# Comprehensive Introduction to Linear Algebra 

## PART II - POLYNOMIALS and CANONICAL FORMS

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$$
H=\left(\begin{array}{ccccccccccccccc}
0 & \cdots & 0 & h_{1 j_{1}} & X & \cdots & X & X & X & \cdots & X & X & X & \cdots & X \\
0 & \cdots & 0 & 0 & 0 & \cdots & 0 & h_{2 j_{2}} & X & \cdots & X & X & X & \cdots & X \\
\vdots & \vdots & \vdots & \vdots & \vdots & & \vdots & \vdots & \vdots & & \vdots & \vdots & \vdots & & \vdots \\
0 & \cdots & 0 & 0 & 0 & \cdots & 0 & 0 & 0 & \cdots & 0 & h_{r j_{r}} & X & \cdots & X \\
- & - & - & - & - & - & - & - & - & - & - & - & & - & - \\
& & & & & & & & & & & & & & \\
& & & & & & & & &
\end{array}\right)
$$

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PART II - POLYNOMIALS AND CANONICAL FORMS

Joel G. Broida<br>S. Gill Williamson

## Preface (Part II)

This book, Part 2 - Polynomials and Canonical Forms, covers Chapters 6 through 8 of the book A Comprehensive Introduction to Linear Algebra (Addison-Wesley, 1986), by Joel G. Broida and S. Gill Williamson. Chapter 6, Polynomials, will be review to some readers and new to others. Chapters 7 and 8 supplement and extend ideas developed in Part I, Basic Linear Algebra, and introduce the very powerful method of canonical forms. The original Preface, Contents and Index are included.

## Preface (Parts I, II, III)

As a text, this book is intended for upper division undergraduate and beginning graduate students in mathematics, applied mathematics, and fields of science and engineering that rely heavily on mathematical methods. However, it has been organized with particular concern for workers in these diverse fields who want to review the subject of linear algebra. In other words, we have written a book which we hope will still be referred to long after any final exam is over. As a result, we have included far more material than can possibly be covered in a single semester or quarter. This accomplishes at least two things. First, it provides the basis for a wide range of possible courses that can be tailored to the needs of the student or the desire of the instructor. And second, it becomes much easier for the student to later learn the basics of several more advanced topics such as tensors and infinite-dimensional vector spaces from a point of view coherent with elementary linear algebra. Indeed, we hope that this text will be quite useful for self-study. Because of this, our proofs are extremely detailed and should allow the instructor extra time to work out exercises and provide additional examples if desired.

A major concern in writing this book has been to develop a text that addresses the exceptional diversity of the audience that needs to know something about the subject of linear algebra. Although seldom explicitly acknowledged, one of the central difficulties in teaching a linear algebra course to advanced students is that they have been exposed to the basic background material from many different sources and points of view. An experienced mathematician will see the essential equivalence of these points of view, but these same differences seem large and very formidable to the students. An engineering student for example, can waste an inordinate amount of time because of some trivial mathematical concept missing from their background. A mathematics student might have had a concept from a different point of view and not realize the equivalence of that point of view to the one currently required. Although such problems can arise in any advanced mathematics course, they seem to be particularly acute in linear algebra.

To address this problem of student diversity, we have written a very selfcontained text by including a large amount of background material necessary for a more advanced understanding of linear algebra. The most elementary of this material constitutes Chapter 0, and some basic analysis is presented in three appendices. In addition, we present a thorough introduction to those aspects of abstract algebra, including groups, rings, fields and polynomials over fields, that relate directly to linear algebra. This material includes both points that may seem "trivial" as well as more advanced background material. While trivial points can be quickly skipped by the reader who knows them already, they can cause discouraging delays for some students if omitted. It is for this reason that we have tried to err on the side of over-explaining concepts, especially when these concepts appear in slightly altered forms. The more advanced reader can gloss over these details, but they are there for those who need them. We hope that more experienced mathematicians will forgive our repetitive justification of numerous facts throughout the text.

A glance at the Contents shows that we have covered those topics normally included in any linear algebra text although, as explained above, to a greater level of detail than other books. Where we differ significantly in content from most linear algebra texts however, is in our treatment of canonical forms (Chapter 8), tensors (Chapter 11), and infinite-dimensional vector spaces (Chapter 12). In particular, our treatment of the Jordan and rational canonical forms in Chapter 8 is based entirely on invariant factors and the

Smith normal form of a matrix. We feel this approach is well worth the effort required to learn it since the result is, at least conceptually, a constructive algorithm for computing the Jordan and rational forms of a matrix. However, later sections of the chapter tie together this approach with the more standard treatment in terms of cyclic subspaces. Chapter 11 presents the basic formalism of tensors as they are most commonly used by applied mathematicians, physicists and engineers. While most students first learn this material in a course on differential geometry, it is clear that virtually all the theory can be easily presented at this level, and the extension to differentiable manifolds then becomes only a technical exercise. Since this approach is all that most scientists ever need, we leave more general treatments to advanced courses on abstract algebra. Finally, Chapter 12 serves as an introduction to the theory of infinite-dimensional vector spaces. We felt it is desirable to give the student some idea of the problems associated with infinite-dimensional spaces and how they are to be handled. And in addition, physics students and others studying quantum mechanics should have some understanding of how linear operators and their adjoints are properly defined in a Hilbert space.

One major topic we have not treated at all is that of numerical methods. The main reason for this (other than that the book would have become too unwieldy) is that we feel at this level, the student who needs to know such techniques usually takes a separate course devoted entirely to the subject of numerical analysis. However, as a natural supplement to the present text, we suggest the very readable "Numerical Analysis" by I. Jacques and C. Judd (Chapman and Hall, 1987).

The problems in this text have been accumulated over 25 years of teaching the subject of linear algebra. The more of these problems that the students work the better. Be particularly wary of the attitude that assumes that some of these problems are "obvious" and need not be written out or precisely articulated. There are many surprises in the problems that will be missed from this approach! While these exercises are of varying degrees of difficulty, we have not distinguished any as being particularly difficult. However, the level of difficulty ranges from routine calculations that everyone reading this book should be able to complete, to some that will require a fair amount of thought from most students.

Because of the wide range of backgrounds, interests and goals of both students and instructors, there is little point in our recommending a particular
course outline based on this book. We prefer instead to leave it up to each teacher individually to decide exactly what material should be covered to meet the needs of the students. While at least portions of the first seven chapters should be read in order, the remaining chapters are essentially independent of each other. Those sections that are essentially applications of previous concepts, or else are not necessary for the rest of the book are denoted by an asterisk (*).

Now for one last comment on our notation. We use the symbol $\square$ to denote the end of a proof, and // to denote the end of an example. Sections are labeled in the format "Chapter.Section," and exercises are labeled in the format "Chapter.Section.Exercise." For example, Exercise 2.3.4 refers to Exercise 4 of Section 2.3, i.e., Section 3 of Chapter 2. Books listed in the bibliography are referred to by author and copyright date.

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## CHAPTER 6

## Polynomials

Before continuing with our treatment of linear operators and transformations, we must make a digression and consider the theory of polynomials in some detail. The subject matter of this chapter will be quite important throughout much of the remainder of this text. Our basic goal is to discuss the factorization of polynomials in detail, including many of the elementary properties that we all learned in high school.

### 6.1 DEFINITIONS

Let $\mathcal{F}$ be a field. In high school (or earlier), we all learned that a polynomial $\mathrm{p}(\mathrm{x})$ in the indeterminate (or variable) x is basically an expression of the form

$$
p(x)=a_{0}+a_{1} x+a_{2} x^{2}+\cdots+a_{n} x^{n}
$$

where n is any nonnegative integer and $\mathrm{a}_{0}, \ldots, \mathrm{a}_{\mathrm{n}}$ are all elements of $\mathcal{F}$. Note that our elementary experience with polynomials tells us that if

$$
\mathrm{q}(\mathrm{x})=\mathrm{b}_{0}+\mathrm{b}_{1} \mathrm{x}+\cdots+\mathrm{b}_{\mathrm{m}} \mathrm{x}^{\mathrm{m}}
$$

is another polynomial in $x$ then, assuming without loss of generality that $n \geq$ $m$, we have (where we define $a_{j}=0$ for $\mathrm{j}>\mathrm{n}$ and $\mathrm{b}_{\mathrm{j}}=0$ for $\mathrm{j}>\mathrm{m}$ )

$$
\mathrm{p}(\mathrm{x})+\mathrm{q}(\mathrm{x})=\left(\mathrm{a}_{0}+\mathrm{b}_{0}\right)+\left(\mathrm{a}_{1}+\mathrm{b}_{1}\right) \mathrm{x}+\cdots+\left(\mathrm{a}_{\mathrm{n}}+\mathrm{b}_{\mathrm{n}}\right) \mathrm{x}^{\mathrm{n}}
$$

and

$$
\begin{aligned}
\mathrm{p}(\mathrm{x}) \mathrm{q}(\mathrm{x})=\left(\mathrm{a}_{0} \mathrm{~b}_{0}\right) & +\left(\mathrm{a}_{0} \mathrm{~b}_{1}+\mathrm{a}_{1} \mathrm{~b}_{0}\right) \mathrm{x}+\left(\mathrm{a}_{0} \mathrm{~b}_{2}+\mathrm{a}_{1} \mathrm{~b}_{1}+\mathrm{a}_{2} \mathrm{~b}_{0}\right) \mathrm{x}^{2} \\
& +\cdots+\left(\mathrm{a}_{0} \mathrm{~b}_{\mathrm{k}}+\mathrm{a}_{1} \mathrm{~b}_{\mathrm{k}-1}+\cdots+\mathrm{a}_{\mathrm{k}-1} \mathrm{~b}_{1}+\mathrm{a}_{\mathrm{k}} \mathrm{~b}_{0}\right) \mathrm{x}^{\mathrm{k}} \\
& +\cdots+\mathrm{a}_{\mathrm{n}} \mathrm{~b}_{\mathrm{m}} \mathrm{x}^{\mathrm{n}+\mathrm{m}} .
\end{aligned}
$$

While this has a definite intuitive appeal (to previous experience), the term "expression" in the above definition of polynomial is rather nebulous, and it is worth making this definition somewhat more precise. To accomplish this, we focus our attention on the coefficients $\mathrm{a}_{\mathrm{i}}$.

We define a polynomial over $\mathcal{F}$ to be an (infinite) sequence of scalars

$$
\mathrm{p}=\left\{\mathrm{a}_{0}, \mathrm{a}_{1}, \mathrm{a}_{2}, \ldots\right\}
$$

such that $a_{n}=0$ for all but finitely many $n$. The scalars $a_{i}$ are called the coefficients of the polynomial. If

$$
\mathrm{q}=\left\{\mathrm{b}_{0}, \mathrm{~b}_{1}, \mathrm{~b}_{2}, \ldots\right\}
$$

is another polynomial in $\mathcal{F}$, then $p=q$ if and only if $a_{i}=b_{i}$ for every $i$. As we did for vector $n$-tuples, we define the addition of two polynomials $p$ and $q$ by

$$
\mathrm{p}+\mathrm{q}=\left\{\mathrm{a}_{0}+\mathrm{b}_{0}, \mathrm{a}_{1}+\mathrm{b}_{1}, \ldots\right\}
$$

Furthermore, we now also define the multiplication of $p$ and $q$ by

$$
\mathrm{pq}=\left\{\mathrm{c}_{0}, \mathrm{c}_{1}, \mathrm{c}_{2}, \ldots\right\}
$$

where

$$
c_{k}=\sum_{i+j=k} a_{i} b_{j}=\sum_{t=0}^{k} a_{t} b_{k-t}=a_{0} b_{k}+a_{1} b_{k-1}+\cdots+a_{k} b_{0} .
$$

Since $p$ and $q$ have a finite number of nonzero terms, so do both $p+q$ and $p q$, and hence both $\mathrm{p}+\mathrm{q}$ and pq are also polynomials.

We claim that the set of all polynomials over $\mathcal{F}$ forms a ring. Indeed, if we define the zero polynomial to be the sequence $\{0,0, \ldots\}$, and the negative of any polynomial $\left\{\mathrm{a}_{0}, \mathrm{a}_{1}, \ldots\right\}$ to be the polynomial $\left\{-\mathrm{a}_{0},-\mathrm{a}_{1}, \ldots\right\}$, then axioms (R1) - (R6) for a ring given in Section 1.4 are clearly satisfied. As to axiom (R7), let $p=\left\{\mathrm{a}_{0}, \mathrm{a}_{1}, \ldots\right\}, \mathrm{q}=\left\{\mathrm{b}_{0}, \mathrm{~b}_{1}, \ldots\right\}$ and $\mathrm{r}=\left\{\mathrm{c}_{0}, \mathrm{c}_{1}, \ldots\right\}$. Then the k th coefficient of $(\mathrm{pq}) \mathrm{r}$ is the sum (using the associative property of $\mathcal{F}$ )

$$
\begin{aligned}
\sum_{i+j=k}\left(\sum_{m+n=i} a_{m} b_{n}\right) c_{j} & =\sum_{m+n+j=k}\left(a_{m} b_{n}\right) c_{j}=\sum_{m+n+j=k} a_{m}\left(b_{n} c_{j}\right) \\
& =\sum_{m+i=k} a_{m}\left(\sum_{n+j=i} b_{n} c_{j}\right)
\end{aligned}
$$

But this last expression is just the $\mathrm{k} t h$ coefficient of $\mathrm{p}(\mathrm{qr})$. Finally, to prove axiom (R8), we use the distributive property of $\mathcal{F}$ to see that the $\mathrm{k} t h$ coefficient of $p(q+r)$ is

$$
\sum_{i+j=k} a_{i}\left(b_{j}+c_{j}\right)=\sum_{i+j=k} a_{i} b_{j}+\sum_{i+j=k} a_{i} c_{j} .
$$

Again we see that this last expression is just the $\mathrm{k} t h$ coefficient of $\mathrm{pq}+\mathrm{pr}$. Similarly, it is easy to see that $(\mathrm{p}+\mathrm{q}) \mathrm{r}=\mathrm{pr}+\mathrm{qr}$.

It should be clear that the ring of polynomials is commutative, and that if 1 is the unit element of $\mathcal{F}$, then $\{1,0,0, \ldots\}$ is a unit element for the ring of polynomials. However, since an arbitrary polynomial does not have a multiplicative inverse, the ring of polynomials does not form a field (see Theorem 6.2 , Corollary 3 below).

Example 6.1 Consider the polynomials

$$
\begin{aligned}
\mathrm{p} & =\{0,1,0,0, \ldots\} \\
\mathrm{q} & =\{1,2,-1,0, \ldots\}
\end{aligned}
$$

Then

$$
p+q=\{1,3,-1,0, \ldots\}
$$

and

$$
\begin{aligned}
\mathrm{pq}= & \{0(1), 0(2)+1(1), 0(-1)+1(2)+0(1), \\
& 0(0)+1(-1)+0(2)+0(1), \ldots\} \\
= & \{0,1,2,-1,0, \ldots\} . /
\end{aligned}
$$

Since the reader probably thought he (or she) already knew what a polynomial was, and since our definition may not be what it was he (or she) had in mind, what we will do now is relate our formal definition to our earlier elementary experience with polynomials. We will explain shortly why we are going through all of this apparently complicated formalism.

Given any element $\mathrm{a} \in \mathcal{F}$, we associate a polynomial $\mathrm{a}^{\prime}$ defined by

$$
a^{\prime}=\{a, 0,0, \ldots\} .
$$

This is clearly a one-to-one mapping of $\mathcal{F}$ into the set of all polynomials with coefficients in $\mathcal{F}$. We also note that if $\mathrm{a}, \mathrm{b} \in \mathcal{F}$, then $\mathrm{a}^{\prime}=\{\mathrm{a}, 0, \ldots\}$ and $\mathrm{b}^{\prime}=$ $\{b, 0, \ldots\}$ so that

$$
(a+b)^{\prime}=\{a+b, 0, \ldots\}=a^{\prime}+b^{\prime}
$$

and

$$
(\mathrm{ab})^{\prime}=\{\mathrm{ab}, 0, \ldots\}=\mathrm{a}^{\prime} \mathrm{b}^{\prime} .
$$

If $\mathcal{F}^{\prime}$ denotes the set of all polynomials $\mathrm{a}^{\prime}$ obtained in this manner, then $\mathcal{F}^{\prime}$ is a field isomorphic to $\mathcal{F}$. Because of this isomorphism, we shall identify the elements of $\mathcal{F}$ with their corresponding polynomials, and write $\mathrm{a}=\{\mathrm{a}, 0, \ldots\}$.

Now let the symbol $x$ denote the polynomial $\{0,1,0,0, \ldots\}$. We call the symbol $x$ an indeterminate. Applying our definition of polynomial multiplication, we see that $x^{2}=\{0,0,1,0, \ldots\}$ and, in general,

$$
\mathrm{x}^{\mathrm{n}}=\{0, \ldots, 0,1,0, \ldots\}
$$

where the 1 is in the n th position (remember that we start our numbering with 0 ). We also see that for any $\mathrm{a} \in \mathcal{F}$ we have (applying our multiplication rule)

$$
a^{n}=\{a, 0, \ldots\}\{0, \ldots, 1,0, \ldots\}=\{0, \ldots, a, 0, \ldots\}
$$

This means that an arbitrary polynomial $p=\left\{a_{0}, a_{1}, \ldots, a_{n}, 0, \ldots\right\}$ can be uniquely expressed in the familiar form

$$
\mathrm{p}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\mathrm{a}_{2} \mathrm{x}^{2}+\cdots+\mathrm{a}_{\mathrm{n}} \mathrm{x}^{\mathrm{n}}
$$

This discussion has now established a precise meaning to the term "expression" used at the beginning of this chapter. We will denote the commutative ring of all polynomials over $\mathcal{F}$ by $\mathcal{F}[\mathrm{x}]$. In summary, we see that that while a polynomial was actually defined as a sequence, we showed that any polynomial $\mathrm{p}=\left\{\mathrm{a}_{0}, \mathrm{a}_{1}, \ldots\right\} \in \mathcal{F}[\mathrm{x}]$ could be uniquely expressed in terms of the indeterminate x as

$$
p=a_{0}+a_{1} x+\cdots+a_{n} x^{n}=\sum_{i=1}^{n} a_{i} x^{i}
$$

Now suppose we are given both a polynomial $\mathrm{p}=\sum \mathrm{a}_{\mathrm{i}} \mathrm{X}^{\mathrm{i}} \in \mathcal{F}[\mathrm{x}]$ and any element $\mathrm{c} \in \mathcal{F}$. We define the element $\mathrm{p}(\mathrm{c}) \in \mathcal{F}$ by $\mathrm{p}(\mathrm{c})=\sum \mathrm{a}_{\mathrm{i}} \mathrm{c}^{\mathrm{i}}$. In other words, given a polynomial $\mathrm{p} \in \mathcal{F}[\mathrm{x}]$, the polynomial function $\mathrm{p}(\mathrm{x})$ is the mapping from $\mathcal{F}$ to $\mathcal{F}$ that takes $\mathrm{c} \in \mathcal{F}$ into the element $\mathrm{p}(\mathrm{c}) \in \mathcal{F}$. We call p (c)
the value of the polynomial $p$ when $c$ is substituted for $x$. Because of this, a polynomial $\mathrm{p} \in \mathcal{F}[\mathrm{x}]$ is frequently denoted by $\mathrm{p}(\mathrm{x})$.

The reason for this apparently complicated technical definition is that it is possible for two different polynomials in $\mathcal{F}[\mathrm{x}]$ to result in the same polynomial function (see Exercise 6.1.1).

Theorem 6.1 Suppose $\mathrm{p}, \mathrm{q} \in \mathcal{F}[\mathrm{x}]$ and $\mathrm{c} \in \mathcal{F}$. Then
(a) $(\mathrm{p} \pm \mathrm{q})(\mathrm{c})=\mathrm{p}(\mathrm{c}) \pm \mathrm{q}(\mathrm{c})$.
(b) $(\mathrm{pq})(\mathrm{c})=\mathrm{p}(\mathrm{c}) \mathrm{q}(\mathrm{c})$.

Proof (a) Writing $\mathrm{p}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{m}} \mathrm{x}^{\mathrm{m}}$ and $\mathrm{q}=\mathrm{b}_{0}+\mathrm{b}_{1} \mathrm{x}+\cdots+\mathrm{b}_{\mathrm{n}} \mathrm{x}^{\mathrm{n}}$ we have

$$
\begin{aligned}
(p \pm q)(c) & =\left(a_{0} \pm b_{0}\right)+\left(a_{1} \pm b_{1}\right) c+\left(a_{2} \pm b_{2}\right) c^{2}+\cdots \\
& =\left(a_{0}+a_{1} c+a_{2} c^{2}+\cdots\right) \pm\left(b_{0}+b_{1} c+b_{2} c^{2}+\cdots\right) \\
& =p(c) \pm q(c) .
\end{aligned}
$$

(b) Using p and q from part (a) and the definition of pq , we have

$$
\begin{aligned}
(p q)(c) & =a_{0} b_{0}+\left(a_{0} b_{1}+a_{1} b_{0}\right) c+\left(a_{0} b_{2}+a_{1} b_{1}+a_{2} b_{0}\right) c^{2}+\cdots \\
& =\left(a_{0}+a_{1} c+a_{2} c^{2}+\cdots\right)\left(b_{0}+b_{1} c+b_{2} c^{2}+\cdots\right) \\
& =p(c) q(c)
\end{aligned}
$$

It should now be clear that the definitions given above for the algebraic properties of polynomials in terms of sequences are just those that we all learned in high school for adding and multiplying polynomials together. It should be easy for the reader to show that Example 6.1 may be repeated in terms of our elementary notion of polynomial addition and multiplication.

If $p=a_{0}+a_{1} x+\cdots+a_{n} x^{n} \neq 0$ and $a_{n} \neq 0$, then we say that the degree of the polynomial p is n , and write we write $\operatorname{deg} \mathrm{p}=\mathrm{n}$. The term $\mathrm{a}_{\mathrm{n}}$ is called the leading coefficient of the polynomial, and if $a_{n}=1$, then the polynomial is said to be monic. If $\operatorname{deg} p=0$, then $p$ is said to be constant. By convention, the degree of the zero polynomial is not defined.

Theorem 6.2 Suppose $\mathrm{p}, \mathrm{q} \in \mathcal{F}[\mathrm{x}]$ are nonzero. Then
(a) $\operatorname{deg}(p+q) \leq \max \{\operatorname{deg} p, \operatorname{deg} q\} \quad($ where $p+q \neq 0)$.
(b) $\operatorname{deg}(\mathrm{pq})=\operatorname{deg} \mathrm{p}+\operatorname{deg} \mathrm{q}$.

