 0 & 1 & -8 & -1.5 & 5.5 & 0 & 0 \\
-1 & 1 & 0 & 2 & 0.5 & -2.5 & 0 & 0 \\
1 & 0 & 0 & 0 & 0 & 0 & 1 & 1 \\
\hline 29 & 0 & 0 & -18 & -15 & 93 & 0 & 0
\end{array}\right]
$$

4. Pivot on $(2,4)$.

$$
\left[\begin{array}{rrrrrrr|r}
-2 & 4 & 1 & 0 & 0.5 & -4.5 & 0 & 0 \\
-0.5 & 0.5 & 0 & 1 & 0.25 & -1.25 & 0 & 0 \\
1 & 0 & 0 & 0 & 0 & 0 & 1 & 1 \\
\hline 20 & 9 & 0 & 0 & -10.5 & 70.5 & 0 & 0
\end{array}\right]
$$

5. Pivot on(1, 5).

$$
\left[\begin{array}{rrrrrrr|r}
-4 & 8 & 2 & 0 & 1 & -9 & 0 & 0 \\
0.5 & -1.5 & -0.5 & 1 & 0 & 1 & 0 & 0 \\
1 & 0 & 0 & 0 & 0 & 0 & 1 & 1 \\
\hline-22 & 93 & 21 & 0 & 0 & -24 & 0 & 0
\end{array}\right]
$$

6. Pivot on $(2,6)$.

$$
\left[\begin{array}{rrrrrrr|r}
0.5 & -5.5 & -2.5 & 9 & 1 & 0 & 0 & 0 \\
0.5 & -1.5 & -0.5 & 1 & 0 & 1 & 0 & 0 \\
1 & 0 & 0 & 0 & 0 & 0 & 1 & 1 \\
\hline-10 & 57 & 9 & 24 & 0 & 0 & 0 & 0
\end{array}\right]
$$

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The tableau obtained after iteration 6 is identical to the initial tableau and so this cycle of iterations would repeat and the simplex algorithm would never terminate. It is easy to see that this is the only way that the simplex algorithm can fail to terminate. That is
THEOREM 14.4 If the simplex algorithm fails to terminate, then it must cycle.
Several methods have been proposed to avoid this cycling phenomenon. The easiest is to adopt the smallest index rule.
Always pivot on the column with the first negative entry in the bottom row (i.e., the one with the smallest index), and choose the first row in that column that achieves the smallest ratio.
We leave it as an exercise to prove the following result:
THEOREM 14.5 (R.G. Bland, 1977) The simplex method always terminates if the smallest index rule is adopted.
PROOF Exercise 14.3.2.

## Exercises

14.3.1 Solve the following linear programs using the simplex algorithm.
(a)

Maximize: $z=3 \times 1+2 \times 2+x 3+4 \times 4$
Subject to: $4 \times 1+5 \times 2+\times 3-3 \times 4=5$, $2 \times 1-3 \times 2-4 \times 3+5 \times 4=7$, $x 1+4 \times 2+2.5 \times 3-4 \times 4=6$, $x 1, x 2, x 3, x 4 \geq 0$
(b)

# Maximize: $z=3 x 1+4 \times 2+x 3+7 \times 4$ <br> Subject to: $8 \times 1+3 \times 2+4 \times 3+x 4 \leq 5$, $2 \times 1+6 \times 2+\times 3+5 \times 4 \leq 7$, $x 1+4 \times 2+5 \times 3+2 \times 4 \leq 6$, $x 1, x 2, x 3, x 4 \geq 0$ 

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(c)

Maximize: $z=2 x 1-3 \times 2+4 \times 3+x 4$
Subject to: $x 1+5 \times 2+9 \times 3-6 \times 4 \geq-2$,
$3 \times 1-1 \times 2+\times 3+3 \times 4 \leq 10$,
$-2 \times 1-3 \times 2+7 \times 3-8 \times 4 \geq 0$,
$x 1, x 2, x 3, x 4 \geq 0$
14.3.2 Prove Theorem 14.5.

### 14.4 Notes

The topic of linear programming appears in a variety of different subjects, for example operations research, mathematical programming, and combinatorial optimization. There are thus numerous books in which it is discussed and among them are CHVÁTAL [27], HADLEY [57], NEMHAUSER and WOLSEY [92],
PAPADIMITRIOU and STEI GLITZ [94], TAHA [110], and WALKER [121].
In this chapter we have only discussed the simplex algorithm which was invented in the late 1940s by DANZIG [31] (see also DANTZG [32]). A thorough discussion of the history of linear programming can be found in DANTZIG'S celebrated work [33].

The example of cycling in the simplex method found in Section 14.3, is from the book by CHVÁTAL [27]. Theorem 14.5 appears in BLAND [16].
It can be shown that in the worst case the running time of simplex algorithm is not polynomial bounded (see KLEE and MINTY [72]) and hence the simplex algorithm is theoretically not satisfactory. In practice it is eminently useful and except for very contrived problems exceedingly fast. In 1979, KHACHIAN [71] provided a method called the ellipsoid method that solves linear programs in polynomial time. This is a marvelous, elegant, and simple jewel of pure mathematics. However, we believe that it is unlikely that the ellipsoid method will ever be a serious challenger to the simplex method.

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The Primal-Dual Algorithm

### 15.1 I ntroduction

In Chapter 14 we found it convenient to convert every linear program consisting of constraints that are a mix of inequalities and equalities:

$$
\sum a_{i j} x_{i}\{\leq=\geq\} b_{i}, \quad i=1,2, \ldots, m
$$

to a system of equations $A x=b, b \geq 0$. A slightly different formulation of constraints will prove useful here. We convert every equality to the equivalent pair of inequalities, so that

$$
\begin{aligned}
& \sum a_{i j} x_{i}=b_{i} \\
& \sum a_{i j} x_{i} \geq b_{i} \\
& \sum a_{i j} x_{i} \leq b_{i}
\end{aligned}
$$

We then multiply all inequalities with the relation $\geq$ through by a -1 so that each has the form

$$
\sum a_{i j} x_{i} \leq b_{i}
$$

Now we have a linear program of the form Maximize $z=c T X$ subject to: $D X \leq d$
$X \geq 0$
which we call the primal linear program. Corresponding to the primal linear program is another linear program which we call the dual linear program.
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is equivalent to :

$$
\begin{array}{lc}
\text { Maximize } & z^{\star}=(-d)^{T} W \\
\text { subject to: } & (-D)^{T} W \leq-c . \\
& W \geq 0
\end{array}
$$

This is in the form of a primal linear program. The dual linear program corresponding to it is:

$$
\begin{array}{cc}
\text { Minimize } & Z^{\star}=(-c)^{T} X \\
\text { subject to: } & -D X \geq-d \\
& X \geq 0
\end{array}
$$

This linear program is equivalent to

LEMMA 15.2 If $X$ is a feasible solution to the primal and $W$ is a feasible solution to the dual, then $c T X \leq d T W$ ( implying $z \leq Z$ ).
PROOF Suppose $X$ is a feasible solution to the primal. Then $D X \leq d$. If $W$ is a feasible solution to the dual, then $W \geq 0$, so we see

$$
W T D X<W T d=d T W
$$

Similarly, because $X \geq 0$ and $W$ is a feasible solution,
$D T W>C$
$X T D T W \geq X T c$
$W T D X \geq c T X$
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Therefore

$$
\begin{equation*}
c T X \leq W T D X \leq d T W . \tag{15.1}
\end{equation*}
$$

LEMMA 15.3 If $\widehat{X}$ is a feasible solution to the primal and $\widehat{W}$ is a feasible solution to the dual such that $c^{T} \widehat{X}=d^{T} \widehat{W}$, then $\widehat{X}$ and $\widehat{W}$ are optimal solutions to the primal and dual, respectively.
PROOF By assumption, $c^{T} \widehat{X}=d^{T} \widehat{W}$, but for any feasible solution $X$ to the primal,

$$
c^{T} X \leq d^{T} \widehat{W}=c^{T} \widehat{X}
$$

(with the inequality from Lemma 15.2). Therefore, $\widehat{X}$ is an optimal solution to the primal.
By the same logic, for any feasible solution $W$ of the dual,

$$
d^{T} W \geq c^{T} \widehat{X}=d^{T} \widehat{W}
$$

Thus, $\widehat{W}$ is an optimal solution to the dual.
LEMMA 15.4 If the dual or the primal has an optimal solution, then the other also must have an optimal solution and their optimal values are the same.
PROOF Because of Lemma 15.1, we need only show that if the primal linear program has an optimal solution, then so does the dual linear program. Recall that the primal linear program is:

$$
\begin{gathered}
\text { Maximize } z=c T X \\
\text { subject to: } D X \leq d \\
X \geq 0 \text {. }
\end{gathered}
$$

Add slack variables $X$ s and convert it to a minimization problem to get the standard form linear program:

$$
\begin{gathered}
\text { Minimize }(-z)=(-c) T X \\
\text { subject to: } D X+I X s=d \text {, } \\
X, X s \geq 0
\end{gathered}
$$

where $I=[E 1, E 2, \ldots, E m]$ is the $m$ by $m$ identity matrix. Let $D=[D 1, \ldots, D n]$. The tableau for the simplex algorithm is

$$
\left[\left.\begin{array}{cccccccc|c}
D_{1} & \cdots & D_{j} & \cdots & D_{n} & E_{1} & \cdots & E_{j^{\prime}} & \cdots \\
E_{m} & d \\
\hline-c_{1} \cdots & -c_{j} \cdots & \cdots & -c_{n} & 0 & \cdots & 0 & \cdots & 0
\end{array} \right\rvert\, 0\right]
$$

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If the linear program 15.2 has an optimal solution $X B$, with optimal value $(-z)=(-C B) T X B$, then there is a rank $m$ submatrix $B$ of the columns $D 1, D 2, \ldots, D n, E 1, E 2, \ldots, E m$ such that $X B=B-1 d$ and after multiplying the tableau by

$$
\left[\begin{array}{l|l}
B^{-1} & 0 \\
\hline c_{B}^{T} B^{-1} & 1
\end{array}\right]
$$

we obtain the tableau

$$
\left[\begin{array}{ccccccccc|c}
\cdots & \cdots & Y_{j} & \cdots & \cdots & \cdots & Y_{j^{\prime}} & \cdots & \cdots & X_{B} \\
\hline \cdots & \cdots & z_{j}-c_{j} & \cdots & \cdots & \cdots & z_{j^{\prime}} & \cdots & \cdots & z
\end{array}\right]
$$

where,

$$
\begin{aligned}
z & =c_{B}^{T} B^{-1} d \\
Y_{j} & =B^{-1} D_{j} \\
Y_{j^{\prime}} & =B^{-1} E_{j^{\prime}}
\end{aligned}
$$

and because of optimality,

$$
\begin{aligned}
& 0 \leq z_{j}-c_{j}=c_{B}^{T} B^{-1} D_{j}-c_{j} \\
& 0 \leq z_{j^{\prime}}=c_{B}^{T} E_{j^{\prime}}
\end{aligned}
$$

So these equations show that

$$
\begin{aligned}
c_{B}^{T} B^{-1} D-c^{T} & \geq 0 \\
c_{B}^{T} B^{-1} I & \geq 0, \text { and } \\
z & =c_{B}^{T} B^{-1} d .
\end{aligned}
$$

If we let $W^{T}=c_{B}^{T} B^{-1}$, then these equations show that $W$ satisfies

$$
\begin{gathered}
D T W \geq c \\
W \geq 0, \text { and } \\
z=d T W .
\end{gathered}
$$

Thus $W$ is a an optimal feasible solution to the dual linear program. Optimality follows from the last equation and Lemma 15.3.
Observe the similarity between Lemma 15.4 and the max-flow-min-cut Theorem (Theorem 8.4). Indeed the Ford-Fulkerson algorithm for solving the maximum network flow problem that was presented in Section 8.2 is a primal-dual algorithm. We re-examine this algorithm in Section 15.6.3.
In the proof of Lemma 15.4 we discovered how to construct an optimal solution of the dual linear program given an optimal solution to the primal. We record this useful fact as the following corollary:
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COROLLARY 15.5 If $X$ is an optimal solution to

> Maximize $z=c T X$
> subject to: $D X \leq d$
> $X \geq 0$
with basis $B$, then $W=B-T C B$ is an optimal solution to the dual linear program
Minimize $Z=d T W$
subject to: $D T W \geq c$
$W \geq 0$
THEOREM 15.6 Given a primal-dual pair, exactly one of the following can occur:
a. Both the primal and the dual have a finite optimum.
b. The primal is unbounded and the dual is infeasible.
c. The primal is infeasible but the dual is unbounded.
d. Both the primal and the dual are infeasible.