Proof (a) Let $\mathrm{p}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{m}} \mathrm{x}^{\mathrm{m}}$ and $\mathrm{q}=\mathrm{b}_{0}+\mathrm{b}_{1} \mathrm{x}+\cdots+\mathrm{b}_{\mathrm{n}} \mathrm{x}^{\mathrm{n}}$ where $\mathrm{a}_{\mathrm{m}}, \mathrm{b}_{\mathrm{n}} \neq 0$. Then

$$
\mathrm{p}+\mathrm{q}=\left(\mathrm{a}_{0}+\mathrm{b}_{0}\right)+\left(\mathrm{a}_{1}+\mathrm{b}_{1}\right) \mathrm{x}+\cdots+\left(\mathrm{a}_{\mathrm{k}}+\mathrm{b}_{\mathrm{k}}\right) \mathrm{x}^{\mathrm{k}}
$$

where $\mathrm{k}=\max \{\mathrm{m}, \mathrm{n}\}$. Therefore $\operatorname{deg}(\mathrm{p}+\mathrm{q}) \leq \max \{\operatorname{deg} \mathrm{p}, \operatorname{deg} \mathrm{q}\}$ where the inequality follows since $a_{k}+b_{k}$ could equal 0 .
(b) From the definition of pq , we see that (using the same p and q from part (a)) the $k$ th term is $c_{k} x^{k}$ where $c_{k}=\sum_{i+j=k} a_{i} b_{j}$. But if $k>m+n$, then necessarily either $a_{i}$ or $b_{j}$ is zero, and therefore $c_{k}=0$ for $k>m+n$. Since $\mathcal{F}$ is a field and therefore also a division ring, it follows that $a_{m}, b_{n} \neq 0$ implies $a_{m} b_{n} \neq 0$ (if $a_{m} b_{n}=0$, then multiplying from the left by $a_{m}{ }^{-1}$ says that $b_{n}=0$, a contradiction). Thus $\operatorname{deg} \mathrm{pq}=\mathrm{m}+\mathrm{n}=\operatorname{deg} \mathrm{p}+\operatorname{deg} \mathrm{q}$.

Corollary 1 If $\mathrm{p}, \mathrm{q} \in \mathcal{F}[\mathrm{x}]$ are both nonzero, then $\operatorname{deg} \mathrm{p} \leq \operatorname{deg} \mathrm{pq}$.

Proof Since p and q are nonzero, $\operatorname{deg} \mathrm{p} \geq 0$ and $\operatorname{deg} \mathrm{q} \geq 0$, and hence $\operatorname{deg} \mathrm{p} \leq$ $\operatorname{deg} \mathrm{p}+\operatorname{deg} \mathrm{q}=\operatorname{deg} \mathrm{pq}$.

Corollary 2 If $\mathrm{p}, \mathrm{q}, \mathrm{r} \in \mathcal{F}[\mathrm{x}]$, then
(a) $\mathrm{pq}=0$ implies that either $\mathrm{p}=0$ or $\mathrm{q}=0$.
(b) If $\mathrm{pq}=\mathrm{rq}$ where $\mathrm{q} \neq 0$, then $\mathrm{p}=\mathrm{r}$.

Proof (a) If $\mathrm{p} \neq 0$ and $\mathrm{q} \neq 0$, then deg $\mathrm{pq} \geq 0$ implies that $\mathrm{pq} \neq 0$.
(b) Since $\mathcal{F}[x]$ is a ring, we see that if $p q=r q$, then $(p-r) q=0$. But $q \neq 0$, so that by (a) we must have $\mathrm{p}-\mathrm{r}=0$, or $\mathrm{p}=\mathrm{r}$.

We note that part (a) of this corollary shows that $\mathcal{F}[x]$ has no zero divisors, and hence $\mathcal{F}[x]$ forms an integral domain (see Section 1.5).

Corollary 3 Let $\mathrm{p} \neq 0$ be an element of $\mathcal{F}[\mathrm{x}]$. Then there exists $\mathrm{q} \in \mathcal{F}[\mathrm{x}]$ such that $\mathrm{pq}=1$ if and only if $\operatorname{deg} \mathrm{p}=0$, and hence $\mathcal{F}[\mathrm{x}]$ is not a field.

Proof If $\operatorname{deg} \mathrm{p}=0$ then $\mathrm{p}=\mathrm{a} \in \mathcal{F}$ with $\mathrm{a} \neq 0$, and thus there exists $\mathrm{q}=\mathrm{a}^{-1} \in$ $\mathcal{F}$ (hence $\mathrm{q}=\mathrm{a}^{-1} \in \mathcal{F}[\mathrm{x}]$ ) such that $\mathrm{pq}=\mathrm{aa}^{-1}=1$. On the other hand, if $\mathrm{pq}=1$ we have

$$
0=\operatorname{deg} 1=\operatorname{deg} p q=\operatorname{deg} p+\operatorname{deg} q .
$$

But $p \neq 0$ implies that $\operatorname{deg} p \geq 0$, and if $q \neq 0$, we must also have $\operatorname{deg} q \geq 0$. It then follows that $\operatorname{deg} \mathrm{p}=0$.

We now prove a very useful and fundamental result known as the division algorithm. Essentially, we will show the existence of the polynomial quotient $\mathrm{f} / \mathrm{g}$ (although technically this symbol as of yet has no meaning). The polyno-
mials q and r defined in the following theorem are called the quotient and remainder respectively. After the proof, we will give an example which shows that this is just the "long division" we all learned in elementary school.

Theorem 6.3 (Division Algorithm) Given $\mathrm{f}, \mathrm{g} \in \mathcal{F}[\mathrm{x}]$ with $\mathrm{g} \neq 0$, there exist unique polynomials $\mathrm{q}, \mathrm{r} \in \mathcal{F}[\mathrm{x}]$ such that

$$
\mathrm{f}=\mathrm{qg}+\mathrm{r}
$$

where either $\mathrm{r}=0$ or $\operatorname{deg} \mathrm{r}<\operatorname{deg} \mathrm{g}$.
Proof The basic idea of the proof is to consider all possible degrees for the polynomials f and g , and show that the theorem can be satisfied in each case. After proving the existence of the polynomials $q$ and $r$, we shall prove their uniqueness.

If $\mathrm{f}=0$ we simply choose $\mathrm{q}=\mathrm{r}=0$. Now suppose that

$$
\begin{aligned}
& f=a_{0}+a_{1} x+\cdots+a_{m} x^{m} \\
& g=b_{0}+b_{1} x+\cdots+b_{n} x^{n}
\end{aligned}
$$

where $\mathrm{a}_{\mathrm{m}}, \mathrm{b}_{\mathrm{n}} \neq 0$. If $\mathrm{m}=\mathrm{n}=0$, then by Corollary 3 of Theorem 6.2 , there exists $g^{-1} \in \mathcal{F}[x]$ such that $g^{-1} g=1$, and therefore $f=f\left(g^{-1} g\right)=\left(f^{-1}\right) g+0$ satisfies our requirements. Next, if $m=0$ and $n>0$, then we may write $f=$ $0 \mathrm{~g}+\mathrm{f}$ with $\operatorname{deg} \mathrm{r}=\operatorname{deg} \mathrm{f}<\operatorname{deg} \mathrm{g}$.

We now assume that $\mathrm{m}>0$ and proceed by induction on m . In other words, we assume that $q$ and $r$ can be found for all polynomials $f$ with $\operatorname{deg} f \leq$ $m-1$ and proceed to construct new polynomials $q$ and $r$ for $\operatorname{deg} f=m$. First note that if $\mathrm{n}>\mathrm{m}$ we may again take $\mathrm{f}=0 \mathrm{~g}+\mathrm{f}$ to satisfy the theorem. Thus we need only consider the case of $n \leq m$.

Define the polynomial

$$
\mathrm{f}_{1}=\mathrm{f}-\left(\mathrm{a}_{\mathrm{m}} / \mathrm{b}_{\mathrm{n}}\right) \mathrm{x}^{\mathrm{m}-\mathrm{n}} \mathrm{~g} .
$$

Then the coefficient of the $x^{m}$ term in $f_{1}$ is 0 (it cancels out on the right hand side), and hence $\operatorname{deg} \mathrm{f}_{1} \leq \mathrm{m}-1$. Therefore, by our induction hypothesis, there exist polynomials $q_{1}$ and $r_{1}$ in $\mathcal{F}[x]$ with either $r_{1}=0$ or $\operatorname{deg} r_{1}<\operatorname{deg} g$ such that $f_{1}=q_{1} g+r_{1}$. Substituting the definition of $f_{1}$ in this equation yields

$$
\mathrm{f}=\left[\left(\mathrm{a}_{\mathrm{m}} / \mathrm{b}_{\mathrm{n}}\right) \mathrm{x}^{\mathrm{m}-\mathrm{n}}+\mathrm{q}_{1}\right] \mathrm{g}+\mathrm{r}_{1} .
$$

If we define $r=r_{1}$ and $q=\left(a_{m} / b_{n}\right) x^{m-n}+q_{1}$, we see that $f=q g+r$ where either $r=0$ or $\operatorname{deg} r<\operatorname{deg} g$. This proves the existence of the polynomials $q$ and r , and all that remains is to prove their uniqueness.

Suppose that

$$
\mathrm{f}=\mathrm{qg}+\mathrm{r}=\mathrm{q}^{\prime} \mathrm{g}+\mathrm{r}^{\prime}
$$

where both $r$ and $r^{\prime}$ satisfy the theorem, and assume that $r \neq r^{\prime}$. Then

$$
r-r^{\prime}=\left(q^{\prime}-q\right) g \neq 0
$$

where

$$
\operatorname{deg}\left(r-r^{\prime}\right)<\operatorname{deg} g
$$

by Theorem 6.2(a). On the other hand, from Theorem 6.2(b) we see that

$$
\operatorname{deg}\left(r-r^{\prime}\right)=\operatorname{deg}\left[\left(q^{\prime}-q\right) g\right]=\operatorname{deg}\left(q^{\prime}-q\right)+\operatorname{deg} g \geq \operatorname{deg} g .
$$

This contradiction shows that in fact $r^{\prime}=r$. We now have $\left(q^{\prime}-q\right) g=0$ with $\mathrm{g} \neq 0$, and hence by Corollary 2(a) of Theorem 6.2, we have $\mathrm{q}^{\prime}-\mathrm{q}=0$, and therefore $\mathrm{q}^{\prime}=\mathrm{q}$.

Let us give an example of the division algorithm that should clarify what was done in the theorem.

Example 6.2 Consider the polynomials

$$
\begin{aligned}
& f=2 x^{4}+x^{2}-x+1 \\
& g=2 x-1 .
\end{aligned}
$$

Following the proof of Theorem 6.3 we have

$$
\mathrm{f}_{1}=\mathrm{f}-\mathrm{x}^{3} \mathrm{~g}=\mathrm{x}^{3}+\mathrm{x}^{2}-\mathrm{x}+1
$$

Now let

$$
\mathrm{f}_{2}=\mathrm{f}_{1}-(1 / 2) \mathrm{x}^{2} \mathrm{~g}=(3 / 2) \mathrm{x}^{2}-\mathrm{x}+1
$$

Again, we let

$$
\mathrm{f}_{3}=\mathrm{f}_{2}-(3 / 4) \mathrm{xg}=(-1 / 4) \mathrm{x}+1
$$

so that

$$
\mathrm{f}_{4}=\mathrm{f}_{3}+(1 / 8) \mathrm{g}=7 / 8
$$

Since $\operatorname{deg}(7 / 8)<\operatorname{deg} \mathrm{g}$, we are finished with the division. Combining the above polynomials we see that

$$
\mathrm{f}=\left[\mathrm{x}^{3}+(1 / 2) \mathrm{x}^{2}+(3 / 4) \mathrm{x}-(1 / 8)\right] \mathrm{g}+\mathrm{f}_{4}
$$

and therefore

$$
\begin{aligned}
& q=x^{3}+(1 / 2) x^{2}+(3 / 4) x-(1 / 8) \\
& r=7 / 8
\end{aligned}
$$

This may also be written out in a more familiar form as

\[

\]

It should be noted that at each step in the division, we eliminated the highest remaining power of f by subtracting the appropriate multiple of g . //

## Exercises

1. Let $\mathcal{F}=\{0,1\}$ be the field consisting of only two elements, and define addition and multiplication on these elements in the obvious way (see Exercise 1.5.17). Show that the distinct polynomials $x^{2}-x$ and 0 define the same polynomial function.
2. Use the division algorithm to find the quotient and remainder when $f=$ $2 x^{4}-x^{3}+x-1 \in \mathbb{R}[x]$ is divided by $g=3 x^{3}-x^{2}+3 \in \mathbb{R}[x]$.
3. Consider the polynomials $\mathrm{p}=\{2,0,1,1\}$ and $\mathrm{q}=\{1,1,-1\}$ over $\mathbb{R}$. Evaluate the product pq by applying the definition. Show that this yields the same result as directly multiplying together the polynomial functions $p$ and $q$.
4. Given a polynomial $p=a_{n} x^{n}+\cdots+a_{1} x+a_{0}$, we define its formal derivative to be the polynomial $D p=n a_{n} x^{n-1}+\cdots+2 a_{2} x+a_{1}$. In other words, $\mathrm{D}: \mathcal{F}[\mathrm{x}] \rightarrow \mathcal{F}[\mathrm{x}]$ is a differentiation operator. Prove $\mathrm{D}(\mathrm{p}+\mathrm{q})=$ $\mathrm{Dp}+\mathrm{Dq}$ and $\mathrm{D}(\mathrm{pq})=\mathrm{p}(\mathrm{Dq})+(\mathrm{Dp}) \mathrm{q}$.
5. Find the remainder when $i \mathrm{x}^{9}+3 \mathrm{x}^{7}+\mathrm{x}^{6}-2 i \mathrm{x}+1 \in \mathbb{C}[\mathrm{x}]$ is divided by $\mathrm{x}+i \in \mathbb{C}[\mathrm{x}]$.

### 6.2 FACTORIZATION OF POLYNOMIALS

If $\mathrm{f}(\mathrm{x})$ is a polynomial in $\mathcal{F}[\mathrm{x}]$, then $\mathrm{c} \in \mathcal{F}$ is said to be a zero (or root) of f if $\mathrm{f}(\mathrm{c})=0$. We shall also sometimes say that c is a solution of the polynomial equation $f(x)=0$. We will see that information about the roots of a polynomial plays an extremely important role throughout much of the remainder of this text.

If $\mathrm{f}, \mathrm{g} \in \mathcal{F}[\mathrm{x}]$ and $\mathrm{g} \neq 0$, then we say that f is divisible by g (or g divides f ) over $\mathcal{F}$ if $\mathrm{f}=\mathrm{qg}$ for some $\mathrm{q} \in \mathcal{F}[\mathrm{x}]$. In other words, f is divisible by g if the remainder in the division of $f$ by $g$ is zero. In this case we also say that $g$ is a factor of f (over $\mathcal{F}$ ). It is standard notation to write $\mathrm{g} \mid \mathrm{f}$ when we wish to say that $g$ divides $f$, or to write $g \nmid f$ when $g$ does not divide $f$.

The next theorem is known as the remainder theorem, and its corollary is known as the factor theorem.

Theorem 6.4 (Remainder Theorem) Suppose $\mathrm{f} \in \mathcal{F}[\mathrm{x}]$ and $\mathrm{c} \in \mathcal{F}$. Then the remainder in the division of $f$ by $x-c$ is $f(c)$. In other words,

$$
\mathrm{f}(\mathrm{x})=(\mathrm{x}-\mathrm{c}) \mathrm{q}+\mathrm{f}(\mathrm{c})
$$

Proof We see from the division algorithm that $\mathrm{f}=(\mathrm{x}-\mathrm{c}) \mathrm{q}+\mathrm{r}$ where either $r=0$ or $\operatorname{deg} r<\operatorname{deg}(x-c)=1$, and hence either $r=0$ or $\operatorname{deg} r=0$ (in which case $r \in \mathcal{F}$ ). In either case, we may substitute c for x to obtain

$$
\mathrm{f}(\mathrm{c})=(\mathrm{c}-\mathrm{c}) \mathrm{q}(\mathrm{c})+\mathrm{r}=\mathrm{r} .
$$

Corollary (Factor Theorem) If $\mathrm{f} \in \mathcal{F}[\mathrm{x}]$ and $\mathrm{c} \in \mathcal{F}$, then $\mathrm{x}-\mathrm{c}$ is a factor of $f$ if and only if $f(c)=0$.

Proof Rephrasing the statement of the corollary as $\mathrm{f}=\mathrm{q}(\mathrm{x}-\mathrm{c})$ if and only if $\mathrm{f}(\mathrm{c})=0$, it is clear that this follows directly from the theorem.

Example 6.3 If we divide $\mathrm{f}=\mathrm{x}^{3}-5 \mathrm{x}^{2}+7 \mathrm{x}$ by $\mathrm{g}=\mathrm{x}-2$, we obtain $\mathrm{q}=\mathrm{x}^{2}-$ $3 x+1$ and $r=2$. It is also easy see that $f(2)=8-5(4)+7(2)=2$ as it should according to Theorem 6.4. //

Let R be a commutative ring with unit element. An element $\mathrm{u} \in \mathrm{R}$ is called a unit (not a unit element) if there exists $r \in R$ such that ur $=1$. Other ways to say this are that $u$ divides 1 , or that a unit is an element whose inverse is also in the ring. We leave it to the reader to show that $u \in R$ is a unit if and only if it is a factor of every element of R (see Exercise 6.2.1). An element $p \in R$ that is neither zero nor a unit is said to be prime if $p=a b$ implies that either a or b is a unit. Thus a prime element is one that can not be factored in a nontrivial way.

Example 6.4 Since for any integer $\mathrm{n} \neq \pm 1$ the number $1 / \mathrm{n}$ is not an integer, it should be clear that the ring of integers $\mathbb{Z}$ has only the units 1 and -1 . On the other hand, if $\mathcal{F}$ is any field and $\mathrm{a} \in \mathcal{F}$, then $\mathrm{a}^{-1}$ is also a member of $\mathcal{F}$, and hence any nonzero element of a field is a unit. In particular, the units of the ring $\mathcal{F}[\mathrm{x}]$ are just the polynomials of degree zero (i.e., the nonzero constant polynomials).

If we consider the ring of integers $\mathbb{Z}$, then a number $p \in \mathbb{Z}$ with $p \neq \pm 1$ or 0 will be prime if the only divisors of $p$ are $\pm 1$ and $\pm p$. However, if we consider the field $\mathbb{R}$, then any nonzero element of $\mathbb{R}$ (i.e., any nonzero real number) is a unit, and hence the notion of a prime real number is not very useful. //

In the particular case of $\mathrm{R}=\mathcal{F}[\mathrm{x}]$, a prime polynomial is frequently called an irreducible polynomial. A polynomial that is not irreducible is said to be reducible. Two polynomials $f, g \in \mathcal{F}[x]$ are said to be associates if $\mathrm{f}=\mathrm{cg}$ for some nonzero $\mathrm{c} \in \mathcal{F}$, and in general, two elements $\mathrm{a}, \mathrm{b} \in \mathrm{R}$ are said to be associates if $a=u b$ where $u$ is a unit of $R$. We leave it to the reader to show that this defines an equivalence relation on a commutative ring with unit element (Exercise 6.2.3).

It should be clear that any nonzero polynomial has exactly one monic polynomial as an associate since we can always write

$$
a_{0}+a_{1} x+\cdots+a_{n} x^{n}=a_{n}\left(a_{0} a_{n}{ }^{-1}+a_{1} a_{n}{ }^{-1} x+\cdots+x^{n}\right) .
$$

It should also be clear that any polynomial $f$ with $\operatorname{deg} \mathrm{f} \geq 1$ has its associates and the set of nonzero constant polynomials as divisors. Thus we see that a nonzero polynomial $f$ is prime if and only if $f=g h$ implies that either $g$ or $h$ is of degree zero, and hence the other is an associate of $f$.

It is important to realize that whether or not a polynomial is prime depends on the particular field $\mathcal{F}$. For example, since $x^{2}-2=(x-\sqrt{2})(x+\sqrt{2})$, we see that $x^{2}-2$ is prime over $\mathbb{Q}$, but is not prime over $\mathbb{R}$.

Returning to our commutative ring R with unit element, we say that an element $d \in R$ is a greatest common divisor (frequently denoted simply by gcd ) of the elements $\mathrm{a}_{1}, \ldots, \mathrm{a}_{\mathrm{n}} \in \mathrm{R}$ if
(1) $\mathrm{d} \mid \mathrm{a}_{\mathrm{i}}$ for every $\mathrm{i}=1, \ldots, \mathrm{n}$ (i.e., d is a common divisor of the $\mathrm{a}_{\mathrm{i}}$ );
(2) If $\mathrm{c} \in \mathrm{R}$ is such that $\mathrm{c} \mid \mathrm{a}_{\mathrm{i}}$ for every $\mathrm{i}=1, \ldots, \mathrm{n}$, then $\mathrm{c} \mid \mathrm{d}$.

Two distinct elements $\mathrm{a}, \mathrm{b} \in \mathrm{R}$ are said to be relatively prime if their greatest common divisor is a unit of R. Note that we have referred to $a$ greatest common divisor, implying that there may be more than one. This is only true in a certain sense as the next theorem shows.

Theorem 6.5 Let $\mathrm{f}_{1}, \ldots, \mathrm{f}_{\mathrm{n}}$ be nonzero polynomials in $\mathcal{F}[\mathrm{x}]$. Then there exists at least one greatest common divisor d of the set $\left\{\mathrm{f}_{1}, \ldots, \mathrm{f}_{\mathrm{n}}\right\}$. Moreover, this greatest common divisor is unique up to a unit factor, and can be expressed in the form $\mathrm{d}=\sum_{\mathrm{i}=1}^{\mathrm{n}} \mathrm{h}_{\mathrm{i}} \mathrm{f}_{\mathrm{i}}$ for some set of polynomials $\mathrm{h}_{\mathrm{i}} \in \mathcal{F}[\mathrm{x}]$.

Proof Consider the set S of all polynomials in $\mathcal{F}[\mathrm{x}]$ of the form $\mathrm{g}_{1} \mathrm{f}_{1}+\cdots+$ $\mathrm{g}_{\mathrm{n}} \mathrm{f}_{\mathrm{n}}$ where each $\mathrm{g}_{\mathrm{i}}$ is an arbitrary element of $\mathcal{F}[\mathrm{x}]$. Then in particular, each $\mathrm{f}_{\mathrm{i}}$ is an element of $S$, and in addition, if $p \in S$ and $q \in \mathcal{F}[x]$, then $p q \in S$. Let $D$ be the set of degrees of all nonzero polynomials in S . Then D is just a collection of nonnegative integers, and hence by the well-ordering principle (Section 0.5 ), D has a least element $\alpha$. This means that there exists a nonzero polynomial $d=h_{1} f_{1}+\cdots+h_{n} f_{n} \in S$ such that $\alpha=\operatorname{deg} d \leq \operatorname{deg} c$ for all nonzero polynomials $c \in S$. We first show that $d \mid f_{i}$ for every $i=1, \ldots, n$.

By the division algorithm, for each $i=1, \ldots, n$ we have $f_{i}=q_{i} d+r_{i}$ where either $r_{i}=0$ or deg $r_{i}<\operatorname{deg} d$. But $r_{i}=f_{i}-q_{i} d \in S$ so that if $r_{i} \neq 0$, we would have deg $r_{i}<\operatorname{deg} d$ which contradicts the definition of $d$. Therefore we must have $r_{i}=0$ and hence $f_{i}=q_{i} d$ so that $d \mid f_{i}$ for every $i=1, \ldots, n$. While this shows that $d$ is a common divisor of $\left\{\mathrm{f}_{\mathrm{i}}\right\}$, we must show that it is in fact a greatest common divisor.

If $c$ is any other common divisor of $\left\{f_{1}, \ldots, f_{n}\right\}$, then by definition there exist polynomials $g_{i}$ such that $f_{i}=g_{i} c$ for each $i=1, \ldots, n$. But then

$$
\mathrm{d}=\sum \mathrm{h}_{\mathrm{i}} \mathrm{f}_{\mathrm{i}}=\sum \mathrm{h}_{\mathrm{i}}\left(\mathrm{~g}_{\mathrm{i}} \mathrm{c}\right)=\left(\sum \mathrm{h}_{\mathrm{i}} \mathrm{~g}_{\mathrm{i}}\right) \mathrm{c}
$$

so that $\mathrm{c} \mid \mathrm{d}$. This proves that d is a greatest common divisor.
Now suppose $d^{\prime}$ is another greatest common divisor of the set $\left\{\mathrm{f}_{1}, \ldots, \mathrm{f}_{\mathrm{n}}\right\}$. Then by definition of greatest common divisor we must have both d|d' and $\mathrm{d}^{\prime} \mid \mathrm{d}$, so that $\mathrm{d}^{\prime}=\mathrm{ud}$ and $\mathrm{d}=\mathrm{vd}^{\prime}$ for some polynomials $\mathrm{u}, \mathrm{v} \in \mathcal{F}[\mathrm{x}]$. Multiplying the second of these equations by $u$, we see that $u v d^{\prime}=u d=d^{\prime}$, and therefore $\mathrm{d}^{\prime}(1-u v)=0$. By Corollary 2(a) of Theorem 6.2, we then have $1-u v=0$ so that $u v=1$ and hence $u$ and $v$ are units. (Alternatively, the fact that $\operatorname{deg} d=$ deg $\mathrm{d}^{\prime}$ implies that u and v must be of degree zero, and hence are units.)

What we have shown is that a gcd exists, and is unique up to its associates. Therefore, if we restrict ourselves to monic greatest common divisors, then we have proved the existence of a unique gcd.

Corollary 1 Let $\mathrm{q}_{1}, \ldots, \mathrm{q}_{\mathrm{n}} \in \mathcal{F}[\mathrm{x}]$ be relatively prime (i.e., they have no common divisors other than units). Then there exist elements $h_{1}, \ldots, h_{n} \in$ $\mathcal{F}[\mathrm{x}]$ such that $\mathrm{h}_{1} \mathrm{q}_{1}+\cdots+\mathrm{h}_{\mathrm{n}} \mathrm{q}_{\mathrm{n}}=1$.

Proof This is an obvious special case of Theorem 6.5.

Corollary 2 Suppose $\mathrm{f}, \mathrm{g}, \mathrm{p} \in \mathcal{F}[\mathrm{x}]$ where p is prime and $\mathrm{p} \mid \mathrm{fg}$. Then either $\mathrm{p} \mid \mathrm{f}$ or plg.

Proof Since p is prime, its only divisors are units and its associates. Therefore, if we assume that $\mathrm{p} \backslash \mathrm{f}$, then the only greatest common divisor of p and f is a unit, and thus p and f are relatively prime. Applying Theorem 6.5, we may write $u p+v f=1$ for some $u, v \in \mathcal{F}[x]$, and hence multiplying by $g$ yields pug $+\mathrm{fgv}=\mathrm{g}$. But $\mathrm{p} \mid \mathrm{fg}$ so that $\mathrm{fg}=\mathrm{qp}$ for some $\mathrm{q} \in \mathcal{F}[\mathrm{x}]$, and thus we have $p(u g+q v)=g$ so that $p \mid g$. It is obvious that had we started with the assumption that $\mathrm{p} \nmid \mathrm{g}$, we would have found that $\mathrm{p} \mid \mathrm{f}$.

By choosing $\mathrm{f}=\mathrm{f}_{1}$ and $\mathrm{g}=\mathrm{f}_{2} \cdots \mathrm{f}_{\mathrm{n}}$ in Corollary 2 we have the following obvious generalization.

Corollary $2^{\prime}$ If $p, \mathrm{f}_{1}, \mathrm{f}_{2}, \ldots, \mathrm{f}_{\mathrm{n}} \in \mathcal{F}[\mathrm{x}]$ where p is prime and $\mathrm{p} \mid \mathrm{f}_{1} \mathrm{f}_{2} \cdots \mathrm{f}_{\mathrm{n}}$, then $p \mid f_{i}$ for some $\mathrm{i}=1, \ldots, \mathrm{n}$.

While Theorem 6.5 proves the existence of a greatest common divisor, it is not of any help in actually computing one. Given two polynomials, we can find their gcd by a procedure known as the Euclidean algorithm (compare

Section 0.7). This approach, illustrated in the next example, is also an alternative proof of Theorem 6.5, the general case as stated in the theorem following by induction.

Example 6.5 (Euclidean algorithm) Suppose $\mathrm{f}, \mathrm{g} \in \mathcal{F}[\mathrm{x}]$ and $\mathrm{f} \neq 0$. We show the existence of a unique monic polynomial $d \in \mathcal{F}[x]$ such that
(1) $\mathrm{d} \mid \mathrm{f}$ and $\mathrm{d} \mid \mathrm{g}$.
(2) If $\mathrm{c} \in \mathcal{F}[\mathrm{x}]$ is such that $\mathrm{c} \mid \mathrm{f}$ and $\mathrm{c} \mid \mathrm{g}$, then $\mathrm{c} \mid \mathrm{d}$.

First note that if $g=0$ and $a_{m}$ is the leading coefficient of $f$, then the monic polynomial $d=a_{m}{ }^{-1} f$ satisfies both requirements (1) and (2). Now assume that $\mathrm{g} \neq 0$ also. By the division algorithm, there exist unique polynomials $\mathrm{q}_{1}$ and $\mathrm{r}_{1}$ such that

$$
\mathrm{f}=\mathrm{gq}_{1}+\mathrm{r}_{1}
$$

with either $\mathrm{r}_{1}=0$ or $\operatorname{deg} \mathrm{r}_{1}<\operatorname{deg} \mathrm{g}$. If $\mathrm{r}_{1}=0$, then $\mathrm{g} \mid \mathrm{f}$ and $\mathrm{d}=\mathrm{g}$ satisfies (1) and (2). If $r_{1} \neq 0$, then we apply the division algorithm again to obtain polynomials $\mathrm{q}_{2}$ and $\mathrm{r}_{2}$ such that

$$
\mathrm{g}=\mathrm{r}_{1} \mathrm{q}_{2}+\mathrm{r}_{2} .
$$

If $r_{2}=0$, then $r_{1} \mid g$ which implies that $r_{1} \mid f$, and thus $d=r_{1}$ is a common divisor. (It still remains to be shown that this $d$ is a greatest common divisor.) If $r_{2} \neq$ 0 , then we continue the process, thus obtaining the following progression:

$$
\begin{array}{rc}
f & =g q_{1}+r_{1} \\
g= & \operatorname{deg} r_{1}<\operatorname{deg} g \\
r_{2}+r_{2} & \operatorname{deg} r_{2}<\operatorname{deg} r_{1} \\
r_{2} q_{3}+r_{3} & \operatorname{deg} r_{3}<\operatorname{deg} r_{2} \\
\vdots & \vdots \\
r_{k-2} & =r_{k-1} q_{k}+r_{k} \\
r_{k-1} & =r_{k} q_{k+1}
\end{array}
$$

This progression must terminate as shown since the degree of any polynomial is a positive integer and $\operatorname{deg} r_{1}>\operatorname{deg} r_{2}>\cdots>\operatorname{deg} r_{k} \geq 0$. Letting $r_{k}$ be the last nonzero remainder, we claim that $\mathrm{r}_{\mathrm{k}}=\mathrm{d}$.

To see this, first note that $r_{k} \mid r_{k-1}$ since $r_{k-1}=r_{k} q_{k+1}$. Next, we see that

$$
r_{k-2}=r_{k-1} q_{k}+r_{k}=r_{k} q_{k+1} q_{k}+r_{k}
$$

and therefore $r_{k} \mid r_{k-2}$. Continuing this procedure, we find that $r_{k}\left|r_{k-1}, r_{k}\right| r_{k-2}$, $\ldots, r_{k}\left|r_{1}, r_{k}\right| g$ and finally $r_{k} \mid f$. This shows that $d=r_{k}$ satisfies (1). Now suppose that $\mathrm{c} \mid \mathrm{f}$ and $\mathrm{c} \mid \mathrm{g}$. Then, since $\mathrm{r}_{1}=\mathrm{f}-\mathrm{gq}_{1}$ we must have $\mathrm{c} \mid \mathrm{r}_{1}$. Similarly $\mathrm{r}_{2}=$ $\mathrm{g}-\mathrm{r}_{1} \mathrm{q}_{2}$ so that $\mathrm{c} \mid \mathrm{r}_{2}$. Continuing this process, it is clear that we eventually
arrive at the conclusion that $c \mid r_{k}$, thus proving (2) for the choice $d=r_{k}$. Finally, if $r$ is the leading coefficient of $r_{k}$, then $r^{-1} r_{k}$ is a monic polynomial satisfying (1) and (2), and its uniqueness follows exactly as in the proof of Theorem 6.5. //

Example 6.6 As a specific illustration of the preceding example, consider the polynomials $f=x^{4}-x^{3}-x^{2}+1$ and $g=x^{3}-1$ over the field $\mathbb{Q}$. Dividing f by $g$ we obtain

$$
x^{4}-x^{3}-x^{2}+1=\left(x^{3}-1\right)(x-1)+\left(-x^{2}+x\right)
$$

Now divide $g$ by $r_{1}=-x^{2}+x$ to obtain

$$
x^{3}-1=\left(-x^{2}+x\right)(-x-1)+(x-1) .
$$

Lastly, we divide $r_{1}$ by $r_{2}=x-1$ to find

$$
-x^{2}+x=(x-1)(-x)
$$

and therefore the gcd of $f$ and $g$ is $x-1$. //
Our next very important result is known as the unique factorization theorem. Recall that by definition, a prime polynomial is not a unit, and thus has positive degree.

Theorem 6.6 (Unique Factorization Theorem) Every nonzero element $\mathrm{f} \in$ $\mathcal{F}[\mathrm{x}]$ is either a unit, or is expressible as a unique (up to associates) finite product of prime elements.

Proof We first show that $\mathrm{f} \in \mathcal{F}[\mathrm{x}]$ is expressible as a product of prime polynomials. Afterwards we will prove uniqueness. Our approach is by induction on $\operatorname{deg} \mathrm{f}$; in other words, we assume that $\operatorname{deg} \mathrm{f}>1$ (if $\operatorname{deg} \mathrm{f}=0$ then f is a unit, and if $\operatorname{deg} \mathrm{f}=1$ the theorem is obvious), and suppose that the theorem is true for all $\mathrm{g} \in \mathcal{F}[\mathrm{x}]$ with deg $\mathrm{g}<\operatorname{deg} \mathrm{f}$. We will show that the theorem is true for f.

Assume f is reducible (or else there is nothing to prove) so that $\mathrm{f}=\mathrm{pq}$ where neither p nor q is a unit. By Theorem 6.2(b) we have $\operatorname{deg} \mathrm{p}<\operatorname{deg} \mathrm{p}+$ $\operatorname{deg} \mathrm{q}=\operatorname{deg} \mathrm{f}$, and similarly $\operatorname{deg} \mathrm{q}<\operatorname{deg} \mathrm{f}$. Therefore, by our induction hypothesis, both p and q can be written as a finite product of prime elements in $\mathcal{F}[\mathrm{x}]$, and hence the same is true of $\mathrm{f}=\mathrm{pq}$.

To prove the uniqueness of the product, assume that

$$
\mathrm{f}=\mathrm{p}_{1} \mathrm{p}_{2} \cdots \mathrm{p}_{\mathrm{n}}=\mathrm{q}_{1} \mathrm{q}_{2} \cdots \mathrm{q}_{\mathrm{m}}
$$

where each of the $p_{i}$ and $q_{j}$ are prime. Since $p_{1} \mid p_{1} p_{2} \cdots p_{n}$, it follows that $p_{1} \mid q_{1} q_{2} \cdots q_{m}$. By Corollary $2^{\prime}$ of Theorem 6.5, it then follows that $p_{1} \mid q_{j}$ for some $j=1, \ldots, m$. But since both $p_{1}$ and $q_{j}$ are prime and $p_{1} \mid q_{j}$, they must be associates, and hence $q_{j}=u_{1} p_{1}$ where $u_{1}$ is a unit in $\mathcal{F}[x]$. This means that

$$
\mathrm{p}_{1} \mathrm{p}_{2} \cdots \mathrm{p}_{\mathrm{n}}=\mathrm{q}_{1} \mathrm{q}_{2} \cdots \mathrm{q}_{\mathrm{m}}=\mathrm{q}_{1} \mathrm{q}_{2} \cdots \mathrm{q}_{\mathrm{j}-1} \mathrm{u}_{1} \mathrm{p}_{1} \mathrm{q}_{\mathrm{j}+1} \cdots \mathrm{q}_{\mathrm{m}} .
$$

Cancelling $\mathrm{p}_{1}$ from both sides of this equation (using Theorem 6.2, Corollary 2) results in

$$
\mathrm{p}_{2} \cdots \mathrm{p}_{\mathrm{n}}=\mathrm{u}_{1} \mathrm{q}_{1} \cdots \mathrm{q}_{\mathrm{j}-1} \mathrm{q}_{\mathrm{j}+1} \cdots \mathrm{q}_{\mathrm{m}} .
$$

Repeating this argument, we next eliminate $\mathrm{p}_{2}$ and one of the remaining factors on the right. Continuing in this manner, we pairwise eliminate one of the $p_{i}$ and one of the $q_{k}$ with each operation, always replacing a $q_{k}$ with a corresponding $u_{k}$. But the primes on one side of this equation can not be eliminated before those on the other side because this would imply that a product of prime polynomials was equal to 1 which is impossible. Therefore $\mathrm{n}=\mathrm{m}$, and the expansion of $f$ as a product of prime elements must be unique up to an associate.

Note that the expansion proved in this theorem for $\mathrm{f} \in \mathcal{F}[\mathrm{x}]$ is completely unique (except for order) if we require that the prime polynomials be monic.

Example 6.7 Consider the polynomial $\mathrm{p}=3 \mathrm{x}^{4}-3 \mathrm{x}^{2}-6 \in \mathcal{F}[\mathrm{x}]$. Using the fields $\mathbb{Q}, \mathbb{R}$ and $\mathbb{C}$ we can factor p three different ways depending on $\mathcal{F}$ :

$$
\begin{aligned}
3\left(x^{2}-2\right)\left(x^{2}+1\right) & \text { in } \mathbb{Q}[\mathrm{x}] \\
3(\mathrm{x}+\sqrt{2})(x-\sqrt{2})\left(x^{2}+1\right) & \text { in } \mathbb{R}[\mathrm{x}] \\
3(\mathrm{x}+\sqrt{2})(x-\sqrt{2})(x+i)(x-i) & \text { in } \mathbb{C}[\mathrm{x}]
\end{aligned}
$$

In each case, $p$ is a product of prime polynomials relative to the appropriate field. //

## Exercises

1. Let R be a commutative ring with unit element. Show that $\mathrm{u} \in \mathrm{R}$ is a unit of $R$ if and only if $u$ is a factor of every element of $R$.
2. Let R be an integral domain with unit element and suppose it is true that $a \mid b$ and $b \mid a$ for some $a, b \in R$. Show that $a=u b$ where $u$ is $a$ unit in $R$.
3. Show that the property of being associates defines an equivalence relation on a commutative ring with unit element.
4. Let $\mathcal{F}$ be an arbitrary field, and suppose $p \in \mathcal{F}[x]$ is of degree $\leq 3$. Prove that p is prime in $\mathcal{F}[\mathrm{x}]$ if and only if p is either of degree 1 , or has no zeros in $\mathcal{F}$. Give an example to show that this result is not true if $\operatorname{deg} \mathrm{p}>$ 3.
5. Factor the following polynomials into their prime factors in both $\mathbb{R}[x]$ and $\mathbb{Q}[\mathrm{x}]$ :
(a) $2 x^{3}-x^{2}+x+1$.
(b) $3 x^{3}+2 x^{2}-4 x+1$.
(c) $\mathrm{x}^{6}+1$.
(d) $x^{4}+16$
6. Let $\mathcal{F}=\{0,1\}$ be the field consisting of only two elements, and define addition and multiplication on $\mathcal{F}$ in the obvious way. Factor the following polynomials into primes in $\mathcal{F}[x]$ :
(a) $x^{2}+x+1$.
(b) $x^{3}+1$.
(c) $\mathrm{x}^{4}+\mathrm{x}^{2}+1$.
(d) $x^{4}+1$.
7. Let $\mathcal{F}$ be as in the previous problem. Find the greatest common divisor of $\mathrm{x}^{3}+\mathrm{x}^{2}+\mathrm{x}+1$ and $\mathrm{x}^{5}+\mathrm{x}^{4}+\mathrm{x}^{3}+\mathrm{x}^{2}+\mathrm{x}+1$ over $\mathcal{F}[\mathrm{x}]$.
8. Find the greatest common divisor of the following pairs of polynomials over $\mathbb{R}[\mathrm{x}]$. Express your result in the form defined in Theorem 6.5.
(a) $4 x^{3}+2 x^{2}-2 x-1$ and $2 x^{3}-x^{2}+x+1$.
(b) $x^{3}-x+1$ and $2 x^{4}+x^{2}+x-5$.
(c) $x^{4}+3 x^{2}+2$ and $x^{5}-x$.
(d) $x^{3}+x^{2}-2 x-2$ and $x^{4}-2 x^{3}+3 x^{2}-6 x$.
9. Use the remainder theorem to find the remainder when $2 x^{5}-3 x^{3}+2 x+$ $1 \in \mathbb{R}[\mathrm{x}]$ is divided by:
(a) $x-2 \in \mathbb{R}[x]$.
(b) $x+3 \in \mathbb{R}[x]$.
10. (a) Is $x-3$ a factor of $3 x^{3}-9 x^{2}-7 x+21$ over $\mathbb{Q}[x]$ ?
(b) Is $x+2$ a factor of $x^{3}+8 x^{2}+6 x-8$ over $\mathbb{R}[x]$ ?
(c) For which $k \in \mathbb{Q}$ is $x-1$ a factor of $x^{3}+2 x^{2}+x+k$ over $\mathbb{Q}[x]$ ?
(d) For which $k \in \mathbb{C}$ is $x+i$ a factor of $i x^{9}+3 x^{7}+x^{6}-2 i x+k$ over $\mathbb{C}[\mathrm{x}]$ ?
11. (a) Construct an example to show that the division algorithm is not true if $\mathcal{F}$ is replaced by the integral domain $\mathbb{Z}$.
(b) Prove that if the division algorithm is true for polynomials over an integral domain D , then D must be a field.
12. Determine the monic associate of:
(a) $2 x^{3}-x+1 \in \mathbb{Q}[x]$.
(b) $-i \mathrm{x}^{2}+\mathrm{x}+1 \in \mathbb{C}[\mathrm{x}]$.
13. Let $f=a_{0}+a_{1} x+\cdots+a_{n} x^{n} \in \mathbb{Z}[x]$ be a polynomial with integer coefficients, and suppose $r / s \in \mathbb{Q}$ is a rational root of $f$. Assume that $r$ and $s$ are relatively prime. Prove that $\mathrm{r} \mid \mathrm{a}_{0}$ and $\mathrm{s} \mid \mathrm{a}_{\mathrm{n}}$.

### 6.3 POLYNOMIAL IDEALS

We now apply the formalism of Section 1.5 to polynomials. If the reader has not yet studied that section (or does not remember it), now is the time to go back and do so (again if necessary).

Example 6.8 Let A and B be ideals of $\mathcal{F}[\mathrm{x}]$. We show that

$$
\mathrm{A} \cap \mathrm{~B}=\{\mathrm{f} \in \mathcal{F}[\mathrm{x}]: \mathrm{f} \in \mathrm{~A} \text { and } \mathrm{f} \in \mathrm{~B}\}
$$

and

$$
\mathrm{A}+\mathrm{B}=\{\mathrm{f}+\mathrm{g} \in \mathcal{F}[\mathrm{x}]: \mathrm{f} \in \mathrm{~A} \text { and } \mathrm{g} \in \mathrm{~B}\}
$$

are both ideals of $\mathcal{F}[x]$. Indeed, if $f, g \in A \cap B$, then $f \pm g \in A$ and $f \pm g \in B$ since $A$ and $B$ are ideals. Therefore $f \pm g \in A \cap B$ so that $A \cap B$ is a subgroup of $\mathcal{F}[x]$ under addition. Similarly, if $f \in A \cap B$ and $g \in \mathcal{F}[x]$, then $\mathrm{fg} \in \mathrm{A}$ and $\mathrm{fg} \in \mathrm{B}$ so that $\mathrm{fg} \in \mathrm{A} \cap \mathrm{B}$, and hence $\mathrm{A} \cap \mathrm{B}$ is an ideal. Now suppose $f_{1}+g_{1}, f_{2}+g_{2} \in A+B$. Then

$$
\left(\mathrm{f}_{1}+\mathrm{g}_{1}\right) \pm\left(\mathrm{f}_{2}+\mathrm{g}_{2}\right)=\left(\mathrm{f}_{1} \pm \mathrm{f}_{2}\right)+\left(\mathrm{g}_{1} \pm \mathrm{g}_{2}\right) \in \mathrm{A}+\mathrm{B}
$$

so that $\mathrm{A}+\mathrm{B}$ is also a subgroup of $\mathcal{F}[\mathrm{x}]$ under addition. Finally, suppose that $\mathrm{f}_{1}+\mathrm{g}_{1} \in \mathrm{~A}+\mathrm{B}$ and $\mathrm{h} \in \mathcal{F}[\mathrm{x}]$. Then

$$
\left(\mathrm{f}_{1}+\mathrm{g}_{1}\right) \mathrm{h}=\mathrm{f}_{1} \mathrm{~h}+\mathrm{g}_{1} \mathrm{~h} \in \mathrm{~A}+\mathrm{B}
$$

so that $\mathrm{A}+\mathrm{B}$ is an ideal also. //

Recall that $\mathcal{F}[\mathrm{x}]$ is not only a ring, it is in fact an integral domain (Corollary 2 of Theorem 6.2). Since $\mathcal{F}[x]$ is a ring, we see that if $p \in \mathcal{F}[x]$, then $p$ can be used to generate the principal ideal (p) of $\mathcal{F}[x]$. Note that by construction, the ideal (p) can not contain any prime polynomials (other than associates of $p$ ) even if $p$ itself is prime. This is because any $q \in(p)$ can be written in the form $\mathrm{q}=\mathrm{pr}$ for some $\mathrm{r} \in \mathcal{F}[\mathrm{x}]$, and hence plq . We will also write the principal ideal $(\mathrm{p})$ in the form

$$
(\mathrm{p})=\mathrm{p} \mathcal{F}[\mathrm{x}]=\{\mathrm{pf}: \mathrm{f} \in \mathcal{F}[\mathrm{x}]\}
$$

Our next theorem is frequently useful when working with quotient rings (see Theorem 1.13) of the general form $\mathcal{F}[x] /(p)$.