PROOF We saw in Chapter 14 that every linear program either (i) has a finite optimum, (ii) is unbounded, or (iii) is unfeasible. Thus for a primal-dual pair there are nine possibilities. Namely:

1. Both the primal and the dual have a finite optimum.
2. The primal has a finite optimum but the dual is unbounded.
3. The primal has a finite optimum but the dual is infeasible.
4. The primal is unbounded but the dual has a finite optimum.
5. Both the primal and the dual are unbounded.
6. The primal is unbounded and the dual is infeasible.
7. The primal is infeasible but the dual has a finite optimum.
8. The primal is infeasible but the dual is unbounded.
9. Both the primal and the dual are infeasible.

Lemma 15.4 shows that possibilities 2, 3, 4, and 7 cannot occur. Equation 15.1 tells us that if either the primal or the dual is unbounded, then the other cannot have a feasible solution and thus possibility 5 is eliminated. It is easy to construct examples of the remaining four possibilities, $1,6,8$, and 9.

1. A primal-dual pair in which both the primal and the dual have a finite opti-

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mum:
Maximize $z=x 1$
subject to: $x 1 \leq 1$ $x 1 \geq 0$
Minimize $Z=w 1$ subject to: $w 1 \geq 1$ $w 1 \geq 0$
6. A primal-dual pair in which the primal is unbounded and the dual is infeasible:

Maximize $Z=-x 1+2 \times 2$
subject to: $-x 1+x 2 \leq 1$ $x 1, x 2 \geq 0$
Minimize $z=w 1$
(Primal)
subect
(Dual)
6. A primal daximize $Z=-x 1+2 x 2$
subject to: $-w 1 \geq-1$
$w 1 \geq 2$
$w 1 \geq 0$
8. A primal-dual pair in which the primal is infeasible and the dual is unbounded:

Maximize $z=x 1 \quad$ (Primal)
subject to: $x 1 \leq 1$
$-x 1 \leq-2$
$x 1 \geq 0$
Mimmize $Z=w 1-2 w 2$
subject to: $w 1-w 2 \geq 1$
$w 1, w 2 \geq 0$
9. A primal-dual pair in which both the primal and the dual are infeasible:

Maximize $z=-x 1+2 x 2$
subject to: $x 1-x 2 \leq 1$

$$
-x 1+x 2 \leq-2
$$

$x 1, x 2 \geq 0$
Minimize $Z=w 1-2 w 2$ subject to: $w 1-w 2 \geq-1$

$$
\begin{gathered}
-w 1+w 2>2 \\
w 1, w 2 \geq 0
\end{gathered}
$$

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15.2 Alternate form of the primal and its dual

It is often more convent to write the primal linear program as:
Maximize $z=c T X$ subject to: $A X=b$ $x \geq 0$.
This is equivalent to

$$
\text { Maximize } z=c T X
$$

$$
\text { subject to: } A X \leq b
$$

$$
\begin{gathered}
-A X \leq-b \\
X \geq 0,
\end{gathered}
$$

and this has the dual linear program

$$
\begin{array}{cc}
\text { Minimize } & Z=\left[b^{T},-b^{T}\right]\left[\begin{array}{l}
W_{1} \\
W_{2}
\end{array}\right] \\
\text { subject to: } & {\left[A^{T},-A^{T}\right]\left[\begin{array}{l}
W_{1} \\
W_{2}
\end{array}\right] \geq c} \\
W_{1}, W_{2} \geq 0
\end{array}
$$

This dual is equivalent to:
Minimize $Z=b T(W 1-W 2)$
subject to: $A T(W 1-W 2) \geq c$
$W 1, W 2 \geq 0$.
If we let $W=W 1-W 2$, then the entries of $W$ are unrestricted in sign and the dual linear program is equivalent to

Similarly, if we take the dual linear program to be

Then its corresponding primal linear program is

A similar proof of Corollary 15.5 found in Lemma 15.4 establishes Corollary 15.7.
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COROLLARY 15.7 If W is an optimal solution to Minimize $Z=b T W$
(Dual: equality form) subject to: $A T W=c$ $W \geq 0$.
with basis $B$, then $X=B-T b B i s$ an optimal solution to the dual linar program
Maximize $z=c T X$
subject to: $A X \leq b$ $X$ unrestricted.

### 15.3 Geometric interpretation

A geometric interpretation can be give to the dual linear program.
Let $A=[A 1, A 2, \ldots, A n]$ and write the primal linear program as:
Maximize $z=c T X$
subject to: $x 1 A 1+x 2 A 2+\ldots+x n A n=b$
$x \geq 0$.
Then the dual is

| Minimize | $Z=b^{T} W$ |  |
| :--- | :--- | :--- |
| subject to: | $A_{1}^{T} W$ | $\geq$ |
|  | $A_{2}^{T} W$ | $\geq$ |
|  |  | $c_{2}$ |
|  |  | $\vdots$ |
|  | $A_{n}^{T} W$ | $\geq$ |
|  |  | $c_{n}$ |
|  | $W$ unrestricted. |  |

The vectors $A j$ in the primal linear program 15.3 are the normals to the the half-spaces that represent the constraints in the dual linear program 15.4. Furthermore the requirement vector of the primal is normal to the hyperplane $Z=b T W$ in the dual. This is easy to illustrate in two dimensions by means of an example.

### 15.3.1 Example

Given the linear program:
Maximize $z=-3 \times 1-23 \times 2-4 \times 3$
subject to: $\times 1 A 1+\times 2 A 2+\ldots+\times 5 A 5=b$
$x 1, x 2, \ldots, x 5 \geq 0$,
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where

$$
A_{1}=\left[\begin{array}{r}
2 \\
-1
\end{array}\right], A_{2}=\left[\begin{array}{l}
-3 \\
-4
\end{array}\right], A_{3}=\left[\begin{array}{r}
-2 \\
3
\end{array}\right], A_{4}=\left[\begin{array}{l}
1 \\
0
\end{array}\right], A_{5}=\left[\begin{array}{l}
0 \\
1
\end{array}\right], \quad \text { and } b=\left[\begin{array}{r}
-2 \\
1
\end{array}\right] .
$$

The dual linear program is:

$$
\begin{gathered}
\text { Minimize } Z=-2 w 1+w 2 \\
\text { subject to: } 2 w 1-1 w 2>-3 \\
-3 w 1-4 w 2 \geq-23 \\
-2 w 1+3 w 2>-4 \\
w 1 \geq 0 \\
w 2 \geq 0
\end{gathered}
$$

(Note that the appearance of slack variables in the primal linear program have caused the variables in the
dual to be non-negative.) In Figure 15.1, we have drawn the requirement-space configuration of the primal and in Figure 15.2, the convex set of feasible solutions is shown as a shaded region.
Whenever two of the constraints hold as strict equalities, the vectors normal to these constraints are a basis for the primal (if the normals are linearly independent). In w1 $w 2$-space the point $w$ where two dual constraints hold as strict equalities is the intersection of the two lines representing these two constraints. A basis solution to the primal can then be associated with the intersection of each pair of bounding lines for the half-spaces representing the dual constraints.
There are $\binom{5}{2}=10$ basis solutions to the primal. They are represented by the points $\operatorname{Pij}, 1 \leq i<j \leq 5$, in Figure 15.2. The point Pij corresponds to having $A i, A j$ in the primal basis. In Table 15.1 we display the simplex tableaux corresponding to the various choice of basis $\{A i, A j\}$. Two of them yield feasible solutions to the primal, but only one corresponds to a point that is also feasible in the dual. This is basis $\{A 2, A 3\}$, corresponding to the point $P 23=(5,2)$. Furthermore this basis yields the value $z=-8$, obtained by setting $x 1=x 4=x 5=0, x 2=0.24$ and $x 3=0.65$. This yields the value $Z=[-2,1] T[5,2]=-8$. Thus by Lemma 15.3 , this an optimal point. Therefore using the dual simplex method we move from one extreme point of the convex polyhedron to an adjacent one until an optimal extreme point is reached. At this point, the corresponding solution to the primal becomes feasible.

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FI GURE 15.1
Requirement-space configuration for the primal


FI GURE 15.2
The convex set of feasible solutions to the dual

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TABLE 15.1
Table of simplex tableaux
$\left[\begin{array}{rrrrr|r}1 & 0 & -1.55 & 0.36 & -0.27 & -1 \\ 0 & 1 & -0.36 & -0.09 & -0.18 & 0 \\ \hline 0 & 0 & -17 & -1 & -5 & -3\end{array}\right]$
$\left[\begin{array}{rrrrr|r}1 & -4.25 & 0 & 0.75 & 0.50 & -1 \\ 0 & -2.75 & 1 & 0.25 & 0.50 & 0 \\ \hline 0 & -46.75 & 0 & 3.25 & 3.50 & -3\end{array}\right]$
$\left[\begin{array}{rrrrr|r}1 & 4 & -3 & 0 & -1 & -1 \\ 0 & -11 & 4 & 1 & 2 & 0 \\ \hline 0 & -11 & -13 & 0 & -3 & -3\end{array}\right]$
$\left[\begin{array}{rrrrr|r}1 & -1.50 & -1 & 0.50 & 0 & -1 \\ 0 & -5.50 & 2 & 0.50 & 1 & 0 \\ \hline 0 & -27.50 & -7 & 1.50 & 0 & -3\end{array}\right]$
$\left[\begin{array}{rrrrr|r}-0.24 & 1 & 0 & -0.18 & -0.12 & 0.24 \\ -0.65 & 0 & 1 & -0.24 & 0.18 & 0.65 \\ \hline-11 & 0 & 0 & -5 & -2 & 8\end{array}\right]$

Basis $\{\mathbf{A} \mathbf{1}, \mathbf{A} \mathbf{2}\}$ is infeasible.

Basis $\{A \mathbf{1}, A \mathbf{3}\}$ is infeasible.

Basis $\{A \mathbf{1}, \mathbf{A} \mathbf{4}\}$ is infeasible.

Basis $\{A \mathbf{1}, \mathbf{A 5}\}$ is infeasible.