Theorem 6.7 Suppose $\mathrm{p}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{n}} \mathrm{x}^{\mathrm{n}} \in \mathcal{F}[\mathrm{x}], \mathrm{a}_{\mathrm{n}} \neq 0$, and let $\mathrm{I}=$ (p). Then every element of $\mathcal{F}[x] / I$ can be uniquely expressed in the form

$$
\mathrm{I}+\left(\mathrm{b}_{0}+\mathrm{b}_{1} \mathrm{x}+\cdots+\mathrm{b}_{\mathrm{n}-1} \mathrm{x}^{\mathrm{n}-1}\right)
$$

where $\mathrm{b}_{0}, \ldots, \mathrm{~b}_{\mathrm{n}-1} \in \mathcal{F}$.

Proof Choose any I $+\mathrm{f} \in \mathcal{F}[\mathrm{x}] / \mathrm{I}$. By the division algorithm (Theorem 6.3) we can write $\mathrm{f}=\mathrm{pq}+\mathrm{r}$ for some $\mathrm{q}, \mathrm{r} \in \mathcal{F}[\mathrm{x}]$ with either $\mathrm{r}=0$ or $\operatorname{deg} \mathrm{r}<\operatorname{deg} \mathrm{p}$. But by definition of I , we have $\mathrm{pq} \in \mathrm{I}$ so that

$$
\mathrm{I}+\mathrm{f}=\mathrm{I}+(\mathrm{pq}+\mathrm{r})=\mathrm{I}+\mathrm{r}
$$

This shows that $\mathrm{I}+\mathrm{f}$ has the form desired.
To prove the uniqueness of this representation, suppose that

$$
\mathrm{I}+\left(\mathrm{b}_{0}+\mathrm{b}_{1} \mathrm{x}+\cdots+\mathrm{b}_{\mathrm{n}-1} \mathrm{x}^{\mathrm{n}-1}\right)=\mathrm{I}+\left(\mathrm{c}_{0}+\mathrm{c}_{1} \mathrm{x}+\cdots+\mathrm{c}_{\mathrm{n}-1} \mathrm{x}^{\mathrm{n}-1}\right)
$$

Then (by adding $\mathrm{I}+\left(-\mathrm{c}_{0}-\mathrm{c}_{1} \mathrm{x}-\cdots-\mathrm{c}_{\mathrm{n}-1} \mathrm{x}^{\mathrm{n}-1}\right)$ to both sides) we see that

$$
\left(\mathrm{b}_{0}-\mathrm{c}_{0}\right)+\left(\mathrm{b}_{1}-\mathrm{c}_{1}\right) \mathrm{x}+\cdots+\left(\mathrm{b}_{\mathrm{n}-1}-\mathrm{c}_{\mathrm{n}-1}\right) \mathrm{x}^{\mathrm{n}-1} \in \mathrm{I} .
$$

But the degree of any nonzero polynomial in I must be greater than or equal to $\mathrm{n}=\operatorname{deg} \mathrm{p}$ (by definition of I and Theorem 6.2(b)), and therefore it follows that

$$
\left(b_{0}-c_{0}\right)+\left(b_{1}-c_{1}\right) x+\cdots+\left(b_{n-1}-c_{n-1}\right) x^{n-1}=0
$$

Since two polynomials are equal if and only if their coefficients are equal, this means that $\mathrm{b}_{\mathrm{i}}=\mathrm{c}_{\mathrm{i}}$ for every $\mathrm{i}=0, \ldots, \mathrm{n}-1$.

It is an interesting fact that every ideal of $\mathcal{F}[\mathrm{x}]$ is actually a principal ideal. We prove this in our next theorem.

Theorem 6.8 Every ideal of the $\operatorname{ring} \mathcal{F}[\mathrm{x}]$ is a principal ideal.
Proof Let I be any ideal of $\mathcal{F}[\mathrm{x}]$. If $\mathrm{I}=\{0\}$, then I is just the principal ideal (0). Now assume that $\mathrm{I} \neq\{0\}$ and let g be any nonzero polynomial of least degree in I. (That $g$ exists follows from the well-ordering principle. In other words, if S is the set of degrees of all polynomials in $\mathcal{F}[\mathrm{x}]$, then S has a least element.) From the definitions it is clear that $(\mathrm{g}) \subset \mathrm{I}$. We now show that $\mathrm{I} \subset$ (g) which will then prove that $\mathrm{I}=(\mathrm{g})$.

By the division algorithm, for any $\mathrm{f} \in \mathrm{I}$ there exist polynomials $\mathrm{q}, \mathrm{r} \in \mathcal{F}[\mathrm{x}]$ such that $\mathrm{f}=\mathrm{gq}+\mathrm{r}$ where either $\mathrm{r}=0$ or $\operatorname{deg} \mathrm{r}<\operatorname{deg} \mathrm{g}$. Since $\mathrm{f} \in \mathrm{I}$ and $\mathrm{gq} \in \mathrm{I}$, it follows that $\mathrm{r}=\mathrm{f}-\mathrm{gq} \in \mathrm{I}$. But if $\mathrm{r} \neq 0$ we have $\operatorname{deg} \mathrm{r}<\operatorname{deg} \mathrm{g}$ which contradicts the definition of $g$ as the polynomial of least degree in $I$. Therefore $r=0$ so that $\mathrm{f}=\mathrm{gq} \in(\mathrm{g})$, and hence $\mathrm{I} \subset(\mathrm{g})$.

Corollary Every ideal of $\mathcal{F}[\mathrm{x}]$ is generated by a unique monic polynomial.
Proof Since every ideal of $\mathcal{F}[x]$ is principal, suppose that $(p)=(q)$ or, equivalently, $\mathrm{p} \mathcal{F}[\mathrm{x}]=\mathrm{q} \mathcal{F}[\mathrm{x}]$. Then clearly $\mathrm{p} \in \mathrm{p} \mathcal{F}[\mathrm{x}]$ so that $\mathrm{p}=\mathrm{qf}_{1}$ for some $\mathrm{f}_{1} \in$ $\mathcal{F}[\mathrm{x}]$. Similarly, we see that $\mathrm{q}=\mathrm{pf}_{2}$ for some $\mathrm{f}_{2} \in \mathcal{F}[\mathrm{x}]$ and hence

$$
\mathrm{p}=\mathrm{qf}_{1}=\mathrm{pf}_{2} \mathrm{f}_{1}
$$

Since $\mathcal{F}[\mathrm{x}]$ is an integral domain it follows that $\mathrm{f}_{1} \mathrm{f}_{2}=1$. But this means that $\mathrm{f}_{1}$ and $\mathrm{f}_{2}$ are units (i.e., constant polynomials), and hence $\mathrm{q}=\mathrm{cp}$ for some $\mathrm{c} \in \mathcal{F}$. Noting that $\mathrm{cp} \mathcal{F}[\mathrm{x}]$ is the same as $\mathrm{p} \mathcal{F}[\mathrm{x}]$, we see that any ideal $\mathrm{p} \mathcal{F}[\mathrm{x}]$ may be written in the form $\mathrm{cp} \mathcal{F}[\mathrm{x}]$ for arbitrary $\mathrm{c} \in \mathcal{F}$. By choosing c to be the inverse of the leading coefficient of p , we have shown that any ideal of $\mathcal{F}[\mathrm{x}]$ has a unique monic generator.

In Section 6.2 we discussed the greatest common divisor of a collection of polynomials. We now treat a related concept that the reader may be wondering about. If $\mathrm{f}, \mathrm{g} \in \mathcal{F}[\mathrm{x}]$ then, by the least common multiple (or simply lcm) of f and g , we mean the polynomial $\mathrm{m} \in \mathcal{F}[\mathrm{x}]$ such that $\mathrm{f} \mid \mathrm{m}$ and $\mathrm{g} \mid \mathrm{m}$, and if $\mathrm{m}^{\prime} \in \mathcal{F}[\mathrm{x}]$ is another polynomial that satisfies $\mathrm{f} \mid \mathrm{m}^{\prime}$ and $\mathrm{g} \mid \mathrm{m}^{\prime}$, then $\mathrm{m} / \mathrm{m}^{\prime}$.

As a useful observation, note that if $\mathrm{f}, \mathrm{g} \in \mathcal{F}[\mathrm{x}]$ and $\mathrm{f} \mid \mathrm{g}$, then $\mathrm{g}=\mathrm{fq}$ for some $\mathrm{q} \in \mathcal{F}[\mathrm{x}]$. But $(\mathrm{f})=\mathrm{f} \mathcal{F}[\mathrm{x}]$ and hence

$$
(\mathrm{g})=\mathrm{g} \mathcal{F}[\mathrm{x}]=\mathrm{fq} \mathcal{F}[\mathrm{x}] \subset(\mathrm{f})
$$

In other words, $\mathrm{f} \mid \mathrm{g}$ implies that $(\mathrm{g}) \subset(\mathrm{f})$.
Example 6.9 Let A and B be ideals of $\mathcal{F}[\mathrm{x}]$. By Theorem 6.8 and Example 6.8 we may write $\mathrm{A}=\mathrm{h} \mathcal{F}[\mathrm{x}]$ and $\mathrm{B}=\mathrm{k} \mathcal{F}[\mathrm{x}]$, and also $\mathrm{A} \cap \mathrm{B}=\mathrm{m} \mathcal{F}[\mathrm{x}]$ and $\mathrm{A}+\mathrm{B}=\mathrm{d} \mathcal{F}[\mathrm{x}]$. We claim that d is a greatest common divisor of h and k , and that m is a least common multiple of h and k .

To see this, first note that since $\mathrm{h} \in \mathrm{h} \mathcal{F}[\mathrm{x}]=\mathrm{A}$, it follows that $\mathrm{h}=\mathrm{h}+0 \in$ $\mathrm{A}+\mathrm{B}$, and hence $\mathrm{h}=\mathrm{dh}_{1}$ for some $\mathrm{h}_{1} \in \mathcal{F}[\mathrm{x}]$. Similarly, we must have $\mathrm{k}=\mathrm{dk}_{1}$ for some $\mathrm{k}_{1} \in \mathcal{F}[\mathrm{x}]$. Therefore $\mathrm{d} \mid \mathrm{h}$ and $\mathrm{d} \mid \mathrm{k}$ so that d is a common divisor of h and $k$. We must show that if $d^{\prime} \mid h$ and $d^{\prime} \mid k$, then $d^{\prime} \mid d$. Now, if $d^{\prime} \mid h$ and $d^{\prime} \mid k$, then $\mathrm{A}=(\mathrm{h}) \subset\left(\mathrm{d}^{\prime}\right)$ and $\mathrm{B}=(\mathrm{k}) \subset\left(\mathrm{d}^{\prime}\right)$. But then $\mathrm{A}+\mathrm{B} \subset\left(\mathrm{d}^{\prime}\right)$ because for any $\mathrm{f}+\mathrm{g} \in \mathrm{A}+\mathrm{B}$ we have $\mathrm{f} \in \mathrm{A} \subset\left(\mathrm{d}^{\prime}\right)$ and $\mathrm{g} \in \mathrm{B} \subset\left(\mathrm{d}^{\prime}\right)$, and therefore $\mathrm{f}+\mathrm{g} \in$ ( $\mathrm{d}^{\prime}$ ) since ( $\mathrm{d}^{\prime}$ ) is an ideal. This means that $\mathrm{d} \in \mathrm{A}+\mathrm{B} \subset\left(\mathrm{d}^{\prime}\right)$ so that $\mathrm{d}=\mathrm{d}^{\prime} \mathrm{p}$ for some $\mathrm{p} \in \mathcal{F}[\mathrm{x}]$, and hence $\mathrm{d}^{\prime} \mid \mathrm{d}$.

Now note that $\mathrm{m} \in \mathrm{A} \cap \mathrm{B}$ so that $\mathrm{m} \in A$ implies $\mathrm{m}=\mathrm{hm}_{1}$ and $\mathrm{m} \in \mathrm{B}$ implies $\mathrm{m}=\mathrm{km}_{2}$ for some polynomials $\mathrm{m}_{1}, \mathrm{~m}_{2} \in \mathcal{F}[\mathrm{x}]$. This means that $\mathrm{h} \mid \mathrm{m}$ and $\mathrm{k} \mid \mathrm{m}$ so that m is a common multiple of h and k . Next, note that if $\mathrm{h} \mid \mathrm{m}^{\prime}$ then $\left(\mathrm{m}^{\prime}\right) \subset(\mathrm{h})=\mathrm{A}$, and if $\mathrm{k} \mid \mathrm{m}^{\prime}$ then $\left(\mathrm{m}^{\prime}\right) \subset(\mathrm{k})=\mathrm{B}$. Therefore we see that $\mathrm{m}^{\prime} \in\left(\mathrm{m}^{\prime}\right) \subset \mathrm{A} \cap \mathrm{B}=(\mathrm{m})$ so that $\mathrm{m}^{\prime}=\mathrm{mq}$ for $\mathrm{q} \in \mathcal{F}[\mathrm{x}]$. Hence $\mathrm{m} \mid \mathrm{m}^{\prime}$ and m is a least common multiple of h and k . //

Greatest common divisors and least common multiples will be of considerable use to us in the next chapter. The following important theorem relates least common multiples and greatest common divisors. Rather than prove it directly, we simply note that it follows easily from Theorem 6.10 below.

Theorem 6.9 Suppose $\mathrm{h}, \mathrm{k} \in \mathcal{F}[\mathrm{x}]$ and let d and m be the greatest common divisor and least common multiple respectively of $h$ and $k$. Then $\mathrm{hk}=\mathrm{dm}$.

Recall from Theorem 6.6 that any polynomial $h \in \mathcal{F}[x]$ is expressible as a unique product of prime polynomials. If $h$ contains the prime factor $g_{i} \in \mathcal{F}[x]$ repeated $r_{i}$ times, then we write $g_{i}{ }^{r_{i}}$ as one of the factors of $h$. Therefore we may write the decomposition of $h$ in the form $h=\prod_{i} \mathrm{~g}_{\mathrm{i}} \mathrm{r}_{\mathrm{i}}$. If $\mathrm{k} \in \mathcal{F}[\mathrm{x}]$ is another polynomial, then it may also be factored in the same manner as $k=\Pi_{j} q_{j}{ }^{s_{j}}$. In fact, we may write both h and k as a product of the same factors if we allow the exponent to be zero for any factor that does not appear in that expansion. In other words, we may write $h=\prod_{1}^{n}=1 p_{i}{ }^{r_{i}}$ and $k=\prod_{i=1}^{n} p_{i}^{s_{i}}$ where $r_{i} \geq 0$ and $\mathrm{s}_{\mathrm{i}} \geq 0$ for each $\mathrm{i}=1, \ldots, \mathrm{n}$.

The next theorem contains Theorem 6.9 as an immediate and obvious corollary.

Theorem 6.10 Suppose that $\mathrm{h}, \mathrm{k} \in \mathcal{F}[\mathrm{x}]$ and write

$$
h=\prod_{i=1}^{n} p_{i}^{r_{i}} \quad k=\prod_{i=1}^{n} p_{i}{ }^{s_{i}}
$$

where each $r_{i} \geq 0$ and each $s_{i} \geq 0$. For the given $r_{i}$ and $s_{i}$, define the polynomials

$$
\alpha=\prod_{r_{i}>s_{i}} p_{i}{ }^{r_{i}} \quad \beta=\prod_{r_{i} \leq s_{i}} p_{i}{ }^{r_{i}} \quad \gamma=\prod_{s_{i}<r_{i}} p_{i}{ }^{s_{i}} \quad \delta=\prod_{s_{i} \geq r_{i}} p_{i}{ }^{s_{i}}
$$

so that $\mathrm{h}=\alpha \beta$ and $\mathrm{k}=\gamma \delta$. Then the least common multiple m of h and k is given by

$$
m=\alpha \delta=\prod_{i=1}^{n} p_{i}^{\max \left(r_{i}, s_{i}\right)}
$$

and the greatest common divisor d is given by

$$
d=\beta \gamma=\prod_{i=1}^{n} p_{i}^{\min \left(r_{i}, s_{i}\right)} .
$$

Proof Since the expressions for h and k are given in terms of the same set of prime polynomials $p_{i}$, a moment's thought should make it clear that the least common multiple is given by

$$
m=\prod_{i=1}^{n} p_{i}^{\max \left(r_{i}, s_{i}\right)}
$$

Formally, we see that $\mathrm{h} \mid \mathrm{m}$ and $\mathrm{k} \mid \mathrm{m}$, and if $\mathrm{m}^{\prime}$ is another polynomial such that $\mathrm{h} \mid \mathrm{m}^{\prime}$ and $\mathrm{k} \mid \mathrm{m}^{\prime}$, then by Theorem 6.6 again we can write $\mathrm{m}^{\prime}=\prod_{1=1}^{\mathrm{n}} 1_{1} \mathrm{p}_{\mathrm{i}}{ }^{\mathrm{t}_{\mathrm{i}}}$ where we must have $t_{i} \geq r_{i}$ and $t_{i} \geq s_{i}$ for each $i=1, \ldots, n$ in order that $m^{\prime}$ be a common multiple of $h$ and $k$. But this means that $m \mid m^{\prime}$ so that $m$ is the least common multiple of $h$ and $k$. In any case, $m$ is exactly the same as the product $\alpha \delta$.

That the greatest common divisor is given by $\prod_{1=1}^{\mathrm{n}} \mathrm{p}_{\mathrm{i}}{ }^{\min \left(\mathrm{r}_{\mathrm{i}}, \mathrm{s}_{\mathrm{i}}\right)}=\beta \gamma$ follows from a similar argument.

By Theorem 1.13, the quotient structure $\mathcal{F}[\mathrm{x}] /(\mathrm{p})$ is a ring for any p , and we now show that it is actually a field for appropriate $p$.

Theorem 6.11 Suppose $\mathrm{p} \in \mathcal{F}[x]$, and let $\mathrm{I}=(\mathrm{p})$. Then $\mathcal{F}[\mathrm{x}] / \mathrm{I}$ is a field if and only if p is prime over $\mathcal{F}$.

Proof We first show that if p is reducible over $\mathcal{F}$, then $\mathcal{F}[\mathrm{x}] / \mathrm{I}$ is not a field. (This is the contrapositive to the statement that if $\mathcal{F}[\mathrm{x}] / \mathrm{I}$ is a field, then p must be prime.) To see this, assume that $\mathrm{p}=\mathrm{ab}$ where neither a nor b is a unit, and each is of degree less than that of $p$. From the definition of a principal ideal, we see that the degree of any polynomial in I must be greater than or equal to deg p , and hence neither a nor b can be an element of I. Since I is the zero element of $\mathcal{F}[x] / I$, we see that $I+a \neq I$ and $I+b \neq I$ must both be nonzero elements of $\mathcal{F}[x] / \mathrm{I}$. But then

$$
(I+a)(I+b)=I+a b=I+p=I
$$

where we used the fact that $p \in I$. This shows that the product of two nonzero elements of $\mathcal{F}[x] / I$ yields the zero element of $\mathcal{F}[x] / I$, and thus the set of nonzero elements of $\mathcal{F}[x] / I$ is not closed under multiplication. Hence $\mathcal{F}[x] / I$ is neither a division ring nor an integral domain, so it certainly is not a field.

Conversely, suppose that p is prime. Since $\mathcal{F}[\mathrm{x}]$ is a commutative ring, it follows that for any $\mathrm{a}, \mathrm{b} \in \mathcal{F}[\mathrm{x}]$ we have that $\mathcal{F}[\mathrm{x}] / \mathrm{I}$ is also a commutative ring. The identity element in $\mathcal{F}[x] / I$ is easily seen to be $I+1$ where 1 is the unit element for the field $\mathcal{F}$. Therefore, all that remains is to show the existence of a multiplicative inverse for each nonzero element in $\mathcal{F}[x] / I$.

If $\mathrm{I}+\mathrm{f}$ is any nonzero element in $\mathcal{F}[\mathrm{x}] / \mathrm{I}$, then $\mathrm{f} \notin \mathrm{I}$ so that $\mathrm{p} \mid \mathrm{f}$. Since p is prime, its only divisors are units and its associates, and therefore the greatest common divisor of p and f is a unit (i.e., p and f are relatively prime). Applying Corollary 1 of Theorem 6.5 we see there exist $u, v \in \mathcal{F}[x]$ such that $u p+v f=1$. Then $1-v f=u p \in I$ and hence

$$
\mathrm{I}+1=\mathrm{I}+\mathrm{up}+\mathrm{vf}=\mathrm{I}+\mathrm{vf}=(\mathrm{I}+\mathrm{v})(\mathrm{I}+\mathrm{f})
$$

This shows that $I+v$ is a multiplicative inverse of $I+f$.