Basis $\{A \mathbf{2}, A \mathbf{3}\}$ yields $z=\mathbf{8}$.
$\left[\begin{array}{rrrrr|r}0.25 & 1 & -0.75 & 0 & -0.25 & -0.25 \\ 2.75 & 0 & -4.25 & 1 & -0.75 & -2.75 \\ \hline 2.75 & 0 & -21.25 & 0 & -5.75 & -5.75\end{array}\right]$
$\left[\begin{array}{rrrrrrr}-0.67 & 1 & 0.67 & -0.33 & 0 & 0.67 \\ -3.67 & 0 & 5.67 & -1.33 & 1 & 3.67 \\ \hline-18.33 & 0 & 11.33 & -7.67 & 0 & 15.33\end{array}\right]$
$\left[\begin{array}{rrrrrr|r}-0.33 & -1.33 & 1 & 0 & 0.33 & 0.33 \\ 1.33 & -5.67 & 0 & 1 & 0.67 & -1.33 \\ \hline-4.33 & -28.33 & 0 & 0 & 1.33 & 1.33\end{array}\right]$
$\left[\begin{array}{rrrrrr|r}-1 & 1.50 & 1 & -0.50 & 0 & 1 \\ 2 & -8.50 & 0 & 1.50 & 1 & -2 \\ \hline-7 & -17 & 0 & -2 & 0 & 4\end{array}\right]$
$\left[\begin{array}{rrrrrr}2 & -3 & -2 & 1 & 0 & -2 \\ -1 & -4 & 3 & 0 & 1 & 1 \\ \hline-3 & -23 & -4 & 0 & 0 & 0\end{array}\right]$

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### 15.4 Complementary slackness

There is a battle of balance between the primal and dual linear programs.
As the constraints tighten in one, they loosen in the other.
In this section we denote by $[\operatorname{Rowi}(D)]$ and $[\operatorname{Colj}(D)]$ the $i$ th row and $j$ th column of the matrix $D$, respectively.
THEOREM 15.8 (Complementary slackness conditions) A primal-dual feasible solution pair $X, W$ is
optimal if and only if

$$
\begin{align*}
& x j([\operatorname{Colj}(D)] T W-c j)=0 \text { for all } j  \tag{15.5}\\
& \text { wi }(\operatorname{di}-[\operatorname{Rowi}(D)] T X)=0 \text { for all } i \tag{15.6}
\end{align*}
$$

PROOF Let

$$
\begin{gathered}
u j=x j([\operatorname{Colj}(D)] T W-c j), \text { and } \\
u i=w i(d i-[R o w i(D)] T X) .
\end{gathered}
$$

Then, because of the feasibility and the duality relations we have $u j \geq 0$ for all $j$ and $u i \geq 0$ for all $i$. Let

$$
\begin{aligned}
& u=\sum_{j} u_{j} \\
& v=\sum_{i} v_{i}
\end{aligned}
$$

Then $u, u \geq 0, u=0$ if and only if Equation 15.5 holds for all $j$ and $u=0$ if and only if Equation 15.6 holds for all $j$. Observe that

$$
\begin{aligned}
u+v & =\sum_{j} u_{j}+\sum_{i} v_{i} \\
& =\sum_{j} x_{j}\left(\left[\operatorname{Col}_{j}(D)\right]^{T} W-c_{j}\right)+\sum_{i} w_{i}\left(d_{i}-\left[\operatorname{Row}_{i}(D)\right]^{T} X\right) \\
& =-\sum_{j} x_{j} c_{j}+\sum_{i} w_{i} d_{i}+\sum_{j} x_{j}\left(\left[\operatorname{Col}_{j}(D)\right]^{T} W-\sum_{i} w_{i}\left[\operatorname{Row}_{i}(D)\right]^{T} X\right. \\
& =-c^{T} X+d^{T} W+\sum_{j} x_{j} \sum_{i} D[i, j] w_{i}-\sum_{i} w_{i} \sum_{j} D[i, j] x_{j} \\
& =-c^{T} X+d^{T} W+\sum_{i} \sum_{j} w_{i} D[i, j] x_{j}-\sum_{i} \sum_{j} w_{i} D[i, j] x_{j} \\
& =-c^{T} X+d^{T} W
\end{aligned}
$$

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Therefore Equations 15.5 and 15.6 hold for all $j$ and $i$, respectively, if and only if $u+u=0$ if and only if $c T X=d T W$ if and only if $X$ and $W$ are both optimal.
Note that Theorem 15.8 says at optimality if a constraint is not met with equality, i.e. has slack, in the primal linear program, then the corresponding variable in the dual is zero and vice versa.

### 15.5 The dual of the shortest-path problem

In this section we study the shortest-path problem for directed graphs.
Problem 15.1: Shortest Path (directed graph)
I nstance: a directed graph $G=(V, E)$, nodes $s, t \in V$ and non-negative weights $c j$, for each edge $e_{j} \in E$.
Find: a directed path $P$ from $s$ to $t$ with minimum total weight

$$
C(P)=\sum_{e_{j} \in E(P)} c_{j}
$$

Let $V=\{u 1, u 2, \ldots, u n\}$ and $E=\{e 1, e 2, \ldots, e m\}$ define the $m$ by $n$ node-edge incidence matrix $A$ by

$$
A[i, j]=\left\{\begin{aligned}
+1, & \text { if } e_{j}=\left(v_{i}, u\right) \text { for some vertex } u \\
-1, & \text { if } e_{j}=\left(u, v_{i}\right) \text { for some vertex } u \\
0, & \text { otherwise }
\end{aligned}\right.
$$

In Figure 15.3 an example is given.
We model the shortest-path problem as a network flow problem by assigning a capcity of 1 on each edge. Let $w j$ denote the flow on edge $j$.
The conservation of flow constraints are

$$
\left[\operatorname{Row}_{i}(A)\right] \cdot W=\left\{\begin{array}{c}
0, i \notin\{s, t\} \\
+1, i=s \\
-1, i=t
\end{array}\right.
$$

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|  | $(s, a)$ | $(s, b)$ | $(s, c)$ | $(a, t)$ | $(b, a)$ | $(b, d)$ | $(b, t)$ | $(c, b)$ | $(c, d)$ | $(d, t)$ |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| $s$ | +1 | +1 | +1 | 0 | 0 | 0 | 0 | 0 | 0 | 0 |
| $a$ | -1 | 0 | 0 | +1 | -1 | 0 | 0 | 0 | 0 | 0 |
| $b$ | 0 | -1 | 0 | 0 | +1 | +1 | +1 | -1 | 0 | 0 |
| $c$ | 0 | 0 | -1 | 0 | 0 | 0 | 0 | +1 | +1 | 0 |
| $d$ | 0 | 0 | 0 | 0 | 0 | -1 | 0 | 0 | -1 | +1 |
| $t$ | 0 | 0 | 0 | -1 | 0 | 0 | -1 | 0 | 0 | -1 |

$c^{T}=[8, \quad 4, \quad 3, \quad 2, \quad 3, \quad 2, \quad 6, \quad 3, \quad 3, \quad 4]$

## FI GURE 15.3

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Then W satisfies

$$
A W=\left[\begin{array}{r}
+1 \\
-1 \\
0 \\
\vdots \\
0
\end{array}\right] \begin{aligned}
& \leftarrow \text { row } s \\
& \leftarrow \text { row } t
\end{aligned}
$$

and the capacity constrains are

$$
0 \leq w j \leq 1
$$

A solution to the shortest-path problem is given by a flow $W$ that minimizes $Z=c T W$. As far as we know it is possible that the wj in general take on non-integer values, but in Section 16.4 we will show that there is an optimal solution $W$ to linear program 15.7 with only integer entries.

Minimize $\quad Z=c^{T} W$
subject to: $A W=\left[\begin{array}{r}+1 \\ -1 \\ 0 \\ \vdots \\ 0\end{array}\right] \leftarrow$ row $s$
$W \geq 0$,
Indeed it not to hard to see that there is an optimal solution to linear program 15.7 in which each wi is either zero or one. A one represents a unit flow along a shortest path from $s$ to $t$.
The dual linear program for the shortest-path problem is
Maximize $\quad z=x_{s}-x_{t}$
subject to: $\quad x_{i}-x_{j} \leq c_{i j} \quad$ for each $(i, j) \in E$
$X$ unrestricted.
The complementary slackness conditions (Theorem 15.8) are easy to interpret in the shortest-path problem. A path $W$ and an assignment of variables $X$ are jointly optimal if and only if

1. Each edge in the shortest path (i.e., a positive wi in the primal linear program 15.7) corresponds to equality in the corresponding constraint in the dual linear program 15.8, and
2. Each strict inequality in the dual linear program 15.8 corresponds to an edge not in the shortest path.
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For the graph in Figure 15.3 an optimal solution to the primal is
$w(s, a)=0, w(s, b)=1, w(s, c)=0, w(a, t)=1, w(b, a)=1$,
$w(b, d)=0, w(b, t)=0, w(c, b)=0, w(c, d)=0, w(d, t)=0$,
which has cost $=9$. In the dual we see by complementary slackness that this means

$$
\begin{gathered}
x S-x b=4 \\
x b-x a=3 \\
x a-x t=2
\end{gathered}
$$

Summing these equations we obtain $z=x s-x t=9$.

## Exercises

15.5.1 An enginer takes measurements of a variable $y(x)$; the results are in the form of pairs (xi,yi). The engineer wishes to find the straight line that fits this data best in the sense that the maximum vertical distance between any point (xi,yi) and the line is as small as possible. Formulate this as a linear program. Why might you decide to solve the dual?
15.5.2 Consider the node-edge incidence matrix $A$ of the directed graph $G=(V, E)$ as described in Section 15.5. Show that a set of | $V-1$ columns is linearly independent if and only if the corresponding edges, when considered as undirected edges, form a tree. (Thus a basis corresponds to a tree.) If a network problem is
formulated with this graph, what does this reult say about the pivot step?
15.5.3 It was shown in Section 15.2 that the dual of

Maximize $z=c T X$
subject to: $A X=b$
$X \geq 0$.
has unrestricted variables. However, if some slack and/or surplus variables appear in $A$, show that the dual variable for a constraint having a slack variable is non-negative and the dual variable for a constraint having a surplus variable is non-positive. Hence, show that the only dual variables which are really unrestricted are those that correspond to constraints that were originally equations and not inequalities. Thus show that the dual of any linear program is esentially unique and is independent of the particular manner in which we write the primal.
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15.6 The primal-dual algorithm

Consider a linear program in standard form

| Minimize | $Z=c^{T} X$ |
| :--- | :---: |
| subject to: | $A X=b$ |
|  | $X \geq 0$ |

and its dual

| Maximize | $z=b^{T} W$ |  |
| :--- | :--- | :--- |
| subject to: | $A_{1}^{T} W$ | $\leq c_{1}$ |
|  | $A_{2}^{T} W$ | $\leq c_{2}$ |
|  |  | $\vdots$ |
|  | $A_{n}^{T} W$ | $\leq c_{n}$ |
|  | $W$ unrestricted. |  |

We may assume that $b>0$, because the equalities in ( $P$ ) can be multiplied by -1 where necessary. The complementary slackness conditions (Theorem 15.8) are: if $X$ is a feasible solution to ( P ) and $W$ is a feasible solution to (D), then $X$ and $W$ are both optimal if and only if

$$
\begin{equation*}
\text { wi }([\operatorname{Rowi}(A)] T X-b i)=0 \text { for all } i \tag{15.9}
\end{equation*}
$$

and

$$
\begin{equation*}
\left(c_{j}-A_{j}^{T} W\right) x_{j}=0 \text { for all } j . \tag{15.10}
\end{equation*}
$$

Condition 15.9 is automatically satisfied because of the equality in (P), so we will focus on condition 15.10. The main idea of the primal dual algorithm is:
Given a feasible solution W to (D), find a feasible solution $X$ to $(P)$ such that

$$
X_{j}=0 \text {, whenever } A_{j}^{T} W<c_{j} .
$$

In order to construct such a pair $W, X$ we will iteratively improve $W$, while maintaining its feasibility in (D). Suppose $W$ is a feasible solution to (D). Then with respect to $W$ some of the inequalities $A_{j}^{T} W \leq c_{j}$ in (D) still have slack and some do not. Let

$$
J=\left\{j: A_{j}^{T} W=c_{j}\right\}=\left\{j_{1}, j_{2}, \ldots, j_{n^{\prime}}\right\},
$$

be the set of admissible columns. So $X$, a feasible solution to ( P ), is optimal if $x j=0$ for all $j \notin J$. Let

$$
\begin{gathered}
A_{J}=\left[A_{j_{1}}, A_{j_{2}}, \ldots, A_{j_{n^{\prime}}}\right] \\
\text { page_434 }
\end{gathered}
$$

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and

$$
X_{J}=\left[x_{j_{1}}, x_{j_{2}}, \ldots, x_{j_{n_{n}}}\right] .
$$

If we can find $X J$ such that

$$
\begin{gather*}
A J X J=b  \tag{15.11}\\
X J=0,
\end{gather*}
$$

then by complementary slackness, the $X$ defined by

$$
X[j]= \begin{cases}0 & \text { if } j \notin J \\ X_{j_{\ell}} & \text { if } j=j_{\ell} \in J\end{cases}
$$

is optimal in ( P ). To find $X I$ we construct a new linear program called the restricted primal (RP).
Minimize

$$
\begin{array}{cc}
\text { Minimize } & \zeta=\gamma\left[\begin{array}{c}
X_{J} \\
Y
\end{array}\right]=y_{1}+y_{2}+\cdots+y_{m} \\
\text { subject to: } & A_{J} X_{J}+Y=\left[A_{J}, I\right]\left[\begin{array}{c}
X_{J} \\
Y
\end{array}\right]=b \\
& X_{J} \geq 0 \\
Y \geq 0 \tag{RP}
\end{array}
$$

where $Y=[y 1, y 2, . ., y m] T$ are new variables one for each equation in $(P)$ and

$$
\gamma=[\underbrace{0,0, \ldots, 0}_{|J|}, \underbrace{1,1, \ldots, 1}_{m}]
$$

Let $\zeta$ OPT be the optimal value of (RP) and suppose it occurs at $Y=Y O P T$, with basis $B$. If $\zeta O P T=0$, then YOPT=0 and the corresponding $X J$ solves the constraints 15.11 . Thus we have an optimal solution to (P). What happens when $\zeta O P T>0$ ?
The dual of (RP) is

| Maximize | $z=b^{T} W$ |
| :---: | :---: |
| subject to: | $A_{J}^{T} W$ |
|  | $W \leq \frac{0}{1}$ |
|  | $W$ unrestricted. |

where $\overrightarrow{1}=[1,1,1, \ldots, 1]^{T}$. Let WOPT be the solution to (DRP) corresponding to YOPT, that is WOPT $=B-T y B$ (see Theorem 15.7). We call (DRP) the dual-restricted primal. The situation is that we tried to find a feasible $X$ in (P) using only the columns in J, but failed. However, we do have the optimal feasible solution pair YOPT, WOPTto (RP) and (DRP), respectively. We also know 弓OPT>0,
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Page 436
the value of (RP) at YOPT is positive. Let's try correcting $W$ by a linear combination of the old $W$ and WOPT. Let

$$
W_{\star}=W+\theta W_{\text {orr }} .
$$

The value of the objective function at $W_{\star}$ is

$$
b^{T} W_{\star}=b^{T} W+\theta b^{T} W_{\text {or }} .
$$

Now we know $b T W O P T=\zeta O P T>0$, because YOPT, WOPT is a primal-dual feasible solution pair. Thus to maximize the value of the objective function we can take $\theta>0$ and large. We also need to maintain feasibility so we need

Consider the $j$-th equation

$$
\begin{aligned}
& A^{T} W_{\star}=A^{T} W+\theta A^{T} W_{\text {ort }} \leq c \\
& A_{j}^{T} W_{\star}=A_{j}^{T} W+\theta A_{j}^{T} W_{\text {or }} \leq c_{j}
\end{aligned}
$$

If $A_{j}^{T} W_{\text {orr }} \leq 0$, then this equation is satisfied because $W$ is feasible solution to (D). Thus, in particular, $W_{\star}$ is feasible, if $A_{j}^{T} W_{\text {ort }} \leq 0$ for all $j$, but in this case we may take $\theta>0$ arbitrarily large and the value of the objective function at $W_{\star}$ will be unbounded. Therefore the primal (P) is infeasible, by Theorem 15.6. Hence: THEOREM 15.9 If 弓OPT $>0$ in (RP) and WOPT the optimal solution to (DRP) satisfies $A_{j}^{T} W_{\text {orr }} \leq 0$, for all $j \notin J$
then $(P)$ is infeasible.
Therefore we need only worry when

$$
A_{j}^{T}>0 \text { for some } j \notin J .
$$

Consequently, the feasibility conditions are

$$
A_{j}^{T} W_{\star}=A_{j}^{T} W+\theta A_{j}^{T} W_{\text {orr }} \leq c_{j}, \text { whenever } j \notin J \text { and } A_{j}^{T} W_{\text {orr }}>0 .
$$

Thus for all $j \notin J$ and $A_{j}^{T} W_{\text {ort }}>0$ we need to chose $\theta$ such that

$$
\theta \leq \frac{c_{j}-A_{j}^{T} W}{A_{j}^{T} W_{\mathrm{or}}}
$$

This yields the following theorem.
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THEOREM 15.10 When そOPT>0 in (RP) and there is a $j \notin J$ with $A_{j}^{T} W_{\text {orr }}>0$, the largest $\theta$ that maintains feasibility $W_{\star}=W+\theta W_{\text {ort }}$ is

$$
\begin{equation*}
\theta_{\star}=\min \left\{\frac{c_{j}-A_{j}^{T} W}{A_{j}^{T} W_{\text {or }}}: j \notin J \text { and } A_{j}^{T} W_{\text {ort }}>0\right\} \tag{15.12}
\end{equation*}
$$

Given a feasible solution $W$ to (D) Algorithm 15.6.1 constructs in $W$ an optimal solution to (D) or determines if $(P)$ is infeasible.

Algorithm 15.6.1: PRIMAL-DUAL(W)
FEASIBLE $\leftarrow$ true
OPTIMAL $\leftarrow$ false
while FEASIBLE and not OPTIMAL
$\left\{\begin{array}{l}J \leftarrow\left\{j: A_{j}^{T} W=c_{j}\right\} \\ \text { if }|J|=n \\ \text { then OptimaL } \leftarrow \text { true }\end{array}\right.$
$\left\{\begin{array}{l}\text { Solve (RP) by Simplex Algorithm } \\ \text { obtain solution with basis } B \text { and objective value } \zeta_{\text {orr }} \\ \text { if } \zeta_{\text {orr }}=0\end{array}\right.$
else
then Optimal $\leftarrow$ true
else

$$
\left\{\begin{array}{l}
W_{\text {or }} \leftarrow B^{-T} \gamma_{B} \\
\text { if } A_{J}^{T} W_{\text {orr }} \leq 0 \text { for all } j \in J \\
\text { then FEASIBLE } \leftarrow \text { false } \\
\text { else }\left\{\begin{array}{l}
\text { compute } \theta_{\star} \text { using equation } 15.12 \\
W \leftarrow W+\theta_{\star} W_{\text {orr }}
\end{array}\right.
\end{array}\right.
$$

### 15.6.1 I nitial feasible solution

In order to start Algorithm 15.6.1 we must first have a feasible solution $W$ to (D). If $c i \geq 0$ for all $i$, we can take $W=\overrightarrow{0}$ as an initial feasible solution. When $c i<0$ for some $i$, we can use the following method to obtain a feasible solution $W$ to (D). Introduce a new variable $x 0$ to the primal problem (P), set $\mathrm{CO}=0$ and add the constraint:

$$
x 0+x 1+\ldots+x n=b 0,
$$

where $b 0$ is taken to be larger than $\sum_{i}^{n} x_{i}$, for every basis solution $X=[x 1, \ldots, x n]$ to (P). For example, take $b 0=n M$ where $M$ is given in Lemma 15.11. This new primal is

Page 438
Minimize
subject to:

$$
\begin{gather*}
Z=c^{T} X=\left[0, c^{T}\right]\left[\begin{array}{c}
x_{0} \\
X
\end{array}\right] \\
x_{0}+x_{1}+\cdots+x_{n}=b_{0} \\
A X=b \\
x_{0} \geq 0, X \geq 0
\end{gather*}
$$

and its dual is

$$
\begin{array}{lc}
\text { Maximize } & z=w_{0} b_{0}+b^{T} W \\
\text { subject to: } & w_{0} \leq 0 \\
& w_{0}+A_{1}^{T} W \\
& w_{0}+A_{2}^{T} W
\end{array} \leq c_{1} \leq c_{2} .
$$

where $w 0$ is the new variable corresponding to the new equation in the new primal. A feasible solution to this new dual is

$$
W_{i}=\left\{\begin{array}{cl}
\operatorname{MIN}\left\{c_{1}, c_{2}, \ldots, c_{n}\right\}, & \text { if } i=0 \\
0, & \text { if } i=1,2, \ldots, m
\end{array}\right.
$$

Also $[x 0, X]$ is an optimal solution to the new primal ( $\mathrm{P}^{\prime}$ ) if and only if $X$ is an optimal solution to (P). A suitable value for 60 is provided by the following lemma.
LEMMA 15.11 Let $X=[x 1, x 2, \ldots, x n]$ be a basis solution to (P). Then
where
$|x j| \leq M=m!a m-1 \beta$
and

$$
a=\operatorname{MAXi}, j\{|a i, j|\}
$$

$\beta=\operatorname{MAX}\{b 1, b 2, \ldots, b m\}$.
PROOF Without loss of generality we assume that the entries of $A, b$, and $c$ are integers. If $X$ is a basis solution, then there is a set $J$ of $m$ columns such that $X[j]=0$ whenever $j \notin J$, and $B=\left[A_{j}: j \in J\right]$ is nonsingular. Thus $B X]=b$, where $X_{J}=[X[j]: j \in J]$ and so by Cramer's rule

$$
x_{j}=\frac{\operatorname{det}\left(B^{\prime}\right)}{\operatorname{det}(B)},
$$

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where $B^{\prime}$ is the matrix $B$ with column $j$ replaced with $b$. By integrality we have $|\operatorname{det}(B)| \geq 1$, and so

$$
\begin{aligned}
\left|x_{j}\right| & \leq\left|\operatorname{det}\left(B^{\prime}\right)\right| \\
& =\sum_{\sigma \in S_{m}} \operatorname{SIGN}(\sigma) \prod_{h=1}^{m} B^{\prime}[h, \sigma(h)] \\
& \leq m!\alpha^{m-1} \beta,
\end{aligned}
$$

where $S m$ is the set of permutations of $\{1,2, \ldots, m\}$ and

$$
\operatorname{SIGN}(\sigma)= \begin{cases}+1, & \text { if } \sigma \text { is an even permutation; } \\ -1, & \text { if } \sigma \text { is an odd permutation }\end{cases}
$$

We now proceed to show that Algorithm 15.6 .1 solves ( P ) in finitely many iterations.
THEOREM 15.12 Every admissible column in the optimal basis of (RP) remains admissible at the start of the next iteration.
PROOF Suppose that the optimal basis $B$ of (RP) includes the column $A j$ Then by Corollary 15.7 the optimal solution to (DRP) corresponding to $X$ is WOPT $=B-T \nu B$, where $\gamma B$ are the entries of

$$
\gamma=[\underbrace{0,0, \ldots, 0}_{|J|}, \underbrace{1,1, \ldots, 1}_{m}]
$$

corresponding to the columns of $[A J, I]$ that are in $B$. Consequently, for some $\ell, 1 \leq \ell \leq|J|$,

$$
\begin{aligned}
A_{j}^{T} W_{\text {ort }} & =A_{j}^{T}\left(B^{-T} \gamma_{B}\right) \\
& =\left(A_{j}^{T} B^{-T}\right) \gamma_{B} \\
& =\left(B^{-1} A_{J}\right)^{T} \gamma_{B} \\
& =E_{\ell}^{T} \gamma_{B} \\
& =\gamma_{B}[\ell]=0,
\end{aligned}
$$

where $E \ell=[0,0, \ldots, 0,1,0,0, \ldots, 0] T$ with 1 in position $\ell$. This implies that
$A_{j}^{T} W_{\star}=A_{j}^{T}\left(W+\theta W_{\text {or }}\right)$
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$$
\begin{aligned}
& =A_{j}^{T} W+\theta W_{\text {ort }}^{T} A_{j} \\
& =\gamma_{j}+0 \\
& =\gamma_{j}
\end{aligned}
$$

Thus if $j \in J$ at the start of an iteration of the while loop in Algorithm 15.6.1, then $j$ remains in $J$ in the next iteration.
How do we know that the primal-dual algorithm will terminate? If the minimum calculation of $\theta_{\star}$ by Equation 15.12 occurs at $j=j_{\star}$, then

$$
\theta_{\star}=\frac{c_{j_{\star}}-A_{j_{\star}}^{T} W}{A_{j_{\star}}^{T} W_{\text {ort }}}
$$

and so

$$
c_{j_{\star}}=A_{j_{\star}}^{T} W+\theta_{\star} A_{j_{\star}}^{T} W_{\text {orr }}=A_{j_{\star}}^{T} W_{\star} .
$$

Consequently, $j_{\star}$ becomes a new column of $J$, and we see that $|J|$ monotonically increases. If $J=\{1,2, \ldots, n\}$, then $W$ satisfies the complementary slackness condition in Equation 15.10 and is therefore optimal. Consequently we have the following theorem.
THEOREM 15.13 Algorithm 15.6.1 correctly solves (P) in at most $n$ iterations.