## Exercises

1. Referring to Theorem 6.7, show that $\{\mathrm{I}+\mathrm{c}: \mathrm{c} \in \mathcal{F}\}$ is a subfield of $\mathcal{F}[\mathrm{x}] / \mathrm{I}$ isomorphic to $\mathcal{F}$.
2. Let $p=1+x^{2} \in \mathbb{R}[x]$ and let $I=(p)$.
(a) Show $\mathbb{R}[x] / I$ is a field.
(b) Show $\mathbb{R}[x] / I$ is isomorphic to $\mathbb{C}$. [Hint: Justify defining the mapping $\theta$ : $\mathbb{R}[x] / \mathbb{I} \rightarrow \mathbb{C}$ by $\theta(\mathrm{I}+(\mathrm{a}+\mathrm{bx}))=\mathrm{a}+i \mathrm{~b}$. Show that $\theta$ is bijective and preserves addition. To show that $\theta$ preserves multiplication, note that

$$
(\mathrm{I}+(\mathrm{a}+\mathrm{bx}))(\mathrm{I}+(\mathrm{c}+\mathrm{dx}))=\mathrm{I}+\left(\mathrm{ac}+(\mathrm{ad}+\mathrm{bc}) \mathrm{x}+\mathrm{bd} \mathrm{x}^{2}\right) .
$$

Write this in the form $\mathrm{I}+(\mathrm{u}+\mathrm{vx})$ by following the first part of the proof of Theorem 6.7. Now show that

$$
\theta[(\mathrm{I}+(\mathrm{a}+\mathrm{bx}))(\mathrm{I}+(\mathrm{c}+\mathrm{dx}))]=\theta(\mathrm{I}+(\mathrm{a}+\mathrm{bx})) \theta(\mathrm{I}+(\mathrm{c}+\mathrm{dx})) .]
$$

3. Suppose $\mathrm{f}, \mathrm{g} \in \mathcal{F}[\mathrm{x}]$. Prove that $(\mathrm{f})=(\mathrm{g})$ if and only if f and g are associates.
4. Suppose $\mathrm{f}, \mathrm{g} \in \mathcal{F}[\mathrm{x}]$. Prove or disprove the following:
(a) If (f) $=(\mathrm{g})$, then $\operatorname{deg} \mathrm{f}=\operatorname{deg} \mathrm{g}$.
(b) If $\operatorname{deg} f=\operatorname{deg} g$, then (f) $=(\mathrm{g})$.
(c) If $\mathrm{f} \in(\mathrm{g})$ and $\operatorname{deg} \mathrm{f}=\operatorname{deg} \mathrm{g}$, then (f) $=(\mathrm{g})$.
5. Find the greatest common divisor and least common multiple of the following pairs of polynomials:
(a) $(x-1)(x+2)^{2}$ and $(x+2)(x-4)$.
(b) $(\mathrm{x}-2)^{2}(\mathrm{x}-3)^{4}(\mathrm{x}-i)$ and $(\mathrm{x}-1)(\mathrm{x}-2)(\mathrm{x}-3)^{3}$.
(c) $\left(\mathrm{x}^{2}+1\right)\left(\mathrm{x}^{2}-1\right)$ and $(\mathrm{x}+i)^{3}\left(\mathrm{x}^{3}-1\right)$.
6. (a) Suppose $\mathrm{f}_{1}, \ldots, \mathrm{f}_{\mathrm{n}} \in \mathcal{F}[\mathrm{x}]$, and let $\mathrm{I}=\mathrm{f}_{1} \mathcal{F}+\cdots+\mathrm{f}_{\mathrm{n}} \mathcal{F}$ be the set of all polynomials of the form $g=f_{1} g_{1}+\cdots+f_{n} g_{n}$ where $g_{i} \in \mathcal{F}[x]$. Show that $I$ is an ideal. This is called the ideal generated by $\left\{\mathrm{f}_{1}, \ldots, \mathrm{f}_{\mathrm{n}}\right\}$.
(b) Show, in particular, that $\mathcal{F}[x]$ is an ideal generated by $\{1\}$. This is called the unit ideal.
(c) Let d be the unique monic generator of I. Show that d divides each of the $\mathrm{f}_{\mathrm{i}}$.
(d) If $\mathrm{c} \in \mathcal{F}[\mathrm{x}]$ divides each of the $\mathrm{f}_{\mathrm{i}}$, show that $\mathrm{c} \mid \mathrm{d}$.
(e) Suppose $\left\{\mathrm{f}_{1}, \mathrm{f}_{2}, \mathrm{f}_{3}\right\}$ generates the unit ideal. Show that we can always find polynomials $\mathrm{f}_{\mathrm{ij}} \in \mathcal{F}[\mathrm{x}]$ such that

$$
\left|\begin{array}{ccc}
f_{1} & f_{2} & f_{3} \\
f_{21} & f_{22} & f_{23} \\
f_{31} & f_{32} & f_{33}
\end{array}\right|=1
$$

[Hint: Show there exists $\mathrm{g}_{1}, \mathrm{~g}_{2}, \mathrm{~g}_{3} \in \mathcal{F}[\mathrm{x}]$ such that $\sum \mathrm{g}_{\mathrm{i}} \mathrm{f}_{\mathrm{i}}=1$, and let $\alpha=$ $\operatorname{gcd}\left\{\mathrm{g}_{1}, \mathrm{~g}_{2}\right\}$. Next, show there exists $\mathrm{h}_{1}, \mathrm{~h}_{2} \in \mathcal{F}[\mathrm{x}]$ such that $\left(\mathrm{g}_{1} / \alpha\right) \mathrm{h}_{1}+$ $\left(\mathrm{g}_{2} / \alpha\right) \mathrm{h}_{2}=1$. Now use the polynomials $\mathrm{g}_{\mathrm{i}}, \mathrm{h}_{\mathrm{i}}$ and $\alpha$ to form the $\mathrm{f}_{\mathrm{ij}}$ ]

### 6.4 POLYNOMIALS OVER ALGEBRAICALLY CLOSED FIELDS

We now turn to a discussion of polynomials over the fields $\mathbb{R}$ and $\mathbb{C}$. These are well worth considering in more detail since most practical applications in mathematics and physics deal with these two special cases. By way of terminology, a field $\mathcal{F}$ is said to be algebraically closed if every polynomial $\mathrm{f} \in$ $\mathcal{F}[\mathrm{x}]$ with $\operatorname{deg} \mathrm{f}>0$ has at least one zero (or root) in $\mathcal{F}$.

Our next theorem is called the Fundamental Theorem of Algebra. While most proofs of this theorem involve the theory of complex variables, this result is so fundamental to our work that we present a proof in Appendix A that depends only on some relatively elementary properties of metric spaces. Basically, if the reader knows that a continuous function defined on a compact space takes its maximum and minimum values on the space, then there should be no problem understanding the proof. However, if the reader does not even know what a compact space is, then Appendix A presents all of the necessary formalism for a reasonably complete understanding of the concepts involved.

Theorem 6.12 (Fundamental Theorem of Algebra) The complex number field $\mathbb{C}$ is algebraically closed.

Proof See Appendix A.

Let $\mathrm{p}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{n}} \mathrm{x}^{\mathrm{n}}$ be a polynomial of degree $\mathrm{n} \geq 1$ over an algebraically closed field $\mathcal{F}$. Then there exists an element $\alpha_{1} \in \mathcal{F}$ such that $p\left(\alpha_{1}\right)=0$. Hence applying the factor theorem (Corollary to Theorem 6.4) we have

$$
\mathrm{p}=\left(\mathrm{x}-\alpha_{1}\right) \mathrm{q}
$$

where q is a polynomial of degree $\mathrm{n}-1$. Again, if $\mathrm{n}-1>0$, we see that q has a zero $\alpha_{2}$ in $\mathcal{F}$, and continuing this process we obtain

$$
\mathrm{p}=\mathrm{c}\left(\mathrm{x}-\alpha_{1}\right)\left(\mathrm{x}-\alpha_{2}\right) \cdots\left(\mathrm{x}-\alpha_{\mathrm{n}}\right)
$$

where c is a unit of $\mathcal{F}$. We thus see that any polynomial of degree $\mathrm{n} \geq 1$ over an algebraically closed field has exactly $n$ roots (although they need not be distinct). We repeat this statement as part of the next theorem.

Theorem 6.13 Let $\mathcal{F}$ be an algebraically closed field. Then every prime polynomial $\mathrm{p} \in \mathcal{F}[\mathrm{x}]$ has (up to a unit factor) the form $\mathrm{x}-\mathrm{a}$ where $\mathrm{a} \in \mathcal{F}$. Moreover, every monic polynomial $\mathrm{f} \in \mathcal{F}[\mathrm{x}]$ can be factored into the form

$$
f=\prod_{i=1}^{n}\left(x-a_{i}\right)
$$

where each $\mathrm{a}_{\mathrm{i}} \in \mathcal{F}$.

Proof Let $\mathrm{p} \in \mathcal{F}[\mathrm{x}]$ be prime. Since $\mathcal{F}$ is algebraically closed, there exists $\mathrm{a} \in \mathcal{F}$ such that $\mathrm{p}(\mathrm{a})=0$. By the factor theorem, $\mathrm{x}-\mathrm{a}$ must be a factor of p . But p is prime so its only factors are its associates and units. This proves the first part of the theorem.

Now let $\mathrm{f} \in \mathcal{F}[\mathrm{x}]$ be of degree $\mathrm{n} \geq 1$. The second part of the theorem is essentially obvious from the first part and Theorem 6.6. However, we may proceed as follows. Since $\mathcal{F}$ is algebraically closed there exists $a_{1} \in \mathcal{F}$ such that $\mathrm{f}\left(\mathrm{a}_{1}\right)=0$, and hence by the factor theorem,

$$
\mathrm{f}=\left(\mathrm{x}-\mathrm{a}_{1}\right) \mathrm{q}_{1}
$$

where $\mathrm{q}_{1} \in \mathcal{F}[\mathrm{x}]$ and $\operatorname{deg} \mathrm{q}_{1}=\mathrm{n}-1$ (Theorem 6.2(b)). Now, by the algebraic closure of $\mathcal{F}$ there exists $\mathrm{a}_{2} \in \mathcal{F}$ such that $\mathrm{q}_{1}\left(\mathrm{a}_{2}\right)=0$, and therefore

$$
\mathrm{q}_{1}=\left(\mathrm{x}-\mathrm{a}_{2}\right) \mathrm{q}_{2}
$$

where $\operatorname{deg} \mathrm{q}_{2}=\mathrm{n}-2$. It is clear that we can continue this process a total of n times, finally arriving at

$$
\mathrm{f}=\mathrm{c}\left(\mathrm{x}-\mathrm{a}_{1}\right)\left(\mathrm{x}-\mathrm{a}_{2}\right) \cdots\left(\mathrm{x}-\mathrm{a}_{\mathrm{n}}\right)
$$

where $\mathrm{c} \in \mathcal{F}$ is a unit. In particular, $\mathrm{c}=1$ if $\mathrm{q}_{\mathrm{n}-1}$ is monic.

While Theorem 6.13 shows that any polynomial of degree $n$ over an algebraically closed field has exactly n (not necessarily distinct) roots, a more general result is the following.

Theorem 6.14 Any polynomial $\mathrm{p} \in \mathcal{F}[\mathrm{x}]$ of degree $\mathrm{n} \geq 1$ over $\mathcal{F}$ has at most n roots in $\mathcal{F}$.

Proof We proceed by induction on the degree $n$ of $p$. If $n=1$, then $p=a_{0}+$ $a_{1} x$ so that $-a_{1} / a_{0}$ is the unique root of $p$, and the theorem thus holds in this case. Now assume that $\mathrm{n}>1$ and that the theorem holds for all polynomials of degree less than $n$. If $p$ has no roots, then the conclusion of the theorem is valid, so we assume that $p$ has at least one root $c$. Then $(x-c) \mid p$ so that $p=$ $(\mathrm{x}-\mathrm{c}) \mathrm{q}$ for some $\mathrm{q} \in \mathcal{F}[\mathrm{x}]$ with $\operatorname{deg} \mathrm{q}=\mathrm{n}-1$ (Theorem 6.2(b)). By our induction hypothesis, q has at most $\mathrm{n}-1$ roots in $\mathcal{F}$, so the proof will be finished if we can show that p has no roots in $\mathcal{F}$ other than c and the roots of q. Suppose that $b \in \mathcal{F}$ is such that $\mathrm{p}(\mathrm{b})=(\mathrm{b}-\mathrm{c}) \mathrm{q}(\mathrm{b})=0$. Since the field $\mathcal{F}$ can have no zero divisors (Exercise 1.5.12), it must be true that either $\mathrm{b}-\mathrm{c}=0$ or $q(b)=0$. In other words, if $p(b)=0$, then either $b=c$ or else $b$ is a root of $q$.

Corollary Every polynomial p of degree $\mathrm{n} \geq 1$ over an algebraically closed field $\mathcal{F}$ has n roots in $\mathcal{F}$.

Proof While this was proved in Theorem 6.13, we repeat it here in a slightly different manner. As was done in the proof of Theorem 6.14, we proceed by induction on the degree of p . The case $\mathrm{n}=1$ is true as above, so we assume that $\mathrm{n}>1$. Since $\mathcal{F}$ is algebraically closed, there exists at least one root $\mathrm{c} \in \mathcal{F}$ such that $\mathrm{p}=(\mathrm{x}-\mathrm{c}) \mathrm{q}$ where $\operatorname{deg} \mathrm{q}=\mathrm{n}-1$. By our induction hypothesis, q has $\mathrm{n}-1$ roots in $\mathcal{F}$ which are also clearly roots of p . It therefore follows that p has at least $\mathrm{n}-1+1=\mathrm{n}$ roots in $\mathcal{F}$, while Theorem 6.14 shows that p has at most n roots in $\mathcal{F}$. Therefore p must have exactly n roots in $\mathcal{F}$.

While we proved in Theorem 6.12 that the field $\mathbb{C}$ is algebraically closed, it is not true that $\mathbb{R}$ is algebraically closed. This should be obvious because any quadratic equation of the form $\mathrm{ax}^{2}+\mathrm{bx}+\mathrm{c}=0$ has solutions given by the quadratic formula

$$
\frac{-b \pm \sqrt{b^{2}-4 a c}}{2 a}
$$

and if $\mathrm{b}^{2}-4 \mathrm{ac}<0$, then there is no solution for x in the real number system. (Recall that the quadratic formula follows by writing

$$
0=x^{2}+b x / a+c / a=(x+b / 2 a)^{2}-b^{2} / 4 a^{2}+c / a
$$

and solving for x .) However, in the case of $\mathbb{R}[\mathrm{x}]$, we do have the following result.

Theorem 6.15 Suppose $f=a_{0}+a_{1} x+\cdots+a_{n} x^{n} \in \mathbb{R}[x]$. If $\alpha \in \mathbb{C}$ is a root of $f$, then so is $\alpha^{*}$. Furthermore, if $\alpha \neq \alpha^{*}$, then $(x-\alpha)\left(x-\alpha^{*}\right)$ is a factor of $f$.

Proof If $\alpha \in \mathbb{C}$ is a root of $f$, then $a_{0}+a_{1} \alpha+\cdots+a_{n} \alpha^{n}=0$. Taking the complex conjugate of this equation and remembering that each $a_{i} \in \mathbb{R}$, we obtain $a_{0}+a_{1} \alpha^{*}+\cdots+a_{n} \alpha^{* n}=0$ so that $\alpha^{*}$ is also a root of $f$. The second part of the theorem now follows directly from the factor theorem.

Corollary Every prime polynomial in $\mathbb{R}[\mathrm{x}]$ is (up to a unit factor) either of the form $\mathrm{x}-\mathrm{a}$ or $\mathrm{x}^{2}+\mathrm{ax}+\mathrm{b}$ where $\mathrm{a}, \mathrm{b} \in \mathbb{R}$ and $\mathrm{a}^{2}-4 \mathrm{~b}<0$.

Proof Let $\mathrm{f} \in \mathbb{R}[\mathrm{x}]$ be prime, and let $\alpha \in \mathbb{C}$ be a root of f (that $\alpha$ exists follows from Theorem 6.13). Then $x-\alpha$ is a factor of $f$ so that if $\alpha \in \mathbb{R}$, then $f=$ $\mathrm{c}(\mathrm{x}-\alpha)$ where $\mathrm{c} \in \mathbb{R}$ (since f is prime). But if $\alpha \notin \mathbb{R}$, then $\alpha \in \mathbb{C}$ and $\alpha^{*} \neq \alpha$ so that by Theorem 6.15, f has the factor

$$
(\mathrm{x}-\alpha)\left(\mathrm{x}-\alpha^{*}\right)=\mathrm{x}^{2}-\left(\alpha+\alpha^{*}\right) \mathrm{x}+\alpha \alpha^{*} .
$$

Writing $\alpha=u+i v$ we see that

$$
-\mathrm{a}=\alpha+\alpha^{*}=2 \mathrm{u} \in \mathbb{R}
$$

and

$$
\mathrm{b}=\alpha \alpha^{*}=u^{2}+v^{2} \in \mathbb{R}
$$

so that f has the form (up to a unit factor) $\mathrm{x}^{2}+\mathrm{ax}+\mathrm{b}$. Finally, note that

$$
\mathrm{a}^{2}-4 \mathrm{~b}=4 \mathrm{u}^{2}-4\left(\mathrm{u}^{2}+\mathrm{v}^{2}\right)=-4 \mathrm{v}^{2}<0 .
$$

## Exercises

1. Suppose $x^{n}+a_{n-1} x^{n-1}+\cdots+a_{1} x+a_{0} \in \mathbb{C}[x]$ has zeros $\alpha_{1}, \ldots, \alpha_{n}$. Prove that $a_{0}= \pm \alpha_{1} \cdots \alpha_{n}$ and $a_{n-1}=-\left(\alpha_{1}+\cdots+\alpha_{n}\right)$.
2. Let $\mathrm{V}_{\mathrm{n}} \subset \mathcal{F}[\mathrm{x}]$ denote the set of all polynomials of degree $\leq \mathrm{n}$, and let $\mathrm{a}_{0}$, $\mathrm{a}_{1}, \ldots, \mathrm{a}_{\mathrm{n}} \in \mathcal{F}$ be distinct.
(a) Show that $\mathrm{V}_{\mathrm{n}}$ is a vector space over $\mathcal{F}$ with basis $\left\{1, \mathrm{x}, \mathrm{x}^{2}, \ldots, \mathrm{x}^{\mathrm{n}}\right\}$, and hence that $\operatorname{dim} \mathrm{V}_{\mathrm{n}}=\mathrm{n}+1$.
(b) For each $\mathrm{i}=0, \ldots, \mathrm{n}$, define the mapping $\mathrm{T}_{\mathrm{i}}: \mathrm{V}_{\mathrm{n}} \rightarrow \mathcal{F}$ by $\mathrm{T}_{\mathrm{i}}(\mathrm{f})=\mathrm{f}\left(\mathrm{a}_{\mathrm{i}}\right)$. Show that the $T_{i}$ are linear functionals on $V_{n}$, i.e., that $T_{i} \in V_{n}$.
(c) For each $\mathrm{k}=0, \ldots, \mathrm{n}$ define the polynomial

$$
\begin{aligned}
p_{k}(x) & =\frac{\left(x-a_{0}\right) \cdots\left(x-a_{k-1}\right)\left(x-a_{k+1}\right) \cdots\left(x-a_{n}\right)}{\left(a_{k}-a_{0}\right) \cdots\left(a_{k}-a_{k-1}\right)\left(a_{k}-a_{k+1}\right) \cdots\left(a_{k}-a_{n}\right)} \\
& =\prod_{i \neq k}\left(\frac{x-a_{i}}{a_{k}-a_{i}}\right) \in V_{n} .
\end{aligned}
$$

Show that $\mathrm{T}_{\mathrm{i}}\left(\mathrm{p}_{\mathrm{j}}\right)=\delta_{\mathrm{ij}}$.
(d) Show that $p_{0}, \ldots, p_{n}$ forms a basis for $V_{n}$, and hence that any $f \in V_{n}$ may be written as

$$
f=\sum_{i=0}^{n} f\left(a_{i}\right) p_{i}
$$

(e) Now let $\mathrm{b}_{0}, \mathrm{~b}_{1}, \ldots, \mathrm{~b}_{\mathrm{n}} \in \mathcal{F}$ be arbitrary, and define $\mathrm{f}=\sum \mathrm{b}_{\mathrm{i}} \mathrm{p}_{\mathrm{i}}$. Show that $f\left(a_{j}\right)=b_{j}$ for $0 \leq j \leq n$. Thus there exists a polynomial of degree $\leq n$ that takes on given values at $\mathrm{n}+1$ distinct points.
(f) Now assume that $f, g \in \mathcal{F}[x]$ are of degree $\leq n$ and satisfy $f\left(a_{j}\right)=b_{j}=$ $g\left(a_{j}\right)$ for $0 \leq j \leq n$. Prove that $f=g$, and hence that the polynomial defined in part (e) is unique. This is called the Lagrange interpolation formula.
3. Suppose $\mathrm{Q} \subset \mathrm{M}_{2}(\mathbb{C})$ is the set of all complex matrices of the form

$$
\left(\begin{array}{cc}
z & w \\
-w^{*} & z^{*}
\end{array}\right) .
$$

(a) Prove that Q is a division ring (i.e., that the nonzero elements of Q form a multiplicative group). Q is called the ring of quaternions.
(b) Prove that Q is not a field.
(c) Prove that $\mathrm{x}^{2}+1 \in \mathrm{Q}[\mathrm{x}]$ has infinitely many roots in Q (where 1 denotes the unit element of Q , i.e., the 2 x 2 identity matrix).
4. Prove that $f, g \in \mathbb{C}[x]$ are relatively prime if and only if they have no root in common.
5. Let D be the differentiation operator defined in Problem 6.1.4, and suppose $f \in \mathbb{C}[x]$ is a monic polynomial. Prove that $f=\left(x-\alpha_{1}\right) \cdots\left(x-\alpha_{n}\right)$
where $\alpha_{1}, \ldots, \alpha_{\mathrm{n}} \in \mathbb{C}$ are distinct if and only if f and Df are relatively prime.
6. If $\mathrm{f} \in \mathcal{F}[\mathrm{x}]$ has a root $\alpha$, and $\mathrm{f}=(\mathrm{x}-\alpha)^{\mathrm{m}} \mathrm{g}$ where $\mathrm{g}(\alpha) \neq 0$, then $\alpha$ is said to be a root of multiplicity m . In other words, m is the largest integer such that $(\mathrm{x}-\alpha)^{\mathrm{m}} \mid \mathrm{f}$. Let $\alpha$ be a root of $\mathrm{f} \in \mathcal{F}[\mathrm{x}]$ and assume that $\operatorname{deg} \mathrm{f} \geq 1$. Show that the multiplicity of f is $>1$ if and only if $\operatorname{Df}(\alpha)=0$, and hence that the multiplicity of $\alpha$ is 1 if $\operatorname{Df}(\alpha) \neq 0$. (See Problem 6.1.4 for the definition of Df.)
7. Show that the following polynomials have no multiple roots in $\mathbb{C}$ (see the previous problem for the definition of multiplicity):
(a) $x^{4}+x$.
(b) $x^{5}-5 x+1$.
(c) $\mathrm{x}^{2}+\mathrm{bx}+\mathrm{c}$ where $\mathrm{b}, \mathrm{c} \in \mathbb{C}$ and $\mathrm{b}^{2}-4 \mathrm{c} \neq 0$.

### 6.5 THE FIELD OF QUOTIENTS

What we will do in this section is show how to construct a field out of the ring $\mathcal{F}[\mathrm{x}]$. Rather than talk about polynomials specifically, we use the fact that $\mathcal{F}[\mathrm{x}]$ is an integral domain and treat the problem on a more general footing.

Notice that the set $\mathbb{Z}$ of all integers has the property that if $a b=0$ for some $a, b \in \mathbb{Z}$, then either a or $b$ must equal 0 . Since $\mathbb{Z}$ is a ring, this shows that $\mathbb{Z}$ is in fact an integral domain. Also note though, for any $a \in \mathbb{Z}, \mathrm{a} \neq 1$, we have $\mathrm{a}^{-1}=1 / \mathrm{a} \notin \mathbb{Z}$, so that $\mathbb{Z}$ is not a field. However, if we enlarge the set $\mathbb{Z}$ to include all of the rational numbers, then we do indeed obtain a field. What this really entails is taking all pairs $a, b \in \mathbb{Z}$ and forming the object $a / b$ with the appropriate algebraic operations defined on it. In this particular case, we say that $\mathrm{a} / \mathrm{b}=\mathrm{c} / \mathrm{d}$ if and only if $\mathrm{ad}=\mathrm{bc}$, and we define the operations of addition and multiplication by

$$
\mathrm{a} / \mathrm{b}+\mathrm{c} / \mathrm{d}=(\mathrm{ad}+\mathrm{bc}) / \mathrm{bd}
$$

and

$$
(\mathrm{a} / \mathrm{b})(\mathrm{c} / \mathrm{d})=(\mathrm{ac}) /(\mathrm{bd}) .
$$

In order to generalize this result, we make the following definition. Let D be an integral domain (i.e., a commutative ring with no zero divisors), let $\mathrm{D}^{\prime}$ denote the set of all nonzero elements of D , and let $Q$ be the set of all ordered pairs

$$
Q=\left\{(\mathrm{a}, \mathrm{~b}) \in \mathrm{D} \times \mathrm{D}^{\prime}\right\}
$$

(You may think of $(\mathrm{a}, \mathrm{b})$ as the quotient $\mathrm{a} / \mathrm{b}$.) We define a relation $\sim$ on $Q$ by $(\mathrm{a}, \mathrm{b}) \sim(\mathrm{c}, \mathrm{d})$ if $\mathrm{ad}=\mathrm{bc}$, and we claim that this is an equivalence relation (for example, $2 / 3$ is "equivalent" to $8 / 12$ ). To prove this, we must verify the three requirements given in Section 0.3. First, for any $(\mathrm{a}, \mathrm{b}) \in Q$ we have $(\mathrm{a}, \mathrm{b}) \sim$ $(\mathrm{a}, \mathrm{b})$ since $\mathrm{ab}=\mathrm{ba}$. Next, for any $(\mathrm{a}, \mathrm{b}),(\mathrm{c}, \mathrm{d}) \in Q$ we see that $(\mathrm{a}, \mathrm{b}) \sim(\mathrm{c}, \mathrm{d})$ implies $\mathrm{ad}=\mathrm{bc}$, and hence $\mathrm{cb}=\mathrm{da}$ which thus implies $(\mathrm{c}, \mathrm{d}) \sim(\mathrm{a}, \mathrm{b})$. Finally, suppose ( $\mathrm{a}, \mathrm{b}$ ), ( $\mathrm{c}, \mathrm{d}),(\mathrm{e}, \mathrm{f}) \in Q$ where $(\mathrm{a}, \mathrm{b}) \sim(\mathrm{c}, \mathrm{d})$ and $(\mathrm{c}, \mathrm{d}) \sim(\mathrm{e}, \mathrm{f})$. Then $\mathrm{ad}=\mathrm{bc}$ and $\mathrm{cf}=\mathrm{de}$, and therefore $\mathrm{bde}=\mathrm{bcf}=\mathrm{adf}$. But D is commutative, and hence this is just afd $=$ bed. By assumption $d \neq 0$ so that, since $D$ is an integral domain, we must have $\mathrm{af}=$ be and thus $(\mathrm{a}, \mathrm{b}) \sim(\mathrm{e}, \mathrm{f})$.