### 15.6.2 The shortest-path problem

We illustrate the primal-dual method with the shortest-path problem introduced in Section 15.5. Observe that the row sum of the coefficient matrix in linear program 15.8 is zero. Hence any row may be omitted because it is redundant. If we choose to omit row t , the resulting linear program 15.13 is our primal.

$$
\text { Minimize } \quad Z=c^{T} X
$$

$$
\text { subject to: } \quad A X=\left[\begin{array}{r}
+1 \\
0 \\
\vdots \\
0
\end{array}\right] \leftarrow \text { row } s
$$

$$
\begin{equation*}
X \geq 0 \tag{15.13}
\end{equation*}
$$

The vector of flow is $X$ and each entry of $X$ is 0 or 1 . The dual linear program for the shortest-path problem is Maximize $z=w_{s}$ subject to: $\quad w_{i}-w_{j} \leq c_{i j}$ for each edge $(i, j) \in E$ $W$ unrestricted, except $w_{t}=0$.
page_440
Page 441
We fix $w t=0$, because its row was omitted from the primal. The set of admissible columns (i.e., edges) is

$$
J=\{(i, j): x i x j=c i j\}
$$

and the restricted primal is

$$
\text { Minimize } \quad \zeta=y_{1}+y_{2}+\cdots+y_{m-1}
$$

$$
\text { subject to: } A X+Y=\left[\begin{array}{r}
+1 \\
0 \\
\vdots \\
0
\end{array}\right] \leftarrow \text { row } s
$$

$$
\begin{align*}
& x_{j} \geq 0 \text { for all } j \in J \\
& x_{j}=0 \text { for all } j \notin J \\
& Y \geq 0 . \tag{15.15}
\end{align*}
$$

Consequently, the dual-restricted primal is

$$
\begin{array}{ll}
\text { Maximize } & z=w_{s} \\
\text { subject to: } & w_{i}-w_{j} \leq c_{i j} \text { for each edge }(i, j) \in J \\
& W \leq 1 \\
& W \text { unrestricted. } \tag{15.16}
\end{array}
$$

It is easy to solve the dual-restricted primal. Note that $w s \leq 1$, and that we are maximizing $w s$, so we can try $w s=1$. If $w i=1$ and $(i, j) \in J$, then we can satisfy the constraint

$$
w i-w j \leq 0,
$$

by also setting $w j=1$. Hence if $P=i 1, i 2 \ldots i h$ is a path with $\left(i_{a}, i_{a+1}\right) \in J$ for all $a=1,2, \ldots, h-1$ and $w_{i_{1}}=1$. Then we can set $w_{i_{a}}=1$ for each a without violating the constraints. Hence if there is no st-path using edges only in J, then an optimal solution to the dual-restricted primal is qiven by

$$
W_{\text {ort }}[i]=\left\{\begin{array}{ll}
1, & \text { if there is an si-path using only edges in } J \\
0, & \text { if there is an it-path using only edges in } J \\
1, & \text { for all other vertices } i
\end{array} .\right.
$$

We then calculate

$$
\theta_{\star}=\operatorname{MIN}\left\{c_{i j}-\left(w_{i}-w_{j}\right):(i, j) \notin J \text { and } W_{\text {orr }}[i]-W_{\text {orr }}[j]>0\right\}
$$

and update $W$ and $J$, and obtain and solve the new dual-restricted primal. If we get to a point where there is an st-path using edges in J, then $w s=0$ and we are

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## FI GURE 15.4

Shortest-path example
at an optimal solution, because minimum $\zeta$ is equal to the maximum $z$ which is ws $=0$. Notice that any st-path that uses only edges in $J$ is optimal.
The primal-dual algorithm reduces the shortest-path problem to the easy calculation which vertices are reachable from the source $s$.
We illustrate the primal-dual algorithm when applied to the shortest-path problem with the directed graph given in Figure 15.4.

## Iteration 1


(D): $W=[0,0,0,0,0,0,0,0]$ $J=\{ \}$

## Iteration 2


(D): $W=[1,1,1,1,1,1,1,0]$
$J=\{(4,7)\}$
(s) $\quad \stackrel{1}{2}$
(3)
$(7)$
1

## 1 (4) + $(t)$ 0

(DP): $W_{\text {ort }}=[1,1,1,1,1,1,1,0]$ $\theta^{\star}=1$ for edge $(4,7)$
(S)

(DP): $W_{\text {ort }}=[1,1,1,0,1,1,1,0]$ $\theta^{\star}=2$ for edges $(3,4)$ and $(7, t)$
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## Iteration 3


(D): $W=[3,3,3,1,3,3,3,0]$
$J=\{(4, t),(3,4),(7, t)\}$

## Iteration 4


(D): $W=[4,4,1,1,4,4,1,0]$
$J=\{(4, t),(3,4),(7, t),(2,3)\}$

## Iteration 5


(D): $W=[5,4,1,1,5,5,1,0]$
$J=\left\{\begin{array}{l}(4, t),(3,4),(7, t),(2,3), \\ (6,2)\end{array}\right\}$

$(\mathrm{DRP}): W_{\text {ort }}=[1,1,0,0,1,1,0,0]$ $\theta^{\star}=1$ for edge $(2,3)$

(DP): $W_{\text {orr }}=[1,0,0,0,1,1,0,0]$
$\theta^{\star}=1$ for edge $(6,2)$

(DRY): $W_{\text {or }}=[1,0,0,1,0,0,0,0]$
$\theta^{\star}=2$ for edge $(5,6)$

## Iteration 6


(D): $W=[7,4,1,3,7,7,1,0]$
$J=\left\{\begin{array}{l}(4, t),(3,4),(7, t),(2,3), \\ (6,2),(5,6)\end{array}\right\}$

(DRP): $W_{\text {or }}=[1,0,0,0,0,0,0,0]$
$\theta^{\star}=1$ for edge $(s, 5)$
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## Page 444

Iteration 7

(D): $W=[8,4,1,3,7,7,1,0]$
$J=\left\{\begin{array}{l}(4, t),(3,4),(7, t),(2,3), \\ (6,2),(5,6),(s, 5)\end{array}\right\}$

$(\mathrm{DRP}): W_{\text {or }}=[0,0,0,0,0,0,0,0]$

### 15.6.3 Maximum flow

A network $N$ with source $s$ and target $t$ can be described as a linear program as follows:
Maximize $v$
subject to: $A f=\left[\begin{array}{r}+v \\ -v \\ 0 \\ \vdots \\ 0\end{array}\right] \leftarrow$ row $s$

$$
\begin{align*}
& f \leq c \\
& f \geq 0 \tag{15.17}
\end{align*}
$$

where $u=\operatorname{VAL}(f)$ is the value of the flow $f$ and $c=[\operatorname{CAP}(e): e \in E(N)]^{T}$ are the capacities of the edges e in $N$. This is of course equivalent to

Maximize $u$
subject to: $A f+u T \leq 0$

$$
\begin{gather*}
f \leq C  \tag{15.18}\\
-f \leq 0,
\end{gather*}
$$

$$
T[x]=\left\{\begin{array}{l}
-1, \text { if } x=s \\
+1, \text { if } x=t \\
0, \text { otherwise }
\end{array}\right.
$$

where
(Note that $A f+u T \leq 0$ implies $A f+u T=0$ because we maximize $u$.) The set of admissible columns (i.e., edges) is

$$
J=\{\overrightarrow{u v}: f(\overrightarrow{u v})=c \text { or }-f(\overrightarrow{u v})=0\} .
$$

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Thus $J$ is the set of saturated edges together with the zero-flow edges and the dual-restricted primal is:

$$
\begin{aligned}
& \text { Maximize } v \\
& \text { subject to: } \quad \begin{aligned}
A f+v T & \leq 0 \\
f(\overrightarrow{u v}) & \leq 0, \text { for all saturated edges } \overrightarrow{u v} \text { in } 15.18 \\
-f(\overrightarrow{u v}) & \leq 0, \text { for all zero flow edges } \overrightarrow{u v} \text { in } 15.18 \\
f & \leq \overrightarrow{1} \\
v & \leq \overrightarrow{1}
\end{aligned}
\end{aligned}
$$

It is easy to solve the dual-restricted primal in 15.19. We wish to maximize $u$ and so we can try $u=1$. Thus we must choose a flow $f$ such that

$$
[\operatorname{Rows}(A)] f=1
$$

Concerning the edges incident to $s$, the inequalities in 15.19 tell us that we can set $f(\overrightarrow{s u})=1$ if the edge $\overrightarrow{s u}$ has zero flow or is unsaturated and we can set $f(\overrightarrow{u s})=-1$ if the edge $\overrightarrow{u s}$ is saturated and/or does not have zero flow. Let $S 1$ be the set of vertices $u$ incident to $s$ satisfying if $u \longrightarrow s$, then $\overrightarrow{s u}$ has zero flow or is unsaturated; if $u \longrightarrow s$, then $\overrightarrow{u s}$ is saturated or does not have zero flow.
Hence we choose $v_{1} \in S_{1}$ and set the flow on the associated edge to 1 if it leaves $s$ and to -1 if it enters $s$. Now we must consider [Romu1](A) and consider the edges incident to $u 1$. Let $S 2$ be the set of vertices $u$ incident to some $u$ in S1 satisfying

$$
\left.\begin{array}{l}
\text { if } u \longrightarrow s \text {, then } \overrightarrow{s u} \text { has zero flow or is unsaturated; }  \tag{15.20}\\
\text { if } u \longrightarrow s \text {, then } \overrightarrow{u s} \text { is saturated or does not have zero flow. }
\end{array}\right\}
$$

In general let $S k+1$ be the set of vertices $u$ incident to some $u$ in $S k$ satisfying conditions in 15.20 and continue until some $S k$ contains the target $t$. When and if it does, we can choose vertices $v_{i} \in S_{i}, i=1,2$, $\ldots, k-1$ such that sulu2u3...uk-1t is a st-path $P$. We obtain an optimal solution fOPT to the dual-restricted primal in 15.19 as follows:

$$
f_{\text {orr }}(\overrightarrow{u v})=\left\{\begin{aligned}
+1, & \text { if } \overrightarrow{u v} \text { is a forward edge of } P \\
-1, & \text { if } \overrightarrow{u v} \text { is a backward edge of } P \\
0, & \text { if } \overrightarrow{u v} \text { is not an edge of } P
\end{aligned}\right.
$$

We then calculate

$$
\theta_{\star}=\operatorname{MIN}_{u v \in E(P)}\left\{\begin{array}{cl}
(c(\overrightarrow{u v})-f(\overrightarrow{u v})), & \text { if } f_{\text {ort }}(\overrightarrow{u v})=1 \\
(0-(-f(\overrightarrow{u v})), & \text { if } f_{\text {ort }}(\overrightarrow{u v})=-1 \\
\text { page_445 }
\end{array}\right.
$$

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$$
\begin{aligned}
& =\operatorname{Min}_{u v \in E(P)} \begin{cases}(\operatorname{CAP}(\overrightarrow{u v})-f(\overrightarrow{u v})), & \text { if } f_{\text {ort }}(\overrightarrow{u v})=1 \\
f(\overrightarrow{u v}), & \text { if } f_{\text {opt }}(\overrightarrow{u v})=-1\end{cases} \\
& =\operatorname{MIN}\{\operatorname{ResCAP}(u v): u v \in E(P)\},
\end{aligned}
$$

where

$$
\operatorname{RESCAP}(u v)= \begin{cases}\operatorname{CAP}(\overrightarrow{u v})-f(\overrightarrow{u v}), & \text { if } u v \text { is a forward edge } \\ f(\overrightarrow{v u}), & \text { if } u v \text { is a backward edge } .\end{cases}
$$

We update the flow by

$$
f \leftarrow f+\theta_{\star} f_{\text {opt }}
$$

recompute the set of admissible edges $J$ to get the new dual-restricted primal and repeat until $J$ is empty at which point the flow $f$ will be maximum.
It is now easy to see the similarity of this method to that in Section 8.2 and realize that the primal-dual algorithm for network flow is exactly the Ford-Fulkerson algorithm.

## Exercises

15.6.1 Consider Problem 15.2 the Weighted Matching problem.

Problem 15.2: Weighted Matching

Instance: undirected graph $G$ and weight we $\geq 0$ for each edge $e$ of $G$.
Find: a matching $M$ of $G$ with maximal possible weight

$$
\mathrm{W} \mathrm{~T}(M)=\sum_{e \in E(M)} w_{e} .
$$

Formulate a primal-dual algorithm for solving Problem 15.2 and give an interpretation for the restricted primal.
15.6.2 Use the primal-dual algorithm as discussed in Section 15.6.2 to find a shortest path between each pair of nodes in the graph given in Figure 15.5.

### 15.7 Notes

The primal-dual algorithm was first described in 1956 by DANTZG; see [33]. Our treatment is similar to that of PAPADIMITRI OU and STIEGLITZ; see [94].

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FI GURE 15.5
An instance of the shortest-path problem
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## page_448

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16

## Discrete Linear Programming

### 16.1 Introduction

An integer linear program (ILP) is a linear program in which the variables have been constrained to be integers.

$$
\begin{gathered}
\text { Minimize: } Z=c T X \\
\text { subject to: } A X \leq b \\
X>0 \\
X \text { integral. }
\end{gathered}
$$

If all of the variables are each constrained to a finite set of values, we say that the integer linear program is discrete. Notice that frequently the equality constraints force the variables to be discrete, for if bi/ai, $j>0$ for all $j$ in the constraint

$$
\sum_{i=1}^{n} a_{i, j} x_{j}=b_{i},
$$

then $x j$ cannot exceed $m_{j}=\left\lfloor b_{i} / a_{i}, j\right\rfloor$. Hence $x_{j} \in\left\{0,1,2, \ldots, m_{j}\right\}$ for all $j$.
DEFINITION 16.1: A discrete linear program (DLP) is an integer linear program in which the variables are a
bounded
subject to: $A X=b$
$0 \leq x j \leq m j, j=1,2, \ldots, n$ $X$ integral.
Consider Problem 16.1, the Knapsack problem. This problem is motivated by what to carry on a proposed hiking trip. The weight limit on how much can be carried is the capacity $M$. Each of the $n$ objects under consideration have a
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certain weight wi and each has a certain value or profit pi, $i=1,2, \ldots, n-1$. Furthermore each object can be either carried or left behind. We cannot choose to carry a fraction of an object.

Problem 16.1: Knapsack
I nstance: profits p1, p1, p2,..., pn;
weights w1, w1, w2,..., wn; and capacity $M$;
Find: the maximum value of

$$
\begin{aligned}
& P=\sum_{i=1}^{n} p_{i} x_{i} \\
& \text { subject to } \\
& \sum_{i=1}^{n} w_{i} x_{i} \leq M
\end{aligned}
$$

and $\left[x_{1}, \ldots, x_{n}\right] \in\{0,1\}^{n}$.
This problem can be formulated as the discrete linear program:

$$
\begin{gathered}
\text { Minimize: } Z=-(p 1 \times 1+p 1 \times 1+\ldots+p n \times n) \\
\text { subject to: } w 1 \times 1+w 1 \times 1+\ldots+w n x n \leq M \\
0 \leq x j \leq 1, j=1,2, \ldots, n \\
X \text { integral. }
\end{gathered}
$$

### 16.2 Backtracking

A discrete integer linear program can be solved with a backtracking search. For example, the Knapsack problem can be solved with Algorithm 16.2.1.

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Algorithm 16.2.1: KNAPSACK 1( $\ell)$ if $\ell>n$
if $\sum_{i=1}^{n} w_{i} x_{i} \leq M$
then $\left\{\begin{array}{l}\text { CurrentProfit } \leftarrow \sum_{i=1}^{n} p_{i} x_{i} \\ \text { if } \text { CurrentProfit }>\text { Optima }\end{array}\right.$
else
$\left\{\begin{array}{l}x_{\ell} \leftarrow 1 \\ \text { KNAPSACK } 1(\ell+1) \\ x_{\ell} \leftarrow 0 \\ \text { KNAPSACK } 1(\ell+1)\end{array}\right.$
Initially Algorithm 16.2.1 is started with $\ell=1$.
In general the backtracking method to solve a discrete integer linear program is performed by computing for a partial solution, in which values for $x 1, x 2, \ldots, x \ell-1$ have been assigned, the set of possible values $C \ell$ for $x \ell$. Each possible value is examined to see whether the partial solution can be extended with it. The general backtracking algorithm is provided in Algorithm 16.2.2

Algorithm 16.2.2: BACKTRACK( $\ell)$
if $[x 1, x 1, \ldots, x \ell]$ is a feasible solution
then process it Compute $C l$
for each $x \in \mathcal{C}_{\ell}$
do $\left\{\begin{array}{l}x_{\ell} \leftarrow x \\ \operatorname{BACKTRACK}(\ell+1)\end{array}\right.$
Algorithm 16.2.1, is an application of Algorithm 16.2.2 with $C \ell=\{1,0\}$. We can improve the running time of a backtracking algorithm if we can find efficient ways of reducing the size of the choice sets $C l$ This process is called pruning.
For the Knapsack problem, one simple method is to observe that we must have

$$
\sum_{i=1}^{\ell} w_{i} x_{i} \leq M
$$

for any partial solution $[x 1, x 2, \ldots, x \ell-1]$. In other words, we can check partial solutions to see if the feasibility condition is satisfied. Consequently, if $\ell \leq n$ and

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we set

$$
\text { CurrentWt }=\sum_{i=1}^{\ell-1} w_{i} x_{i},
$$

then we have

$$
\mathcal{C}_{\ell}= \begin{cases}\{1,0\}, & \text { if CurrentWt }+w_{\ell} \leq M ; \\ \{0\}, & \text { otherwise. }\end{cases}
$$

Using Algorithm 16.2.2 as a template, we obtain Algorithm 16.2.3, which is invoked with $\ell=1$ and CurrentW $=0$.

Algorithm 16.2.3: KNAPSACK2( $\ell)$ if $\ell>n$


A backtracking algorithm with simple pruning for solving the discrete integer linear program 16.2 in which the coefficient matrix $A=[A i, A 2, \ldots, A n]$ and $b$ consists of only non-negative entries is given as Algorithm 16.2.4.
then $\left\{\begin{array}{l}\text { if } \sum_{i=1}^{n} c_{i} x_{i}<\text { Optimal } Z \\ \text { then }\left\{\begin{array}{l}\text { Optimal } Z \leftarrow \sum_{i=1}^{n} c_{i} x_{i} \\ \text { Optimal } X \leftarrow\left[x_{1}, \ldots, x_{n}\right]\end{array}\right.\end{array}\right.$
comment: Compute Cl
if $\ell=n$

$$
\text { then } C_{\ell} \leftarrow \emptyset
$$

else
$\left\{\begin{array}{l}C_{\ell} \leftarrow\{ \} \\ \text { for } x=0 \text { to } m_{j} \\ \text { do }\left\{\begin{array}{c}\text { if } \text { Current } W t[i]+A_{i} x \leq b \\ \text { then } C_{\ell} \leftarrow C_{\ell} \cup\{x\}\end{array}\right.\end{array}\right.$
for each $x \in \mathcal{C}_{\ell}$

$$
\text { do }\left\{\begin{array}{l}
x_{\ell} \leftarrow x \\
\text { CurrentWt }=\text { CurrentWt }+A_{i} x_{\ell} \\
\text { BACKTRACK2 }(\ell+1) \\
\text { CurrentWt }=\text { CurrentWt }-A_{i} x_{\ell}
\end{array}\right.
$$

### 16.3 Branch and bound

Another strategy for solving an integer linear program is to first ignore the integer constraints and solve the corresponding linear program. This linear program is called the relaxation of the integer linear program. If a solution is found to the relaxed integer linear program, it can be rounded to the nearest integer solution. Although this may seem plausible, it generally leads to solutions that are either infeasible or far from the optimal solution. We illustrate this difficulty with the following integer linear program:

Minimize: $Z=-x 1-x 2$
subject to: $-2 \times 1+7 \times 2 \leq 14$
$2 \times 1-2 \times 2 \leq 1$
$x 1, x 2 \geq 0$
$x 1, x 2$ integral.
This problem is solved graphically in Figure 16.1 and we see that there are six


## FI GURE 16.1

The integer points closest to the optimal solution to the standard linear program are infeasible in the integer linear program.
feasible solutions to this integer linear program: $(0,0),(0,1),(1,1),(0,2),(1,2)$ and $(2,2)$. The value of the object function, respectively, at these points is: $0,-1,-2,-2,-3$, and -4 . Hence $(2,2)$ is the optimal solution to this integer linear program. On the other hand, the optimal solution to the relaxed integer linear program is $(3.5,3)$ and the value of the objective function at this point is -6.5 . The nearest integer points are $(3,3)$ and $(4,3)$ having values -6 and -7 . Neither of these points are feasible and the value of the objective function at them is far from the value at the true optimal solution.
Thus in practice the relaxation strategy can result in misleading information. However, not all is lost. It is not to difficult to see that a bound on the value of the objective function is obtained.
THEOREM 16.1 Let $\widehat{X}$ be an optimal solution to relaxation of the integer linear program

$$
\begin{gather*}
\text { Minimize: } Z=c T X  \tag{16.4}\\
\text { subject to: } A X=b \\
X \geq 0 \\
X \text { integral }
\end{gather*}
$$

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Then any solution $X$ to the integer linear program 16.4 satisfies

$$
c^{T} X \leq c^{T} \hat{X}
$$

PROOF If $X$ is a feasible solution to the integer linear program, it is also a feasible solution to the linear program, in which the integer constraints have been removed.
If values have been assigned to each of $x 1, x 2, \ldots, x \ell-1$, then the remaining variables $x \ell, x \ell+1, \ldots, x n$ must satisfy

$$
\begin{array}{lc}
\text { Minimize: } & \widehat{Z}=\widehat{c}^{T} \widehat{X} \\
\text { subject to: } & \widehat{A} \widehat{X} \leq \widehat{b} \\
& \widehat{X} \geq 0 \\
& \widehat{X} \text { integral, }
\end{array}
$$

where

$$
\begin{aligned}
\widehat{A} & =\left[A_{\ell}, A_{\ell+1}, \ldots, A_{n}\right] \\
\widehat{X} & =\left[x_{\ell}, x_{\ell+1}, \ldots, x_{n}\right] \\
\widehat{c} & =\left[c_{\ell}, c_{\ell+1}, \ldots, c_{n}\right] \\
\widehat{b} & =b-\text { CurrentWt }
\end{aligned}
$$

and

$$
\text { CurrentWt }=\sum_{j=1}^{\ell-1} x_{j} A_{j}
$$

(Here $A j$ is the $j$ th column of the coefficient matrix $A$.) Thus according to Theorem 16.1, any feasible solution $X$ that extends $x 1, x 2, \ldots, x \ell-1$ has value $Z=c T X$ no larger than

$$
B=\text { Current } Z+\left\lfloor\widehat{Z}_{\text {orr }}\right\rfloor
$$

where $\widehat{Z}_{\text {or }}$ is value of the objective function $\widehat{Z}=\widehat{c}^{T} \widehat{X}$ for the linear program

$$
\begin{aligned}
& \text { Minimize: } \\
& \widehat{Z}=\widehat{c}^{T} \widehat{X} \\
& \text { subject to: } \hat{A} \hat{X} \leq \hat{b} \\
& \widehat{X} \geq 0 \text {. }
\end{aligned}
$$