We are now in a position to show that any integral domain can be enlarged in a similar manner to form a field. By way of terminology, if there is a one-to-one homomorphism (i.e., an isomorphism) of a ring $R$ into a ring $R^{\prime}$, then we say that R can be embedded in $\mathrm{R}^{\prime}$. Furthermore, if R and $\mathrm{R}^{\prime}$ are both rings with unit elements 1 and $1^{\prime}$ respectively, then we require that the embedding take 1 into $1^{\prime}$. The ring $\mathrm{R}^{\prime}$ is also called an extension of R .

The proof of the next theorem appears to be quite involved, but it is actually nothing more than a long series of simple steps.

Theorem 6.16 Every integral domain D can be embedded in a field.
Proof Let $\mathrm{D}^{\prime}$ be the nonzero elements of D , and let $Q$ be the set of all ordered pairs $(a, b) \in D \times D^{\prime}$ as defined above. We let $[a, b]$ denote the equivalence class in $Q$ of $(\mathrm{a}, \mathrm{b})$ as constructed above. In other words,

$$
[\mathrm{a}, \mathrm{~b}]=\left\{(\mathrm{x}, \mathrm{y}) \in \mathrm{D} \times \mathrm{D}^{\prime}:(\mathrm{a}, \mathrm{~b}) \sim(\mathrm{x}, \mathrm{y})\right\} .
$$

We claim that the set $\mathcal{F}_{\mathrm{D}}$ of all such equivalence classes forms a field. To prove this, we must first define addition and multiplication in $\mathcal{F}_{\mathrm{D}}$.

Guided by the properties of $\mathbb{Z}$, we define addition in $\mathcal{F}_{\mathrm{D}}$ by the rule

$$
[\mathrm{a}, \mathrm{~b}]+[\mathrm{c}, \mathrm{~d}]=[\mathrm{ad}+\mathrm{bc}, \mathrm{bd}] .
$$

Since D is an integral domain, we know that $\mathrm{bd} \neq 0$ for any nonzero $\mathrm{b}, \mathrm{d} \in \mathrm{D}$, and hence $[\mathrm{ad}+\mathrm{bc}, \mathrm{bd}] \in \mathcal{F}_{\mathrm{D}}$. We still must show that this addition is welldefined, i.e., if $[a, b]=\left[a^{\prime}, b^{\prime}\right]$ and $[c, d]=\left[c^{\prime}, d^{\prime}\right]$, then

$$
[\mathrm{a}, \mathrm{~b}]+[\mathrm{c}, \mathrm{~d}]=\left[\mathrm{a}^{\prime}, \mathrm{b}^{\prime}\right]+\left[\mathrm{c}^{\prime}, \mathrm{d}^{\prime}\right]
$$

This is equivalent to showing that

$$
[\mathrm{ad}+\mathrm{bc}, \mathrm{bd}]=\left[\mathrm{a}^{\prime} \mathrm{d}^{\prime}+\mathrm{b}^{\prime} \mathrm{c}^{\prime}, \mathrm{b}^{\prime} \mathrm{d}^{\prime}\right]
$$

or, alternatively, that

$$
(a d+b c) b^{\prime} d^{\prime}=b d\left(a^{\prime} d^{\prime}+b^{\prime} c^{\prime}\right)
$$

From $[a, b]=\left[a^{\prime}, b^{\prime}\right]$ we have $a b^{\prime}=b a^{\prime}$, and similarly $c d^{\prime}={d c^{\prime}}^{\prime}$. Therefore we indeed have

$$
\begin{aligned}
(a d+b c) b^{\prime} d^{\prime} & =a d b^{\prime} d^{\prime}+b c b^{\prime} d^{\prime}=a b^{\prime} d d^{\prime}+b b^{\prime} c d^{\prime}=b a^{\prime} d d^{\prime}+b b^{\prime} d c^{\prime} \\
& =b d a^{\prime} d^{\prime}+b d b^{\prime} c^{\prime}=b d\left(a^{\prime} d^{\prime}+b^{\prime} c^{\prime}\right)
\end{aligned}
$$

Since D is commutative, it should be clear that if $\mathrm{c} \neq 0$ then

$$
[\mathrm{a}, \mathrm{~b}]=[\mathrm{ac}, \mathrm{bc}]=[\mathrm{ca}, \mathrm{cb}]
$$

Therefore

$$
[\mathrm{a}, \mathrm{~b}]+[0, \mathrm{c}]=[\mathrm{ac}+\mathrm{b} 0, \mathrm{bc}]=[\mathrm{ac}, \mathrm{bc}]=[\mathrm{a}, \mathrm{~b}]
$$

so that $[0, c]$ is a zero element for addition. We now see that

$$
[\mathrm{a}, \mathrm{~b}]+[-\mathrm{a}, \mathrm{~b}]=[\mathrm{ab}-\mathrm{ba}, \mathrm{bb}]=[0, \mathrm{~b}]
$$

and hence $[-a, b]$ is the negative of $[a, b]$. The reader should now have no trouble showing that $\mathcal{F}_{\mathrm{D}}$ is an abelian group under addition.

To complete our ring structure, we define multiplication in $\mathcal{F}_{\mathrm{D}}$ by the rule

$$
[\mathrm{a}, \mathrm{~b}][\mathrm{c}, \mathrm{~d}]=[\mathrm{ac}, \mathrm{bd}] .
$$

As was the case with addition, the fact that $b, d \neq 0$ means that $b d \neq 0$, and hence the product is an element of $\mathcal{F}_{\mathrm{D}}$. We leave it to the reader to show that the product is well-defined (see Exercise 6.5.1). It is also easy to see that [ $\mathrm{x}, \mathrm{x}]$ is a unit element in $\mathcal{F}_{\mathrm{D}}$ for any nonzero $\mathrm{x} \in \mathrm{D}$, and that the nonzero elements of $\mathcal{F}_{\mathrm{D}}$ (i.e., those of the form $[\mathrm{a}, \mathrm{b}]$ with $\mathrm{a} \neq 0$ ) form an abelian group under multiplication with the inverse of an element $[a, b]$ given $b y[a, b]^{-1}=$ $[\mathrm{b}, \mathrm{a}]$ (where $[\mathrm{b}, \mathrm{a}] \in \mathcal{F}_{\mathrm{D}}$ since $\mathrm{a} \neq 0$ ).

As to the ring axioms, we show only one of the distributive laws, and leave the others to the reader. This will complete the proof that $\mathcal{F}_{\mathrm{D}}$ forms a field. If $[a, b],[c, d],[e, f] \in \mathcal{F}_{D}$, then

$$
\begin{aligned}
{[a, b]([c, d]+[e, f]) } & =[a, b][c f+d e, d f]=[a c f+a d e, b d f] \\
& =[b(a c f+a d e), b b d f]=[(a c)(b f)+(b d)(a e),(b d)(b f)] \\
& =[a c, b d]+[a e, b f]=[a, b][c, d]+[a, b][e, f]
\end{aligned}
$$

What we have accomplished up to this point is the construction of a field $\mathcal{F}_{\mathrm{D}}$ from an arbitrary integral domain D . It still must be shown that D can be embedded in $\mathcal{F}_{\mathrm{D}}$. As was noted above, for any nonzero $\mathrm{x}, \mathrm{y} \in \mathrm{D}$ we have $[a x, x]=[a y, y]$ because $(a x) y=x(a y)$. This means that we can denote the element $[\mathrm{ax}, \mathrm{x}] \in \mathcal{F}_{\mathrm{D}}$ by $[\mathrm{a}, 1]$. (It is important to realize that the ring D does not necessarily contain a unit element, so that there need not exist a unit element $1 \in \mathrm{D}$. What we have just done is define the element $[\mathrm{a}, 1] \in \mathcal{F}_{\mathrm{D}}$. Everything that follows in the remainder of this proof holds if we replace the symbol 1 by an arbitrary nonzero element $x \in D$.)

We now define the mapping $\phi: \mathrm{D} \rightarrow \mathcal{F}_{\mathrm{D}}$ by $\phi(\mathrm{a})=[\mathrm{a}, 1]$ for all $\mathrm{a} \in \mathrm{D}$. If $\phi(a)=\phi\left(a^{\prime}\right)$, then $[a, 1]=\left[a^{\prime}, 1\right]$ so that $a 1=1 a^{\prime}$ and hence $a=a^{\prime}$, thus proving that $\phi$ is one-to-one. (As we just mentioned, the symbol 1 could actually be replaced by any $x \neq 0$ since the fact that $D$ is an integral domain then says that if $a x=x a^{\prime}$, then $x\left(a-a^{\prime}\right)=0$ which also implies that $a=a^{\prime}$.) To finish the proof, we need only show that $\phi$ is a homomorphism. But for any $a, b \in D$ we have

$$
\phi(a+b)=[a+b, 1]=[a 1+b 1,1 \cdot 1]=[a, 1]+[b, 1]=\phi(a)+\phi(b)
$$

and

$$
\phi(\mathrm{ab})=[\mathrm{ab}, 1]=[\mathrm{ab}, 1 \cdot 1]=[\mathrm{a}, 1][\mathrm{b}, 1]=\phi(\mathrm{a}) \phi(\mathrm{b}) .
$$

The field $\mathcal{F}_{\mathrm{D}}$ constructed in this theorem is usually called the field of quotients of D . If we start with the ring of integers $\mathbb{Z}$, then this construction yields the rational field $\mathbb{Q}$. While we have shown that any integral domain D can be embedded in its field of quotients, there can be other fields in which D can also be embedded. However, it can be shown that $\mathcal{F}_{\mathrm{D}}$ is the "smallest" field in which D can be embedded (see Exercise 6.5.2).

## Exercises

1. Referring to the proof of Theorem 6.16 , show that the product in $\mathcal{F}_{\mathrm{D}}$ is well-defined.
2. Show that $\mathcal{F}_{\mathrm{D}}$ is the smallest field in which an integral domain D can be embedded. In other words, show that if $\mathcal{K}$ is any field containing an integral domain isomorphic to D , then $\mathcal{K}$ contains a field isomorphic to $\mathcal{F}_{\mathrm{D}}$.
[Hint: For simplicity, assume that D is actually a subring of $\mathcal{K}$. Now, for any $\mathrm{a}, \mathrm{b} \in \mathrm{D}$ show that the map $\phi: \mathcal{F}_{\mathrm{D}} \rightarrow \mathcal{K}$ defined by $\phi([\mathrm{a}, \mathrm{b}])=\mathrm{ab}^{-1}$ is one-to-one and preserves addition and multiplication. Thus $\phi$ is an isomorphism of $\mathcal{F}_{\mathrm{D}}$ onto a subfield of $\mathcal{K}$.]
3. Show that $\mathcal{F}_{\mathrm{D}}$ obeys all of the axioms for a ring.

### 6.6 POLYNOMIALS OVER FINITE FIELDS *

With very few exceptions (e.g., Exercise 1.5.15), the fields we have been using (such as $\mathbb{R}$ and $\mathbb{C}$ ) contain an infinite number of elements. However, it is also possible to construct many fields that contain only a finite number of elements. This section is meant to be only an introduction to the theory of finite fields.

Recall from Example 1.11 that two integers $a, b \in \mathbb{Z}$ are said to be congruent modulo $n$ (where $n \in \mathbb{Z}^{+}$) if $n \mid(a-b)$, and we write this as

$$
\mathrm{a} \equiv \mathrm{~b}(\bmod \mathrm{n}) .
$$

We also saw in Exercise 1.5 .2 that for each $\mathrm{n} \in \mathbb{Z}^{+}$, this defines an equivalence relation on $\mathbb{Z}$ that decomposes $\mathbb{Z}$ into $n$ distinct congruence classes. We shall denote the congruence class (for a fixed $n$ ) of an integer $k \in$ $\mathbb{Z}$ by $[\mathrm{k}]$. If there is any possible ambiguity as to the value of $n$ under discussion, we will write $[k]_{n}$. For example, if $n=5$ we have

$$
[2]=[7]=[-33]=\{\ldots,-8,-3,2,7,12, \ldots\}
$$

We also refer to any of the integers in [k] as a representative of [k]. For example, 2 is the smallest positive representative of the class [7] (or the class [-33] etc.). We emphasize that [ $k$ ] is a subset of $\mathbb{Z}$ for each $k$.

From Example 1.11, we say that the collection

$$
\{[0],[1],[2], \ldots,[n-1]\}
$$

forms a complete set of congruence classes modulo n , and we denote this collection by $\mathbb{Z}_{\mathrm{n}}$. We first show that $\mathbb{Z}_{\mathrm{n}}$ can be made into an abelian group. For any [a], $[\mathrm{b}] \in \mathbb{Z}_{\mathrm{n}}$ we define a group "addition" operation $[\mathrm{a}] \oplus[\mathrm{b}] \in \mathbb{Z}_{\mathrm{n}}$ by

$$
[\mathrm{a}] \oplus[\mathrm{b}]=[\mathrm{a}+\mathrm{b}] .
$$

(Note that the symbol $\oplus$ is used in an entirely different context when we talk about direct sums.) It should be obvious that [0] will serve as the additive identity element since $[a] \oplus[0]=[a+0]=[a]$ and $[0] \oplus[a]=[0+a]=[a]$. Noting that, for example with $\mathrm{n}=5$ again, we have $[3]=[18]$ and $[4]=[-1]$, we must be sure that $[3] \oplus[4]=[18] \oplus[-1]$. Clearly this is true because $[3] \oplus$ $[4]=[7]=[2]$ and $[18] \oplus[-1]=[17]=[2]$. In other words, we must be sure that this addition operation is well-defined. That this is in fact the case is included in the next theorem.

Theorem 6.17 The set $\mathbb{Z}_{\mathrm{n}}$ is an abelian group with respect to the operation $\oplus$ defined above.

Proof We leave it to the reader to show that $\oplus$ is indeed well-defined (see Exercise 6.6.1). As to the group properties, we prove associativity, leaving the rest of the proof to the reader (see Exercise 6.6.2). We have

$$
\begin{aligned}
{[a] \oplus([b] \oplus[c]) } & =[a] \oplus[b+c]=[a+(b+c)]=[(a+b)+c] \\
& =[a+b] \oplus[c]=([a] \oplus[b]) \oplus[c] .
\end{aligned}
$$

By virtue of this theorem, $\mathbb{Z}_{\mathrm{n}}$ is called the group of integers modulo $\mathbf{n}$ (or simply $\bmod \mathrm{n}$. We can also define another operation on $\mathbb{Z}_{\mathrm{n}}$ that is analogous to multiplication. Thus, we define the "multiplication" operation $\otimes$ on $\mathbb{Z}_{\mathrm{n}}$ by

$$
[\mathrm{a}] \otimes[\mathrm{b}]=[\mathrm{ab}] .
$$

(Again, this symbol should not be confused with the tensor product to be introduced in Chapter 11.) For example, if $\mathrm{n}=6$ we have $[2] \otimes[5]=[10]=$ $[4]$ and $[3] \otimes[-4]=[-12]=[0]$. The closest analogue to Theorem 6.17 that we have for $\otimes$ is the following.

Theorem 6.18 The operation $\otimes$ defined above on $\mathbb{Z}_{\mathrm{n}}$ is well-defined, obeys the associative and commutative laws, and has [1] as the identity element.

Proof See Exercise 6.6.3.

Since [1] is the identity element for $\otimes$, it is easy to see that [0] has no multiplicative inverse in $\mathbb{Z}_{\mathrm{n}}$, and hence $\mathbb{Z}_{\mathrm{n}}$ can not possibly form a group under $\otimes$. Let us denote the set $\mathbb{Z}_{\mathrm{n}}-[0]=\{[1],[2], \ldots,[n-1]\}$ by $\mathbb{Z}_{\mathrm{n}}{ }^{+}$. It turns out that for some (but not all) values of $\mathrm{n}, \mathbb{Z}_{\mathrm{n}}{ }^{+}$will in fact form a group with respect to $\otimes$. We will leave specific examples of this to the exercises at the end of this section.

With the operations $\oplus$ and $\otimes$ defined, it is now easy to see that $\mathbb{Z}_{\mathrm{n}}$ actually forms a commutative ring. All we must do is verify the axioms given in Section 1.4. We will show that the first half of axiom (R8) is obeyed, and leave it to the reader to verify the rest of the ring axioms (see Exercise 6.6.4). We therefore have

$$
\begin{aligned}
{[a] \otimes([b] \oplus[c]) } & =[a] \otimes[b+c]=[a(b+c)]=[a b+a c] \\
& =[a b] \oplus[a c]=([a] \otimes[b]) \oplus([a] \otimes[c]) .
\end{aligned}
$$

Now consider the ring $\mathbb{Z}_{\mathrm{n}}$ and assume that n is not prime. Then we may write $\mathrm{n}=\mathrm{rs}$ where $\mathrm{r}, \mathrm{s}>1$. But then $[\mathrm{r}] \otimes[\mathrm{s}]=[\mathrm{rs}]=[\mathrm{n}]=[0]$ where $[\mathrm{r}],[\mathrm{s}] \neq$ [0]. Since [0] is the zero element of $\mathbb{Z}_{n}$, this shows that $\mathbb{Z}_{n}$ is not an integral domain if n is not prime. On the other hand, suppose that $\mathrm{n}=\mathrm{p}$ is prime. We claim that $\mathbb{Z}_{\mathrm{p}}$ is an integral domain. For, suppose $[\mathrm{a}] \in \mathbb{Z}_{\mathrm{p}}$ and $[\mathrm{a}] \neq[0]$. We may assume that a is the smallest positive representative of the equivalence class [a], and hence $\mathrm{a}<\mathrm{p}$. Now assume that $[\mathrm{b}] \in \mathbb{Z}_{\mathrm{p}}$ is such that $[\mathrm{a}] \otimes[\mathrm{b}]=$ $[\mathrm{ab}]=[0]$ (where we again choose $\mathrm{b}<\mathrm{p}$ to be the smallest positive representative of the class [b]). Then by definition we have plab. But p is prime so (by Theorem 0.9) this implies that either $p \mid a$ or $p \mid b$. Since $a<p$, it is impossible for p to divide a , and therefore $\mathrm{p} \mid \mathrm{b}$. Since $0 \leq b<\mathrm{p}$, we must have $\mathrm{b}=0$, and thus $\mathbb{Z}_{\mathrm{p}}$ is an integral domain. This proves the next result.

Theorem 6.19 The ring $\mathbb{Z}_{\mathrm{n}}$ is an integral domain if and only if n is prime.
Noting that $\mathbb{Z}_{n}$ consists of $n$ equivalence classes, we now claim that $\mathbb{Z}_{n}$ is in fact a field if $n$ is prime. This is an immediate consequence of the following general result. Recall that a field is a commutative ring with identity element in which the nonzero elements form a multiplicative group (i.e., a commutative division ring). Furthermore, any field is necessarily an integral domain (see Exercise 1.5.6 or 1.5.12).

Theorem 6.20 Every finite integral domain is a field.

Proof Let D be a finite integral domain (which is commutative by definition). We must show that $1 \in \mathrm{D}$, and that every nonzero a $\in \mathrm{D}$ has a multiplicative inverse that is also in D . In other words, we must show that for every nonzero $a \in D$ there exists $b \in D$ such that $a b=1 \in D$. Let $\left\{x_{1}, \ldots\right.$, $\left.\mathrm{x}_{\mathrm{n}}\right\}$ denote all the elements of D , and consider the set $\left\{\mathrm{ax}_{1}, \ldots, \mathrm{ax}_{\mathrm{n}}\right\}$ where $a \in D$ and $a \neq 0$. If $a x_{i}=a x_{j}$ for $i \neq j$, then $a\left(x_{i}-x_{j}\right)=0$ which (since $D$ has no zero divisors) implies that $\mathrm{x}_{\mathrm{i}}=\mathrm{x}_{\mathrm{j}}$, contradicting the assumption that $\mathrm{i} \neq \mathrm{j}$. Thus
$\mathrm{ax}_{1}, \ldots, \mathrm{ax}_{\mathrm{n}}$ are all distinct. Since D contains n elements, it follows that in fact we have $D=\left\{\mathrm{ax}_{1}, \ldots, \mathrm{ax}_{\mathrm{n}}\right\}$. In other words, every $\mathrm{y} \in \mathrm{D}$ can be written in the form $\mathrm{ax}_{\mathrm{i}}=\mathrm{x}_{\mathrm{i}} \mathrm{a}$ for some $\mathrm{i}=1, \ldots$, n . In particular, we must have $\mathrm{a}=$ $\mathrm{ax}_{\mathrm{i}_{0}}$ for some $\mathrm{i}_{0}=1, \ldots, n$. Then for any $\mathrm{y}=\mathrm{x}_{\mathrm{i}} \mathrm{a} \in \mathrm{D}$ we have

$$
\mathrm{yx}_{\mathrm{i}_{0}}=\left(\mathrm{x}_{\mathrm{i}} \mathrm{a}\right) \mathrm{x}_{\mathrm{i}_{0}}=\mathrm{x}_{\mathrm{i}}\left(\mathrm{ax}_{\mathrm{i}_{0}}\right)=\mathrm{x}_{\mathrm{i}} \mathrm{a}=\mathrm{y}
$$

so that $\mathrm{x}_{\mathrm{i}_{0}}$ may be taken as the identity element 1 in D. Finally, since we have now shown that $1 \in D$, it follows that $1=\operatorname{ax}_{\mathrm{j}}$ for some particular $\mathrm{j}=1, \ldots$, . Defining $\mathrm{b}=\mathrm{x}_{\mathrm{j}}$ yields $1=\mathrm{ab}$ and completes the proof.

Corollary $\mathbb{Z}_{\mathrm{n}}$ is a field if and only if n is prime.
We now turn our attention to the question of whether there exist any finite fields that do not contain a prime number of elements. As with groups, we refer to the number of elements in a finite field as its order. This is not surprising since any field is a ring, and any ring is an additive group. We will frequently denote a finite field by F rather than by $\mathcal{F}$.