Consequently, if a feasible solution $X 0$ has already been obtained that has value $Z=C T X 0 \leq B$, then no extension of the given assignment to the variables $x 1, x 2, \ldots, x \ell-1$ will lead to an improved $Z$ value. Hence the search for such extensions can be aborted. This method is known as branch and bound and is recorded as Algorithm 16.3.1, which we start with $\ell=1$, CurrentWt=0, CurrentZ=0, and Optimal $Z=\infty$.
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Algorithm 16.3.1: BACKTRACK3( $\ell)$
then $\left\{\begin{array}{l}\text { if CurrentWt }=b \\ \text { then }\left\{\begin{array}{l}\text { if Current } Z<O \text { Optimal } Z \\ \text { then }\left\{\begin{array}{l}\text { for } j=1 \text { to } n \text { do Optimal } X[j] \\ O p t i m a l Z\end{array} \text { Current } Z\right.\end{array} \leftarrow X[j]\right. \\ \text { return }\end{array}\right.$
Compute Cl
Use the simplex algorithm to solve the linear program:
$\begin{array}{lc}\text { Minimize: } & \widehat{Z}=\hat{c}^{T} \hat{X} \\ \text { subject to: } & \widehat{A} \widehat{X}=\widehat{b}, \text { where }\end{array}\left\{\begin{array}{l}\widehat{A}=\left[A_{\ell}, A_{\ell+1}, \ldots, A_{n}\right] \\ \hat{X}=\left[x_{\ell}, x_{\ell+1}, \ldots, x_{n}\right] \\ \hat{C}=\left[c_{\ell}, c_{\ell+1}, \ldots, c_{n}\right] \\ \widehat{b}=b-\text { CurrentWt }\end{array}\right.$
ifthe linear program has optimal value $\widehat{Z}_{\text {orr }}$ at $\widehat{X}_{\text {orr }}$

if the linear program is infeasible then return


## FI GURE 16.2

## The backtracking state space search tree that results when Algorithm 16.3.1 is applied to the integer linear program 16.3.

By way of example we provide the backtracking state space search tree in Figure 16.2 that results when Algorithm 16.3.1 is applied to the integer linear program 16.3. Observe that adding the two constraints of linear program 16.3 we see that $x 2 \leq 3=m 2$ and thus $x 1 \leq 3=m 1$ as well. These bounds means that the linear program is a discrete linear program and indeed we can apply Algorithm 16.3.1. Each leaf-node in Figure 16.2 is decorated with $I, P$, or an arrow. The $I$ indicates that the corresponding reduced integer linear program is infeasible, the $P$ indicates pruning by a previously obtained feasible solution and an arrow indicates that a new optimal solution XOPT has been found. Its value $Z=c T X O P T$ is also given. For this example the final optimal solution is $X O P T=[2,2]$ with value $Z=-4$. This agrees with our previous analysis. As a second example consider the linear program

Minimize: $Z=-2 \times 1-3 \times 2-x 3-2 \times 4$
subject to: $x 1+2 \times 2+3 \times 3+2 \times 4 \leq 8$
$-3 x 1+4 \times 2+x 3+3 \times 4 \leq 8$
$3 \times 1-4 \times 2-6 \times 3-10 \times 4 \leq-20$
$x 1, x 2, x 3, x 4 \geq 0$
$x 1, x 2, x 3, x 4$ integral.
The constraint $x 1+2 \times 2+3 \times 3+2 x 4 \leq 8$ provides the upper bounds $m 1=8, m 2=4, m 3=2$, and $m 4=4$, on the variables $x 1, x 2, x 3$, and $x 4$, respectively. Thus we can take $C 1=\{0,1, \ldots, 8\}, C 2=C 4=\{0,1, \ldots, 4\}$, and $C 3=\{0,1,2\}$. The backtracking state space search tree is given in Figures 16.3 and 16.4.

[0000] [0001][0002] [0003][0004] [0010] [0011] [0012][0013][0014]



FI GURE 16.3
The first part of the backtracking state space search tree that results when Algorithm 16.3.1 is applied to the integer linear program 16.5. So far we see that $X O P T=[0,1,1,1]$ gives $z=-6$.


FI GURE 16.4
Continuation of Figure 16.3. Notice the subtree with root [101] is pruned because [1003] gave $z=-8$. The final solution appears at $[20]$ and is $X O P T=[2,0,0,3]$ with objective value $z=-10$.

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[2] $B=-11$
[1] $B=-10$
[0] $B=-7$

$Z=-10$
$X_{\mathrm{orr}}=[2,0,0,3]$

## FI GURE 16.5

## Dramatic pruning when the subtrees are processed in order of the bounds

Notice in this second example, that if we were to process the subtree with root [2] prior to the subtree [0] and [1], pruning would be dramatic and the search would be quite short. This is because the bound obtained when the relaxed linear program with $x 1=2$ leads to a bound of -10 , and there is an integer solution that achieves this bound, namely, $X=[2,0,0,3]$ in the subtree with root [2]. The bound obtained when the root is [0] and [1] is -7 and -10 , respectively, so these subtrees will be pruned if [2] is examined first. The search tree with this type of pruning is given in Figure 16.5.
One way to implement this type of pruning is as follows. After values for $x 1, x 2, \ldots, x \ell-1$ have been assigned, precompute the bounds obtained for each possible assignment for $x \ell$ and then process the assignment in order of increasing bound. Algorithm 16.3.2 describes this method. The danger in using this method is that at each node we are required to do sorting. If there are $m \ell$ choices for $x \ell$, this will require $O(m \ell \log (m \ell))$ additional steps at each node that occurs at level $\ell-1$ in the search tree. This computation may be prohibitive. (There no additional calls to the simplex algorithm.)
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Algorithm 16.3.2: BACKTRACK4( $\ell$ )
if $\ell>n$
then $\left\{\begin{array}{l}\text { if CurrentWt }=b \\ \text { then }\left\{\begin{array}{c}\text { if } \operatorname{Current} Z<O p t i m a l Z \\ \text { then }\left\{\begin{array}{l}\text { for } j=1 \text { to } n \text { do Optimal } X[j] \\ \text { Optimal } Z \leftarrow C \text { Current } Z\end{array}\right. \\ \text { return }\end{array}\right.\end{array}\right.$
$N \nleftarrow 0$
if $\ell<n$ then
(for $x \leftarrow 0$ to $m_{\ell}$
Use the simplex algorithm to solve the linear program:
Minimize:

$$
\begin{aligned}
\widehat{Z}=\hat{c}^{T} \widehat{X} \\
\widehat{A} \widehat{X}=\widehat{b} \\
\widehat{X} \geq 0
\end{aligned}, \text { where }\left\{\begin{array}{l}
\widehat{A}=\left[A_{\ell+1}, \ldots, A_{n}\right] \\
\widehat{X}=\left[x_{\ell+1}, \ldots, x_{n}\right] \\
\widehat{c}=\left[c_{\ell+1}, \ldots, c_{n}\right] \\
\widehat{b}=b-\text { CurrentWt }-x * A_{\ell}
\end{array}\right.
$$

if the linear program has optimal value $\widehat{Z}_{\text {or }}$ at $\widehat{X}_{\text {or }}$
do

$$
\left\{\begin{array}{l}
B \leftarrow \text { Current } Z+\left\lfloor\widehat{Z}_{\text {or }}\right\rfloor \\
\text { if } \widehat{X}_{\text {or }} \text { is integer valued }
\end{array}\right.
$$

do Minimize.
subject to:
,

$$
\text { then }\left\{\begin{array}{l}
\text { if } B<O p t i m a l Z \text { then } \\
\left\{\begin{array}{l}
\text { for } i \leftarrow 1 \text { to } \ell-1 \text { do Optimal } X[i] \leftarrow X[i] \\
O \text { optimal } X[\ell] \leftarrow x \\
\text { for } i \leftarrow \ell+1 \text { to } n \text { do Optimal } X[i] \leftarrow \widehat{X}_{\text {orr }}[i] \\
\text { OptimalZ } \leftarrow B
\end{array}\right.
\end{array}\right.
$$

else if $B<O$ Otimal $Z$ then

$$
\left\{\begin{array}{l}
N_{\ell} \leftarrow N_{\ell}+1 \\
C_{\ell}\left[N_{\ell}\right] \leftarrow(x, B)
\end{array}\right.
$$

if the linear program is unbounded

$$
\text { then }\left\{\begin{array}{l}
N_{\ell} \leftarrow N_{\ell}+1 \\
C_{\ell}\left[N_{\ell}\right] \leftarrow(x,-\infty)
\end{array}\right.
$$

if the linear program is infeasible then return
Sort the ordered pairs in Cl in order of increasing second coordinate for $h \leftarrow 1$ to $N \ell$
do $\left\{\begin{array}{l}\text { if } \text { Bound }_{\ell}[h] \geq \text { Optimal } Z \text { then return } \\ x_{\ell} \leftarrow \text { first coordinate of the ordered pair } C_{\ell}[h] \\ \text { CurrentWt } \leftarrow \text { Current } W t+A_{\ell} x_{\ell} \\ \text { Current } Z \leftarrow \text { Current } Z+c_{\ell} x_{\ell} \\ \text { BACKTRACK } 3(\ell+1) \\ \text { CurrentWt } \leftarrow \text { CurrentWt }-A_{\ell} x_{\ell} \\ \text { Current } Z \leftarrow \text { Current } Z-c_{\ell} x_{\ell}\end{array}\right.$
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The relaxed linear program for the Knapsack problem is easy to solve. A straightforward method which uses a greedy strategy, to solve the relaxed Knapsack problem is given in Algorithm 16.3.3. (See Exercises 16.3.4, 16.3.5, and 16.3.6.) It returns the optimal profit for the Knapsack problem, in which the integer constraint has been relaxed to allow non-integer (i.e., rational) solutions.
Algorithm 16.3.3: RELAXEDKNAPSACK ( $p 1, p 2, . ., P n, w 1, \ldots, w n, M$ )
permute the indices so that $p 1 / w 1 \geq p 2 / w 2 \geq P n / w n$

$$
\begin{aligned}
& i \leftarrow 1 \\
& P \leftarrow 0 \\
& W \leftarrow 0 \\
& \text { for } j \leftarrow 1 \text { to } n \\
& \text { do } x j \leftarrow 0 \\
& \text { while } W<M \text { and } i<n
\end{aligned}
$$

return ( $P$ )
To solve an instance of the Knapsack problem, it will be useful to presort the objects in non-decreasing order of the profit/weight ratio, before we begin the backtracking algorithm. Then, when we apply Algorithm 16.3.3, the first step will be unnecessary, and consequently RELAXEDKNAPSACK will run faster. Thus, we will assume that

$$
\frac{p_{1}}{w_{1}} \geq \frac{p_{2}}{w_{2}} \geq \cdots \geq \frac{p_{n}}{w_{n}}
$$

The improved Knapsack algorithm is given as Algorithm 16.3.4.