Example 6.10 Let $\mathrm{S}_{2}(\mathcal{F}) \subset \mathrm{M}_{2}(\mathcal{F})$ denote the set of all matrices of the form

$$
\left(\begin{array}{rr}
x & y \\
-y & x
\end{array}\right)
$$

We will show that $\mathrm{S}_{2}(\mathcal{F})$ is a field when $\mathcal{F}=\mathbb{Z}_{3}$ but not when $\mathcal{F}=\mathbb{Z}_{5}$.
We leave it as a simple exercise for the reader to show that if $\mathrm{A}, \mathrm{B} \in$ $\mathrm{S}_{2}(\mathcal{F})$, then $\mathrm{A}+\mathrm{B}$ and AB are also in $\mathrm{S}_{2}(\mathcal{F})$. Furthermore, $\mathrm{AB}=\mathrm{BA}$ so that $\mathrm{S}_{2}(\mathcal{F})$ is commutative. Note that $\mathrm{S}_{2}(\mathcal{F})$ also contains the zero and identity matrices, and if $\mathrm{A} \in \mathrm{S}_{2}(\mathcal{F})$, then so is -A . Thus $\mathrm{S}_{2}(\mathcal{F})$ is easily seen to be a subring of $\mathrm{M}_{2}(\mathcal{F})$. We now consider the problem of inverses. If

$$
A=\left(\begin{array}{rr}
x & y \\
-y & x
\end{array}\right)
$$

has inverse

$$
A^{-1}=\left(\begin{array}{rr}
a & b \\
-b & a
\end{array}\right)
$$

then $\mathrm{AA}^{-1}=\mathrm{I}$ yields the two simultaneous equations

$$
\begin{aligned}
& a x-b y=1 \\
& a y+b x=0 .
\end{aligned}
$$

Since $\mathrm{x}, \mathrm{y}, \mathrm{a}, \mathrm{b} \in \mathcal{F}$ we can (formally) solve these for $\mathrm{a} a \mathrm{and} \mathrm{b}$ to obtain

$$
\begin{aligned}
& a=x\left(x^{2}+y^{2}\right)^{-1} \\
& b=-y\left(x^{2}+y^{2}\right)^{-1} .
\end{aligned}
$$

The element $\left(x^{2}+y^{2}\right)^{-1}$ will exist as long as $x^{2}+y^{2} \neq 0$.
In $\mathbb{Z}_{3}$ we have $\mathrm{x}, \mathrm{y}=\{0,1,2\}$ and it is easy to see by direct calculation that $x^{2}+y^{2} \neq 0$ as long as $(x, y) \neq(0,0)$. For example, if $x=1$ and $y=2$ we have $x^{2}+y^{2}=1+1=2$. Thus every nonzero matrix in $S_{2}\left(\mathbb{Z}_{3}\right)$ is invertible, and hence $S_{2}\left(\mathbb{Z}_{3}\right)$ is a field with 9 elements.

On the other hand, in $\mathbb{Z}_{5}$ we see that $1^{2}+2^{2}=0$ so that the matrix with $x=$ 1 and $y=2$ is not invertible, and hence $S_{2}\left(\mathbb{Z}_{5}\right)$ is not a field. //

Example 6.11 Since $\mathbb{Z}_{3}$ is a field, we can consider polynomials in $\mathbb{Z}_{3}[x]$. These may be used to generate a field of order 9 as follows. We define the set $\mathrm{F}_{9} \subset \mathbb{Z}_{3}[\mathrm{x}]$ consisting of the nine polynomials of degree $\leq 1$ with coefficients in $\mathbb{Z}_{3}$ :

$$
\mathrm{F}_{9}=\{0,1,2, \mathrm{x}, \mathrm{x}+1, \mathrm{x}+2,2 \mathrm{x}, 2 \mathrm{x}+1,2 \mathrm{x}+2\} .
$$

It is easy to see that this set is closed under ordinary polynomial addition. For example, remembering that our scalars lie in $\mathbb{Z}_{3}$ we have $(x+1)+(2 x+1)=$ 2. However, we must be careful in defining multiplication. This is because, for example, $(x+1)(2 x+1)=2 x^{2}+1 \notin F_{9}$ even though it is in $\mathbb{Z}_{3}[x]$. To ensure that multiplication is closed, we multiply as usual in $\mathbb{Z}_{3}[x]$ and then reduce modulo $x^{2}+1$. In other words, we subtract off multiples of $x^{2}+1$. For example, we have

$$
\begin{aligned}
(x+1)(2 x+1) & =2 x^{2}+1 & & \left(\text { in } \mathbb{Z}_{3}[x]\right) \\
& =2\left(x^{2}+1\right)+2 & & \\
& =2 & & \left(\text { in } \mathrm{F}_{9}\right) .
\end{aligned}
$$

As another example,

$$
\begin{aligned}
(2 x+1)(x) & =2 x^{2}+x & & \left(\text { in } \mathbb{Z}_{3}[x]\right) \\
& =2\left(x^{2}+1\right)+(x+1) & & \\
& =x+1 & & \left(\text { in } \mathrm{F}_{9}\right) .
\end{aligned}
$$

Using the constant polynomials 0 and 1 as the 0 and 1 elements of a ring, it is easy to show that $\mathrm{F}_{9}$ forms a commutative ring. That $\mathrm{F}_{9}$ in fact is a field follows from the observation that each nonzero element of $\mathrm{F}_{9}$ has the inverse shown below:

| Element: | 1 | 2 | x | $\mathrm{x}+1$ | $\mathrm{x}+2$ | 2 x | $2 \mathrm{x}+1$ | $2 \mathrm{x}+2$ |
| ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: | ---: |
| Inverse: | 1 | 2 | 2 x | $\mathrm{x}+2$ | $\mathrm{x}+1$ | x | $2 \mathrm{x}+2$ | $2 \mathrm{x}+1$ |

We leave verification of these facts to the reader. //

Let $G$ be any group, and let e be the identity element of $G$. If there exists an element $x \in G$ such that every element in $G$ is a power of $x$, then $G$ is said to be a cyclic group generated by x . The cyclic group generated by x is usually denoted by $\langle x\rangle$. If we consider the set of all powers of the generator $x$, then this set will consist of either all distinct elements, or else some elements will be repeated. In the case of repeated elements, there will exist a positive integer $m>0$ such that $x^{m}=e$ while no smaller nonzero power of $x$ is also equal to e. For example, if $\mathrm{i} \leq \mathrm{j}$ and $\mathrm{x}^{i}=\mathrm{x}^{j}$, we have $e=x^{j}\left(x^{i}\right)^{-1}=x^{j} x^{-i}=x^{j-i}$. Then let $m$ be the smallest nonzero value of all such differences $j-i$.

We say that $G=\left\{e, x, x^{2}, \ldots, x^{m-1}\right\}$ is a cyclic group of order $m$, and we denote this group by $\mathrm{C}_{\mathrm{m}}$. We are using the letter e to denote the identity element of a general group so that we may distinguish between the multiplicative identity (usually written as 1 ) of some groups and the additive identity (usually written as 0 ) of other groups. Note that given any $k \in \mathbb{Z}$ we may write $\mathrm{k}=\mathrm{qm}+\mathrm{r}$ (where $0 \leq \mathrm{r}<\mathrm{m}$ ), and hence

$$
x^{k}=\left(x^{m}\right)^{q} x^{r}=e^{q} x^{r}=x^{r} .
$$

Thus all powers of x can indeed be written in terms of the first m powers of x .
In the case that all powers of x are distinct, then the group

$$
G=\left\{\ldots, x^{-2}, x^{-1}, e, x, x^{2}, \ldots\right\}
$$

contains an infinite number of elements. We denote such an infinite cyclic group by $\mathrm{C}_{\infty}$. Of importance to us is the fact that many apparently diverse groups are isomorphic to cyclic groups (either finite or infinite). We will see a simple example of this below.

If $x \in G$ and $m>0$ is the smallest integer such that $x^{m}=1$, then $m$ is called the order of x , and will also be denoted by $\mathrm{o}(\mathrm{x})$. (This should not be confused with the order of $G$ which is the number of elements in G. It is for this reason that $\mathrm{o}(\mathrm{G})$ is frequently denoted by $|\mathrm{G}|$.) If $\mathrm{o}(\mathrm{x}) \leq \mathrm{o}(\mathrm{G})$, then it is
easy to see that $\langle\mathrm{x}\rangle$ is simply a subgroup of G , and furthermore that $\mathrm{o}(\mathrm{x})=$ $o(\langle\mathrm{x}\rangle)$ (see Exercise 6.6.9).

Example 6.12 Let us show that the set of integers $\mathbb{Z}$ under addition is isomorphic to $\mathrm{C}_{\infty}=\langle\mathrm{x}\rangle$. We define the mapping $\phi: \mathbb{Z} \rightarrow \mathrm{C}_{\infty}$ by $\phi(\mathrm{n})=\mathrm{x}^{\mathrm{n}}$. Clearly $\phi$ is a homomorphism of groups since

$$
\phi(\mathrm{n}+\mathrm{m})=\mathrm{x}^{\mathrm{n}+\mathrm{m}}=\mathrm{x}^{\mathrm{n}} \mathrm{x}^{\mathrm{m}}=\phi(\mathrm{n}) \phi(\mathrm{m}) .
$$

(Note that $\mathrm{n}+\mathrm{m}$ in $\phi(\mathrm{n}+\mathrm{m})$ denotes the "product" of two group elements in $\mathbb{Z}$ while $\phi(\mathrm{n}) \phi(\mathrm{m})$ denotes the product of two elements in $\mathrm{C}_{\infty}$.) It should also be obvious that $\phi$ is surjective because every $\mathrm{x}^{\mathrm{k}} \in \mathrm{C}_{\infty}$ is just the image of $\mathrm{k} \in$ $\mathbb{Z}$. Finally, $\phi$ is injective since $\operatorname{Ker} \phi=\{0\}$. We have thus constructed an isomorphism of $\mathbb{Z}$ onto $\mathrm{C}_{\infty}$.

We leave it to the reader to show that $\mathbb{Z}_{\mathrm{m}}$ is isomorphic to $\mathrm{C}_{\mathrm{m}}$ (see Exercise 6.6.7). //

Let $\mathcal{F}$ be any field. Since $\mathcal{F}$ is closed under addition and contains the multiplicative identity element 1 , we see that for any $n \in \mathbb{Z}^{+}$, the sum $1+\cdots+1$ of n 1 's is also in $\mathcal{F}$. For example, $2=1+1 \in \mathcal{F}$ as is $3=1+1+1$ and so on. (However, it is important to stress that in an arbitrary field these elements need not be distinct.) Therefore the positive integers form a (not necessarily infinite) cyclic subgroup $\langle 1\rangle$ of the additive group of $\mathcal{F}$. In other words

$$
\langle 1\rangle=\{0,1,2, \ldots\}
$$

where each $\mathrm{n} \in\langle 1\rangle$ denotes $1+\cdots+1$ ( n times) in $\mathcal{F}$.
Now consider the special case where F is a finite field. Note that since $\langle 1\rangle$ is a subgroup of F , Theorem 1.9 tells us that $\mathrm{o}(\langle 1\rangle) \mid \mathrm{o}(\mathrm{F})$. The number $\mathrm{o}(\langle 1\rangle)$ is called the characteristic of F (see Section 1.5). For example, if $\mathrm{F}=\mathbb{Z}_{\mathrm{p}}$, then $\langle 1\rangle=\{0,1,2, \ldots, p-1\}=\mathbb{Z}_{p}$ and hence the characteristic of $\mathbb{Z}_{p}$ is $o(\langle 1\rangle)=$ $\mathrm{p}=\mathrm{o}(\mathrm{F})$. The characteristic of an infinite field may or may not be finite.

Example 6.13 Let us show that the characteristic of any finite field must be a prime number (if it is nonzero). By definition, the characteristic of $F$ is the smallest $\mathrm{m} \in \mathbb{Z}^{+}$such that $\mathrm{m} \cdot 1=1+\cdots+1=0$. If $\mathrm{m} \neq 0$ is not prime, then we may write $\mathrm{m}=\mathrm{rs}$ for some integers $0<\mathrm{r}<\mathrm{m}$ and $0<\mathrm{s}<\mathrm{m}$. But then we have $\mathrm{rs}=0$ which implies (since we are in a field) that either $\mathrm{r}=0$ or $\mathrm{s}=0$. In either case this contradicts the definition of m as the least positive integer such that $\mathrm{m}=0$. Thus m must be prime. Note that this proof actually applies to any finite integral domain with an identity element.

We now wish to prove that if a finite field exists, then its order must be a prime power. For example, we have seen that $\mathbb{Z}_{n}$ is a field if and only if $n$ is prime, and the field $\mathrm{F}_{9}$ discussed above is of order $9=3^{2}$. Our claim will follow as an immediate corollary of the next theorem. The reader should recall that the direct product of two groups was defined in Exercise 1.1.5. The direct product of a finite number of groups follows by an obvious induction argument.

Example 6.14 Consider the cyclic groups $C_{2}=\langle x\rangle=\{1, x\}$ and $C_{3}=\langle y\rangle=$ $\left\{1, \mathrm{y}, \mathrm{y}^{2}\right\}$. Then the product $\mathrm{C}_{2} \times \mathrm{C}_{3}$ consists of the six elements

$$
C_{2} \times C_{3}=\left\{(1,1),(1, y),\left(1, y^{2}\right),(x, 1),(x, y),\left(x, y^{2}\right)\right\} .
$$

To show that this product group is isomorphic to $\mathrm{C}_{6}$, let $\mathrm{z}=(\mathrm{x}, \mathrm{y}) \in \mathrm{C}_{2} \times \mathrm{C}_{3}$. Then by definition of the group product in $\mathrm{C}_{2} \times \mathrm{C}_{3}$ we have $\mathrm{z}^{2}=\left(1, \mathrm{y}^{2}\right), \mathrm{z}^{3}=$ $(x, 1), z^{4}=(1, y), z^{5}=\left(x, y^{2}\right), z^{6}=(1,1)$. Therefore $C_{2} \times C_{3}=\left\{z, z^{2}, \ldots, z^{6}\right\}$ which is just a cyclic group of order 6. //

Theorem 6.21 If $F$ is a finite field of characteristic $p$, then for some $r \geq 1$, the additive group of F is isomorphic to the r -fold direct product $\left(\mathrm{C}_{\mathrm{p}}\right)^{\mathrm{r}}$. Thus $\mathrm{o}(\mathrm{F})=\mathrm{p}^{\mathrm{r}}$.

Proof We leave it to the reader to show that $\mathbb{Z}_{\mathrm{p}}$ is isomorphic to a subfield of F , and hence that F may be considered to be a vector space V over the field $\mathbb{Z}_{\mathrm{p}}$. Since F is finite, $\mathrm{r}=\operatorname{dim} \mathrm{V}$ must also be finite. By the corollary to Theorem 2.8, V is isomorphic to $\left(\mathbb{Z}_{\mathrm{p}}\right)^{\mathrm{r}}=\mathbb{Z}_{\mathrm{p}} \times \cdots \times \mathbb{Z}_{\mathrm{p}}$. Noting that $\mathbb{Z}_{\mathrm{p}}=$ $\{0,1, \ldots, \mathrm{p}-1\}=\langle 1\rangle$ is isomorphic to the (additive) cyclic group $\mathrm{C}_{\mathrm{p}}$, it follows that V is isomorphic to $\left(\mathrm{C}_{\mathrm{p}}\right)^{\mathrm{r}}$ (i.e., the set of all r-tuples of field elements). Thus V has $\mathrm{p}^{\mathrm{r}}$ elements.

Corollary The order of a finite field must be a prime power.
Proof This follows from Example 6.13 and Theorem 6.21.
We now comment briefly on the construction of finite fields. What we shall do is generalize the procedure demonstrated in Example 6.11 where we constructed the field $\mathrm{F}_{9}$. Recall that the problem came in defining a closed multiplication in $\mathbb{Z}_{3}[x]$. A more general way to view the solution to this problem is to define an equivalence relation $\approx o n \mathbb{Z}_{3}[\mathrm{x}]$ by the requirement that $\mathrm{a} \approx \mathrm{b}$ if $\left(\mathrm{x}^{2}+1\right) \mid(\mathrm{a}-\mathrm{b})$. Note that $\mathrm{x}^{2}+1$ is prime in $\mathbb{Z}_{3}[\mathrm{x}]$. We may now
take the elements of $\mathrm{F}_{9}$ to be the equivalence classes of the nine polynomials that were previously used in Example 6.11 to define F9. Another way to say this is that these nine polynomials of degree $\leq 1$ form a complete set of representatives of the classes. In this case, addition and multiplication are defined as expected by

$$
[\mathrm{a}]+[\mathrm{b}]=[\mathrm{a}+\mathrm{b}]
$$

and

$$
[\mathrm{a}][\mathrm{b}]=[\mathrm{ab}] .
$$

Note that if $\left(x^{2}+1\right) \mid(a-b)$, then the remainder of a divided by $x^{2}+1$ must be the same as the remainder of $b$ divided by $x^{2}+1$. Since there are only a finite number of polynomials in $\mathbb{Z}_{3}[x]$, there can be only a finite number of distinct remainders, and the degree of each remainder must be less than that of $x^{2}+1$. Referring to Theorem 6.7 and its proof, a moments thought should convince you that all we are doing is considering the cosets $\mathbb{Z}_{3}[x] /\left(x^{2}+1\right)$ where $\left(x^{2}+1\right)$ denotes the principal ideal generated by $k=x^{2}+1 \in \mathbb{Z}_{3}[x]$. This is because any $\mathrm{p} \in \mathbb{Z}_{3}[\mathrm{x}] /(\mathrm{k})=\mathbb{Z}_{3}[\mathrm{x}] / \mathrm{I}$ is of the form $\mathrm{I}+\mathrm{h}$ where $\mathrm{h} \in$ $\mathbb{Z}_{3}[\mathrm{x}]$. But $\mathrm{h}=\mathrm{qk}+\mathrm{r}$ for some $\mathrm{q} \in \mathbb{Z}_{3}[\mathrm{x}]$ and where $\operatorname{deg} \mathrm{r}<\operatorname{deg} \mathrm{k}$, and hence $q k \in I$. Therefore $p$ must actually be of the form $I+r$, and thus there can be only as many distinct such p as there are distinct r .

The next theorem shows that this approach works in general.
Theorem 6.22 Let $k \in \mathbb{Z}_{p}[x]$ be a prime polynomial of degree $r$, and define an equivalence relation on $\mathbb{Z}_{\mathrm{p}}[\mathrm{x}]$ by $\mathrm{a} \approx \mathrm{b}$ if and only if $\mathrm{k} \mid(\mathrm{a}-\mathrm{b})$. Then the corresponding set of equivalence classes in $\mathbb{Z}_{p}[x]$ is a field of order $p^{r}$.

Proof First note that if $\mathrm{a}_{\mathrm{i}} \in \mathbb{Z}_{\mathrm{p}}[\mathrm{x}]$, then there are $\mathrm{p}^{\mathrm{r}}$ distinct polynomials of the form $a_{0}+a_{1} x+\cdots+a_{r-1} x^{r-1}$. This set of $p^{r}$ polynomials (which consists of all distinct polynomials of degrees $0,1,2, \ldots, r-1$ ) forms a complete set of representatives of the classes, and hence there are $\mathrm{p}^{\mathrm{r}}$ classes in all. Since these equivalence classes are just the cosets $\mathbb{Z}_{\mathrm{p}}[\mathrm{x}] /(\mathrm{k})$ where k is prime, it follows from Theorem 6.11 that $\mathbb{Z}_{\mathrm{p}}[\mathrm{x}] /(\mathrm{k})$ is a field.

One consequence of this theorem is that to construct a field of order $p^{r}$ we need only find a prime polynomial of degree $r$ in $\mathbb{Z}_{p}[x]$. While this is easy enough to do in most common cases, it is fairly hard to prove that there exists at least one prime polynomial for every choice of $p$ and $r$. We refer the interested reader to e.g., the very readable book by Biggs (1985).

## Exercises

1. Show that the operation $\oplus$ defined on $\mathbb{Z}_{\mathrm{n}}$ is well-defined. In other words, show that if $\left[a_{1}\right]=\left[a_{2}\right]$ and $\left[b_{1}\right]=\left[b_{2}\right]$ then $\left[a_{1}+b_{1}\right]=\left[a_{2}+b_{2}\right]$.
2. Finish the proof that $\mathbb{Z}_{\mathrm{n}}$ forms an additive group.
3. Prove Theorem 6.18.
4. Finish the proof that $\mathbb{Z}_{\mathrm{n}}$ forms a ring.
5. Finish the details in Example 6.10.
6. Finish the details in Example 6.11.
7. Prove that $\mathbb{Z}_{\mathrm{m}}$ is isomorphic to $\mathrm{C}_{\mathrm{m}}$.
8. If $m$ and $n$ are relatively prime positive integers, prove that $C_{m} \times C_{n}$ is isomorphic to $C_{m n}$. [Hint: Suppose $C_{m}=\langle x\rangle$ and $C_{n}=\langle y\rangle$. Let $z=$ $(x, y) \in C_{m} \times C_{n}$ have order $r$. Show that $r=m n$ and then conclude that $\mathrm{C}_{\mathrm{m}} \times \mathrm{C}_{\mathrm{n}}$ must be a cyclic group.]
9. Let G be a group, and suppose $\langle\mathrm{x}\rangle \subset \mathrm{G}$. If $\mathrm{o}(\mathrm{x}) \leq \mathrm{o}(\mathrm{G})$, show that $\langle\mathrm{x}\rangle$ is a subgroup of $G$ and that $o(x)=o(\langle x\rangle)$.
10. Fill in the details in the proof of Theorem 6.22.
11. For each of the following expressions in $\mathbb{Z}_{5}$, write the answer as [0], [1], [2], [3] or [4]:
(a) $[3] \oplus[4]$
(b) $[2] \oplus[-7]$
(c) $[17] \oplus[76]$
(d) $[3] \otimes[4]$
(e) $[2] \otimes[-7]$
(f) $[17] \otimes[76]$
(g) $[3] \otimes([2] \oplus[4])$
(h) $([3] \otimes[2]) \oplus([3] \otimes[4])$
12. Repeat the previous problem in $\mathbb{Z}_{6}$.
13. (a) Which elements of $\mathbb{Z}_{4}$ are zero divisors?
(b) Which elements of $\mathbb{Z}_{10}$ are zero divisors?
14. Show that $([2],[0])$ is a zero divisor in $\mathbb{Z}_{3} \times \mathbb{Z}_{3}$.
15. (a) Show that $1+x+x^{2} \in \mathbb{Z}_{2}[x]$ is prime over $\mathbb{Z}_{2}$.
(b) Show that $1+x^{2} \in \mathbb{Z}_{3}[x]$ is prime over $\mathbb{Z}_{3}$.
[Hint: Use the factor theorem.]
16. (a) Show that $1+x^{2}+x^{3} \in \mathbb{Z}_{2}[x]$ is prime over $\mathbb{Z}_{2}$, and use this to construct a field of order 8 .
(b) What is the order of its multiplicative group?
17. Prove that for every prime number p there exist fields of order $\mathrm{p}^{2}$ and $\mathrm{p}^{3}$.
18. For which of the following primes $p$ can we construct a field of order $p^{2}$ by using the polynomial $1+\mathrm{x}^{2}$ ?

$$
\mathrm{p}=3,5,7,11,13,19,23
$$

Describe the multiplicative group for the first two cases in which the field can be constructed.