Algorithm 16.3.4: KNAPSACK3 ( $\ell$ )
if $\ell=n$
then $\left\{\begin{array}{l}\text { if CurrentProfit }>\text { OptimalProfit } \\ \text { then }\left\{\begin{array}{l}\text { OptimalProfit } \leftarrow \text { CurrentProfit } \\ \text { OptimalX } \leftarrow\left[x_{1}, \ldots, x_{n}\right]\end{array}\right.\end{array}\right.$ if $\ell=n$ then $C_{\ell} \leftarrow \emptyset$ $x \in \mathcal{C}_{\ell}$ $B \leftarrow$ CurrentProfit +RELAXEDKNAPSACK ( $p \ell, p n, w \ell, \ldots, w n, M-C u r r e n t W t)$
for each $x \in \mathcal{C}_{\ell}$
do $\left\{\begin{array}{l}\text { if } B \leq \text { OptimalProfit then return } \\ x_{\ell} \leftarrow x \\ \text { CurrentWt }=\text { CurrentWt }+w_{\ell} x_{\ell} \\ \text { CurrentProfit }=\text { CurrentProfit }+p_{\ell} x_{\ell} \\ \text { KNAPSACK3 }(\ell+1) \\ \text { CurrentWt }=\text { CurrentWt }-w_{\ell} x_{\ell} \\ \text { CurrentProfit }=\text { CurrentProfit }-p_{\ell} x_{\ell}\end{array}\right.$

## Exercises

16.3.1 Solve graphically the following integer linear programs:

$$
\begin{aligned}
& \text { (b) }\left\{\right.
\end{aligned}
$$

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Use Algorithm 16.3.4 to solve the following instances of the Knapsack problem.

| (a) | Profits | 122 | 2 | 144 | 133 | 52 | 172 | 169 | 50 | 11 | 87 | 127 | 31 | 10 | 132 | 59 |
| :--- | :--- | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
|  | Weights | 63 | 1 | 71 | 73 | 24 | 79 | 82 | 23 | 6 | 43 | 66 | 17 | 5 | 65 | 29 |
| Capacity | 323 |  |  |  |  |  |  |  |  |  |  |  |  |  |  |  |
| (b) | Profits | 143 | 440 | 120 | 146 | 266 | 574 | 386 | 512 | 106 | 418 | 376 | 124 | 48 | 535 | 55 |
|  | Weights | 72 | 202 | 56 | 73 | 144 | 277 | 182 | 240 | 54 | 192 | 183 | 67 | 23 | 244 | 29 |
| (c)Capacity 1019 | 818 | 460 | 267 | 75 | 621 | 280 | 555 | 214 | 721 | 427 | 78 | 754 | 704 | 44 | 371 |  |
|  | Profits | Weights | 380 | 213 | 138 | 35 | 321 | 138 | 280 | 118 | 361 | 223 | 37 | 389 | 387 | 23 |
|  | Capacity | 1617 |  |  |  |  |  |  |  |  |  |  | 191 |  |  |  |

16.3.2 Algorithm 16.3.4 does not take advantage of the fact that given a partial solution $X^{\prime}$, if the optimal solution to the corresponding relaxed knapsack problem is integer-valued it gives that best solution $X$ that extends $X^{\prime}$. Hence there is no need to pursue further extensions, and the search can be pruned. This type of pruning was done for the general problem in Algorithm 16.3.1. Construct a new algorithm that takes advantage of this pruning for the Knapsack problem. Test your algorithm on the data in Exercise 1. How does it compare with Algorithm 16.3.4?
16.3.3 Program Algorithm 16.3.1 and use it to solve the following integer linear program.

$$
\begin{array}{lrl}
\text { Minimize: } & Z=x_{1}-3 x_{2}+x 5 \\
\text { subject to: } & x_{1}+2 x_{2}+3 x_{4} & \leq 6 \\
& 5 x_{1}+4 x_{2}+9 x_{5} & \leq 20 \\
& X & \geq 0 \\
& X & \text { integral. }
\end{array}
$$

16.3.4 Prove that Algorithm 16.3.3 does indeed solve

Problem 16.2: Relaxed Knapsack
I nstance: profits p1, p1, p2,... pn;
weights $w 1, w 1, w 2, \ldots w n$; and capacity M;
Find: the maximum value of

$$
P=\sum_{i=1}^{n} p_{i} x_{i}
$$

subject to

$$
\sum_{i=1}^{n} w_{i} x_{i} \leq M
$$

and $x 1, x 2 \ldots, x n$ are rational
as was claimed.
16.3.5 Verify that the simplex algorithm when applied to Problem 16.2 gives exactly the same result as Algorithm 16.3.3.
16.3.6 Determine the running time complexity of Algorithm 16.3.3.
16.3.7 In Section 9.6 we studied the traveling salesman problem. Let $W$ (uiuj) be the non-negative weight on the edge viuj of the complete graph with vertices $V=$
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$\{U 1, U 2, \ldots, u n\}$. The traveling salesman problem is to find a hamilton cycle $C$ that has minimum weight

$$
\sum_{u v \in E(C)} W(u v) .
$$

Let $x_{u v} \in\{0,1\}$ variable that denotes an edge $u v$ in the hamilton cycle if $x u u=1$ and an edge not in the cycle if $x u u=0$. Show that the optimal solution to the discrete linear program

| Minimize: | $Z=\sum_{u v} x_{u v} W(u v)$ |  |  |
| ---: | :--- | ---: | :--- |
| subject to: | $\sum_{u \in V \backslash\{v\}} x_{u v} W(u v)$ | $=2$, | $v \in V$ |
|  | $\sum_{u v \in\{S, \widehat{S}]} x_{u v} W(u v)$ | $\geq 1$, | $\emptyset \neq S \subset V$ |
|  | $x_{u v} \in\{0,1\}$, | for each edge $u v$ |  |

solves the traveling salesman problem.
16.3.8 Prove that the the problem

Problem 16.3: ILP decision
Instance: an integer linear program.
Question:does the the given integer linear program have a feasible solution?
is NP-complete. (Hint: Transform from 3-Sat.)
16.3.9 Use Algorithm 16.3.1 to determine the maximum number of edge disjoint triangles in the complete graph $K n$, for $n=7,8,9, \ldots, 13$. Hint: Use the $\binom{n}{2}$ by $\binom{n}{3}$ matrix $A$ whose rows are labeled by the edges, whose columns are labeled by the triangles and whose [ $e, t$ ]-entry is 1 if e is an edge on triangle $t$ and is 0 otherwise. When $n=1,3(\bmod 6)$, then the maximum number edge disjoint triangles is $n(n-1) / 6$. The corresponding collection of edge disjoint triangles is called a Steiner triple system.

### 16.4 Unimodular matrices

In Sections 15.6.2 and 15.6.3 we studied primal-dual algorithms for the Shortest Path and Max-Flow problems, respectively. Surprisingly we found that their optimal solutions were always integral although we never required that the variables be constrained to be integers. The reason for this is that the node-edge incidence matrix of any digraph is totally unimodular.
DEFINITION 16.2: An $m$ by $n$ integer valued matrix is totally unimodular if the determinant of each square submatrix is equal to 0,1 , or -1 .
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THEOREM 16.2 Every basis feasible solution to the linear program
Minimize: $Z=c T X$
subject to: $A X=b$
where $A$ is a totally unimodular $m$ by $n$ matrix, and $b$ is integer-valued, is integer-valued.
PROOF If $X$ is the basis feasible solution corresponding to the submatrix $B$ composed of $m$ linearly independent columns $A_{j_{1}}, A_{j_{2}}, \ldots, A_{j_{m}}$, then

$$
X_{B}=B^{-1} b=\frac{\operatorname{ADJ}(B)}{\operatorname{det}(B)} b
$$

where $A D J(B)$ is the adjoint of $B$. Hence $X B$ has integer values, because the total unimodularity of $A$ implies that the $\operatorname{det}(B)= \pm 1$. Finally

$$
X[j]= \begin{cases}X_{B}\left[j_{\ell}\right], & \text { if } j_{\ell}=j \\ 0, & \text { otherwise }\end{cases}
$$

and so the entries of $X$ are integers.
THEOREM 16.3 Every basis feasible solution to the linear program
Minimize: $Z=c T X$
subject to: $A X \leq b$ $X \geq 0$,
where $A$ is a totally unimodular $m$ by $n$ matrix, and $b$ is integer-valued, is integer-valued.
PROOF Adding slack variables $Y$ we obtain the following equivalent linear pro gram:

$$
\begin{array}{rlrl}
\text { Minimize: } & Z=c^{T} X & \\
\text { subject to: } \quad\left[A, I_{m}\right]\left[\begin{array}{c}
X \\
Y
\end{array}\right] & =b \\
X, Y & \geq 0 .
\end{array}
$$

Thus we need only show that if $A$ is a totally unimodular, then $[A, I m$ ] is totally unimodular, where $I m$ is the $m$ by $m$ identity matrix. Then the result follows
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from Theorem 16.2. Let $M$ be a square nonsingular submatrix of $[A, I m]$; then after a suitable permutation of the rows and columns we see that $M$ has the form

$$
\left[\begin{array}{c|c}
B & 0 \\
\hline N & I_{m-k}
\end{array}\right]
$$

where $B$ is a square $k$ by $k$ submatrix of $A$ and $l \ell$ is the $\ell$ by $\ell$ identity matrix, for some $k$ and $\ell$. The determinant of $B$ is $\pm 1$, because $A$ is totally unimodular and permutations of the rows and columns of $M$ only change the determinant of $M$ by a factor of $\pm 1$. Thus

$$
\operatorname{det}(M)= \pm \operatorname{det}(B) \operatorname{det}(I K)= \pm 1 .
$$

THEOREM 16.4 The node-edge incidence matrix of a directed graph is totally unimodular.
PROOF Let $G=(V, E)$ be a directed graph and let $A$ be its node-edge incidence matrix. Then

$$
A[v, e]=\left\{\begin{aligned}
+1, & \text { if } e \text { leaves } v, \\
-1, & \text { if } e \text { enters } v, \\
0, & \text { otherwise }
\end{aligned}\right.
$$

In particular, $A$ has exactly two non-zero entries in each column, one is a -1 and the other is +1 . Let $M$ be any $k$ by $k$ submatrix of $A$. If $k=1$, then clearly $\operatorname{det}(M)=0,+1$, or -1 . So suppose $k>1$ and proceed by induction. If $M$ contains a column of zeros, then $\operatorname{det}(M)=0$. If $M$ contains a column $j$ with a single non-zero entry $a= \pm 1$ say in row $i$, then $\operatorname{det}(M)=a \operatorname{det}(N)$ where $N$ is the $k-1$ by $k-1$ submatrix obtained by removing column $j$ and row $i$. By induction we have $\operatorname{det}(N)=0,+1$ or -1 , and so $\operatorname{det}(M)=0,+1$ or -1 . Finally we have the case when each column has two non-zero entries in each column. One is a -1 and the other is a +1 ; hence each column sums to zero and therefore $M$ is singular and hence has determinant zero.

## Exercises

16.4.1 Show that the following statements are all equivalent:
(a) $A$ is totally unimodular.
(b) The transpose of $A$ is totally unimodular.
(c) $[A, I m]$ is totally unimodular.
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(d) A matrix obtained by deleting a row or column of $A$ is totally unimodular.
(e) A matrix obtained by multiplying a row or column of $A$ by -1 is totally unimodular.
(f) A matrix obtained by duplicating a row or column of $A$ is totally unimodular.
(g) A matrix obtained by pivoting on an entry of $A$ is totally unimodular. 16.4.2 Show that the matrix

$$
\left[\begin{array}{rrrrr}
1 & -1 & 0 & 0 & -1 \\
-1 & 1 & -1 & 0 & 0 \\
0 & -1 & 1 & -1 & 0 \\
0 & 0 & -1 & 1 & -1 \\
-1 & 0 & 0 & -1 & 1
\end{array}\right]
$$

is totally unimodular.
16.4.3 Let $G$ be an undirected bipartite graph with bipartition $(X, Y)$. Show that the vertex-edge incidence matrix $M$ of $G$ is totally unimodular.

$$
M[v, e]= \begin{cases}1, & \text { if } v \text { is incident to } e \\ 0, & \text { otherwise }\end{cases}
$$

### 16.5 Notes

An excellent treatment of backtracking algorithms is given in the book by KREHER and STINSON [81]. The treatment of the Knapsack problem and exercise 16.3.1 is taken from this book. Two other good books that discuss general integer linear programming are PAPIDIMITRIOU and STEIGLITZ [94] and NEMHAUSER and WOLSEY [92].
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