## CHAPTER 7

## Linear Transformations and Polynomials

We now turn our attention to the problem of finding the basis in which a given linear transformation has the simplest possible representation. Such a representation is frequently called a canonical form. Although we would almost always like to find a basis in which the matrix representation of an operator is diagonal, this is in general impossible to do. Basically, in this chapter as well as in Chapters 8 and 10, we will try and find the general conditions that determine exactly what form it is possible for a representation to take.

In the present chapter, we focus our attention on eigenvalues and eigenvectors, which is probably the most important characterization of a linear operator that is available to us. We also treat the triangular form theorem from two distinct viewpoints. Our reason for this is that in this chapter we discuss both quotient spaces and nilpotent transformations, and the triangular form theorem is a good application of these ideas. However, since we also treat this theorem from an entirely different (and much simpler) point of view in the next chapter, the reader should feel free to skip Sections 7.10 to 7.12 if desired. (We also point out that Section 7.9 on quotient spaces is completely independent of the rest of this chapter, and may in fact be read immediately after Chapter 2.)

In Chapter 8 we give a complete discussion of canonical forms of matrices under similarity. All of the results that we prove in the present chapter for canonical forms of operators also follow from the development in Chapter 8. The reason for treating the "operator point of view" as well as the "matrix
point of view" is that the proof techniques and way of thinking can be quite different. The matrix point of view leads to more constructive and insightful proofs, while the operator point of view leads to techniques that are more likely to extend to infinite-dimensional analogs (although there is no complete extension to the infinite-dimensional version).

### 7.1 MINIMAL POLYNOMIALS

Let $\mathrm{f}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{n}} \mathrm{x}^{\mathrm{n}} \in \mathcal{F}[\mathrm{x}]$ be any polynomial in the indeterminate x . Then, given any linear operator $T \in L(V)$, we define the linear operator $f(T) \in$ $\mathrm{L}(\mathrm{V})$ as the polynomial in the operator T defined by substitution as

$$
\mathrm{f}(\mathrm{~T})=\mathrm{a}_{0} 1+\mathrm{a}_{1} \mathrm{~T}+\cdots+\mathrm{a}_{\mathrm{n}} \mathrm{~T}^{\mathrm{n}}
$$

where 1 is the identity transformation on V . Similarly, given any matrix $\mathrm{A} \in$ $\mathrm{M}_{\mathrm{m}}(\mathcal{F})$, we define the matrix polynomial $\mathrm{f}(\mathrm{A})$ by

$$
\mathrm{f}(\mathrm{~A})=\mathrm{a}_{0} \mathrm{I}+\mathrm{a}_{1} \mathrm{~A}+\cdots+\mathrm{a}_{\mathrm{n}} \mathrm{~A}^{\mathrm{n}}
$$

where now $I$ is the $m \times m$ identity matrix. If $T$ is such that $f(T)=0$, then we say that T is a root or zero of the polynomial f . This terminology also applies to a matrix A such that $\mathrm{f}(\mathrm{A})=0$.

If $A \in M_{m}(\mathcal{F})$ is the representation of $T \in L(V)$ relative to some (ordered) basis for $V$, then (in view of Theorem 5.13) we expect that $f(A)$ is the representation of $f(T)$. This is indeed the case.

Theorem 7.1 Let A be the matrix representation of an operator $T \in L(V)$. Then $f(A)$ is the representation of $f(T)$ for any polynomial $f \in \mathcal{F}[x]$.

## Proof This is Exercise 7.1.1.

The basic algebraic properties of polynomials in either operators or matrices are given by the following theorem.

Theorem 7.2 Suppose $T \in L(V)$ and let $f, g \in \mathcal{F}[x]$. Then
(a) $\mathrm{f}(\mathrm{T}) \mathrm{T}=\mathrm{Tf}(\mathrm{T})$.
(b) $(\mathrm{f} \pm \mathrm{g})(\mathrm{T})=\mathrm{f}(\mathrm{T}) \pm \mathrm{g}(\mathrm{T})$.
(c) $(\mathrm{fg})(\mathrm{T})=\mathrm{f}(\mathrm{T}) \mathrm{g}(\mathrm{T})$.
(d) $(\mathrm{cf})(\mathrm{T})=\operatorname{cf}(\mathrm{T})$ for any $\mathrm{c} \in \mathcal{F}$.

Furthermore, these same results also hold for any matrix representation $\mathrm{A} \in$ $\mathrm{M}_{\mathrm{n}}(\mathcal{F})$.

Proof In view of Theorem 6.1, we leave this as an easy exercise for the reader (see Exercise 7.1.2).

From this theorem and the fact that the ring of polynomials is commutative, it should be clear that any two polynomials in the operator T (or matrix A) also commute.

This discussion is easily generalized as follows. Let $\mathcal{A}$ be any algebra over $\mathcal{F}$ with unit element $e$, and let $\mathrm{f}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{n}} \mathrm{x}^{\mathrm{n}}$ be any polynomial in $\mathcal{F}[\mathrm{x}]$. Then for any $\alpha \in \mathcal{A}$ we define

$$
\mathrm{f}(\alpha)=\mathrm{a}_{0} \mathrm{e}+\mathrm{a}_{1} \alpha+\cdots+\mathrm{a}_{\mathrm{n}} \alpha^{\mathrm{n}} \in \mathcal{A}
$$

If $f(\alpha)=0$, then $\alpha$ is a root of $f$ and we say that $\alpha$ satisfies $f$. We now show that in fact every $\alpha \in \mathcal{A}$ satisfies some nontrivial polynomial in $\mathcal{F}[\mathrm{x}]$. Recall that by definition, an algebra $\mathcal{A}$ is automatically a vector space over $\mathcal{F}$.

Theorem 7.3 Let $\mathcal{A}$ be an algebra (with unit element e) of dimension $m$ over $\mathcal{F}$. Then every element $\alpha \in \mathcal{A}$ satisfies some nontrivial polynomial in $\mathcal{F}[x]$ of degree at most m .

Proof Since $\operatorname{dim} \mathcal{A}=\mathrm{m}$, it follows that for any $\alpha \in \mathcal{A}$, the $\mathrm{m}+1$ elements e, $\alpha, \alpha^{2}, \ldots, \alpha^{\mathrm{m}} \in \mathcal{A}$ must be linearly dependent (Theorem 2.6). This means there exist scalars $\mathrm{a}_{0}, \mathrm{a}_{1}, \ldots, \mathrm{a}_{\mathrm{m}} \in \mathcal{F}$ not all equal to zero such that

$$
\mathrm{a}_{0} \mathrm{e}+\mathrm{a}_{1} \alpha+\cdots+\mathrm{a}_{\mathrm{m}} \alpha^{\mathrm{m}}=0 .
$$

But then $\alpha$ satisfies the nontrivial polynomial

$$
\mathrm{f}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{m}} \mathrm{x}^{\mathrm{m}} \in \mathcal{F}[\mathrm{x}]
$$

which is of degree at most m .

Corollary Let V be a finite-dimensional vector space over $\mathcal{F}$, and suppose $\operatorname{dim} \mathrm{V}=\mathrm{n}$. Then any $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ satisfies some nontrivial polynomial $\mathrm{g} \in \mathcal{F}[\mathrm{x}]$ of degree at most $n^{2}$.

Proof By Theorem 5.8, $\mathrm{L}(\mathrm{V})$ is an algebra over $\mathcal{F}$, and by Theorem 5.4 we have $\operatorname{dim} \mathrm{L}(\mathrm{V})=\operatorname{dim} \mathrm{L}(\mathrm{V}, \mathrm{V})=\mathrm{n}^{2}$. The corollary now follows by direct application of Theorem 7.3.

While this corollary asserts that any $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ always satisfies some polynomial $\mathrm{g} \in \mathcal{F}[\mathrm{x}]$ of degree at most $\mathrm{n}^{2}$, we shall see a little later on that g can be chosen to have degree at most n (this is the famous Cayley-Hamilton theorem).

Now, for a given $T \in L(V)$, consider the set of all $f \in \mathcal{F}[x]$ with the property that $f(T)=0$. This set is not empty by virtue of the previous corollary. Hence (by well-ordering) we may choose a polynomial $\mathrm{p} \in \mathcal{F}[\mathrm{x}]$ of least degree with the property that $\mathrm{p}(\mathrm{T})=0$. Such a polynomial is called a minimal polynomial for T over $\mathcal{F}$. (We will present an alternative definition in terms of ideals in Section 7.4.)

Theorem 7.4 Let V be finite-dimensional and suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$. Then there exists a unique monic polynomial $\mathrm{m} \in \mathcal{F}[\mathrm{x}]$ such that $\mathrm{m}(\mathrm{T})=0$ and, in addition, if $\mathrm{q} \in \mathcal{F}[\mathrm{x}]$ is any other polynomial such that $\mathrm{q}(\mathrm{T})=0$, then $\mathrm{m} \mid \mathrm{q}$.

Proof The existence of a minimal polynomial $\mathrm{p} \in \mathcal{F}[\mathrm{x}]$ was shown in the previous paragraph, so all that remains is to prove the uniqueness of a particular (i.e., monic) minimal polynomial. Suppose

$$
\mathrm{p}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{n}} \mathrm{x}^{\mathrm{n}}
$$

so that $\operatorname{deg} \mathrm{p}=\mathrm{n}$. Multiplying p by $\mathrm{a}_{\mathrm{n}}{ }^{-1}$ we obtain a monic polynomial $\mathrm{m} \in$ $\mathcal{F}[\mathrm{x}]$ with the property that $\mathrm{m}(\mathrm{T})=0$. If $\mathrm{m}^{\prime}$ is another distinct monic polynomial of degree $n$ with the property that $m^{\prime}(T)=0$, then $m-m^{\prime}$ is a nonzero polynomial of degree less than $n$ (since the leading terms cancel) that is satisfied by T , thus contradicting the definition of n . This proves the existence of a unique monic minimal polynomial.

Now let q be another polynomial satisfied by T. Applying the division algorithm we have $\mathrm{q}=\mathrm{mg}+\mathrm{r}$ where either $\mathrm{r}=0$ or deg $\mathrm{r}<\operatorname{deg} \mathrm{m}$. Substituting $T$ into this equation and using the fact that $q(T)=0$ and $m(T)=0$ we find that $\mathrm{r}(\mathrm{T})=0$. But if $\mathrm{r} \neq 0$, then we would have a polynomial r with deg $\mathrm{r}<\operatorname{deg} \mathrm{m}$ such that $\mathrm{r}(\mathrm{T})=0$, contradicting the definition of m . We must therefore have $\mathrm{r}=0$ so that $\mathrm{q}=\mathrm{mg}$, and hence $\mathrm{m} \mid \mathrm{q}$.

From now on, all minimal polynomials will be assumed to be monic unless otherwise noted. Furthermore, in Section 7.3 we will show (as a consequence of the Cayley-Hamilton theorem) the existence of a minimal polynomial for matrices. It then follows as a consequence of Theorem 7.1 that
any $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and its corresponding matrix representation A both have the same minimal polynomial (since $m(T)=0$ if and only if $m(A)=0$ ).

Recall that $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is invertible if there exists an element $\mathrm{T}^{-1} \in \mathrm{~L}(\mathrm{~V})$ such that $\mathrm{TT}^{-1}=\mathrm{T}^{-1} \mathrm{~T}=1$ (where 1 is the identity element of $\mathrm{L}(\mathrm{V})$ ). It is interesting to note that for any invertible $\mathrm{T} \in \mathrm{L}(\mathrm{V})$, its inverse $\mathrm{T}^{-1}$ is actually a polynomial in T . This fact is essentially shown in the proof of the next theorem.

Theorem 7.5 Let V be finite-dimensional over $\mathcal{F}$. Then $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is invertible if and only if the constant term in the minimal polynomial for T is not equal to zero.

Proof Let the minimal polynomial for T over $\mathcal{F}$ be

$$
\mathrm{m}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{n}-1} \mathrm{x}^{\mathrm{n}-1}+\mathrm{x}^{\mathrm{n}}
$$

We first assume that $\mathrm{a}_{0} \neq 0$. Since m is the minimal polynomial for T , we have

$$
\mathrm{m}(\mathrm{~T})=\mathrm{a}_{0} 1+\mathrm{a}_{1} \mathrm{~T}+\cdots+\mathrm{a}_{\mathrm{n}-1} \mathrm{~T}^{\mathrm{n}-1}+\mathrm{T}^{\mathrm{n}}=0
$$

and hence multiplying by $\mathrm{a}_{0}{ }^{-1}$ and using Theorem 7.2 yields

$$
0=1+a_{0}^{-1} T\left(a_{1} 1+a_{2} T+\cdots+a_{n-1} T^{n-2}+T^{n-1}\right)
$$

or

$$
\begin{aligned}
1 & =T\left[-a_{0}^{-1}\left(a_{1} 1+a_{2} T+\cdots+a_{n-1} T^{n-2}+T^{n-1}\right)\right] \\
& =\left[-a_{0}^{-1}\left(a_{1} 1+a_{2} T+\cdots+a_{n-1} T^{n-2}+T^{n-1}\right)\right] T .
\end{aligned}
$$

This shows that $T^{-1}=-a_{0}{ }^{-1}\left(a_{1} 1+a_{2} T+\cdots+a_{n-1} T^{n-2}+T^{n-1}\right)$, and hence $T$ is invertible.

Now suppose $T$ is invertible, but that $\mathrm{a}_{0}=0$. Then we have

$$
\begin{aligned}
0 & =a_{1} T+a_{2} T^{2}+\cdots+a_{n-1} T^{n-1}+T^{n} \\
& =\left(a_{1} 1+a_{2} T+\cdots+a_{n-1} T^{n-2}+T^{n-1}\right) T .
\end{aligned}
$$

Multiplying from the right by $\mathrm{T}^{-1}$ yields

$$
0=\mathrm{a}_{1} 1+\mathrm{a}_{2} \mathrm{~T}+\cdots+\mathrm{a}_{\mathrm{n}-1} \mathrm{~T}^{\mathrm{n}-2}+\mathrm{T}^{\mathrm{n}-1}
$$

and hence $T$ satisfies the polynomial $p=a_{1}+a_{2} x+\cdots+a_{n-1} x^{n-2}+x^{n-1} \in$ $\mathcal{F}[\mathrm{x}]$. But $\operatorname{deg} \mathrm{p}=\mathrm{n}-1<\mathrm{n}$ which contradicts the definition of m as the minimal polynomial. Therefore we must have $\mathrm{a}_{0} \neq 0$.

Corollary Let V be finite-dimensional over $\mathcal{F}$, and assume that $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is invertible. Then $\mathrm{T}^{-1}$ is a polynomial in T over $\mathcal{F}$.

Proof If T is invertible, then $m(T)=a_{0} 1+a_{1} T+\cdots+a_{n-1} T^{n-1}+T^{n}=0$ with $\mathrm{a}_{0} \neq 0$. Multiplying by $\mathrm{a}_{0}{ }^{-1}$ then shows that

$$
\mathrm{T}^{-1}=-\mathrm{a}_{0}{ }^{-1}\left(\mathrm{a}_{1} 1+\mathrm{a}_{2} \mathrm{~T}+\cdots+\mathrm{a}_{\mathrm{n}-1} \mathrm{~T}^{\mathrm{n}-2}+\mathrm{T}^{\mathrm{n}-1}\right)
$$

While we have so far shown the existence of minimal polynomials, most readers would be hard-pressed at this point to actually find one given any particular linear operator. Fortunately, we will discover a fairly general method for finding the minimal polynomial of a matrix in Chapter 8 (see Theorem 8.10).

As we stated earlier, V will always denote a finite-dimensional vector space over a field $\mathcal{F}$. In addition, we will let $1 \in \mathrm{~L}(\mathrm{~V})$ denote the identity transformation on $V$ (i.e., the unit element of $L(V)$ ), and we let $I \in M_{n}(\mathcal{F})$ be the identity matrix.

## Exercises

1. Prove Theorem 7.1.
2. Prove Theorem 7.2.
3. Let V be finite-dimensional over $\mathcal{F}$, and suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is singular. Prove there exists a nonzero $\mathrm{S} \in \mathrm{L}(\mathrm{V})$ such that $\mathrm{ST}=\mathrm{TS}=0$.
4. Suppose V has a basis $\left\{\mathrm{e}_{1}, \mathrm{e}_{2}\right\}$. If $\mathrm{T} \in \mathrm{L}(\mathrm{V})$, then $\mathrm{Te}_{\mathrm{i}}=\sum_{\mathrm{j}} \mathrm{e}_{\mathrm{j}} \mathrm{a}_{\mathrm{ji}}$ for some $\left(\mathrm{a}_{\mathrm{ij}}\right) \in \mathrm{M}_{2}(\mathcal{F})$. Find a nonzero polynomial of degree 2 in $\mathcal{F}[\mathrm{x}]$ satisfied by T.
5. Repeat the previous problem, but let $\operatorname{dim} \mathrm{V}=3$ and find a polynomial of degree 3.
6. Let $\alpha \in \mathcal{F}$ be fixed, and define the linear transformation $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ by $\mathrm{T}(\mathrm{v})=\alpha \mathrm{v}$. This is called the scalar mapping belonging to $\alpha$. Show that T is the scalar mapping belonging to $\alpha$ if and only if the minimal polynomial for $T$ is $m(x)=x-\alpha$.

### 7.2 EIGENVALUES AND EIGENVECTORS

We now make a very important definition. If $\mathrm{T} \in \mathrm{L}(\mathrm{V})$, then an element $\lambda \in$ $\mathcal{F}$ is called an eigenvalue (also called a characteristic value or characteristic root) of T if there exists a nonzero vector $\mathrm{v} \in \mathrm{V}$ such that $\mathrm{T}(\mathrm{v})=\lambda \mathrm{v}$. In this case, we call the vector $v$ an eigenvector (or characteristic vector) belonging to the eigenvalue $\lambda$. Note that while an eigenvector is nonzero by definition, an eigenvalue may very well be zero.

Throughout the remainder of this chapter we will frequently leave off the parentheses around vector operands. In other words, we sometimes write Tv rather than $\mathrm{T}(\mathrm{v})$. This simply serves to keep our notation as uncluttered as possible.

An important criterion for the existence of an eigenvalue of T is the following.

Theorem 7.6 A linear operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ has eigenvalue $\lambda \in \mathcal{F}$ if and only if $\lambda 1-\mathrm{T}$ is singular.

Proof Suppose $\lambda 1-\mathrm{T}$ is singular. By definition, this means there exists a nonzero $\mathrm{v} \in \mathrm{V}$ such that $(\lambda 1-\mathrm{T}) \mathrm{v}=0$. But this is just $\mathrm{Tv}=\lambda \mathrm{v}$. The converse should be quite obvious.

Note, in particular, that 0 is an eigenvalue of T if and only if T is singular. In an exactly analogous manner, we say that an element $\lambda \in \mathcal{F}$ is an eigenvalue of a matrix $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ if there exists a nonzero (column) vector $\mathrm{v} \in \mathcal{F}^{\mathrm{n}}$ such that $A v=\lambda v$, and we call $v$ an eigenvector of $A$ belonging to the eigenvalue $\lambda$. Given a basis $\left\{\mathrm{e}_{\mathrm{i}}\right\}$ for $\mathcal{F}^{\mathrm{n}}$, we can write this matrix eigenvalue equation in terms of components as

$$
\sum_{j=1}^{n} a_{i j} v_{j}=\lambda v_{i}, \quad i=1, \ldots, n
$$

Now suppose $T \in L(V)$ and $v \in V$. If $\left\{e_{1}, \ldots, e_{n}\right\}$ is a basis for $V$, then $v$ $=\sum_{i} \mathrm{v}_{\mathrm{i}} \mathrm{e}_{\mathrm{i}}$ and hence

$$
\mathrm{T}(\mathrm{v})=\mathrm{T}\left(\sum_{\mathrm{i}} \mathrm{v}_{\mathrm{i}} \mathrm{e}_{\mathrm{i}}\right)=\sum_{\mathrm{i}} \mathrm{v}_{\mathrm{i}} \mathrm{~T}\left(\mathrm{e}_{\mathrm{i}}\right)=\sum_{\mathrm{i}, \mathrm{j}} \mathrm{e}_{\mathrm{j}} \mathrm{a}_{\mathrm{ji}} \mathrm{v}_{\mathrm{i}}
$$

where $\mathrm{A}=\left(\mathrm{a}_{\mathrm{ij}}\right)$ is the matrix representation of T relative to the basis $\left\{\mathrm{e}_{\mathrm{i}}\right\}$. Using this result, we see that if $T(v)=\lambda v$, then

$$
\Sigma_{\mathrm{i}, \mathrm{j}} \mathrm{e}_{\mathrm{j}} \mathrm{a}_{\mathrm{j} \mathrm{i}} \mathrm{v}_{\mathrm{i}}=\lambda \Sigma_{\mathrm{j}} \mathrm{v}_{\mathrm{j}} \mathrm{e}_{\mathrm{j}}
$$

and hence equating components shows that $\sum_{i} \mathrm{a}_{\mathrm{ji}} \mathrm{v}_{\mathrm{i}}=\lambda \mathrm{v}_{\mathrm{j}}$. We thus see that (as expected) the isomorphism between $\mathrm{L}(\mathrm{V})$ and $\mathrm{M}_{\mathrm{n}}(\mathcal{F})$ (see Theorem 5.13) shows that $\lambda$ is an eigenvalue of the linear transformation T if and only if $\lambda$ is also an eigenvalue of the corresponding matrix representation A. Using the notation of Chapter 5 , we can say that $\mathrm{T}(\mathrm{v})=\lambda \mathrm{v}$ if and only if $[\mathrm{T}]_{\mathrm{e}}[\mathrm{v}]_{\mathrm{e}}=$ $\lambda[v]_{e}$.

Example 7.1 Let us find all of the eigenvectors and associated eigenvalues of the matrix

$$
A=\left(\begin{array}{ll}
1 & 2 \\
3 & 2
\end{array}\right) .
$$

This means that we must find a vector $v=(x, y)$ such that $A v=\lambda v$. In matrix notation, this equation takes the form

$$
\left(\begin{array}{ll}
1 & 2 \\
3 & 2
\end{array}\right)\binom{x}{y}=\lambda\binom{x}{y}
$$

or

$$
\left(\begin{array}{cc}
1-\lambda & 2 \\
3 & 2-\lambda
\end{array}\right)\binom{x}{y}=0
$$

This is equivalent to the system

$$
\begin{align*}
& (1-\lambda) x+2 y=0 \\
& 3 x+(2-\lambda) y=0 \tag{*}
\end{align*}
$$

Since this homogeneous system of equations has a nontrivial solution if and only if the determinant of the coefficient matrix is nonzero (Corollary to Theorem 4.13), we must have

$$
\left|\begin{array}{cc}
1-\lambda & 2 \\
3 & 2-\lambda
\end{array}\right|=\lambda^{2}-3 \lambda-4=(\lambda-4)(\lambda+1)=0 .
$$

We thus see that the eigenvalues are $\lambda=4$ and $\lambda=-1$. (The roots of this polynomial are found either by inspection, or by applying the quadratic formula proved following Theorem 6.14.)

Substituting $\lambda=4$ into $(*)$ yields

$$
\begin{array}{r}
-3 x+2 y=0 \\
3 x-2 y=0
\end{array}
$$

or $\mathrm{y}=(3 / 2) \mathrm{x}$. This means that every eigenvector corresponding to the eigenvalue $\lambda=4$ has the form $v=(x, 3 x / 2)$. In other words, every multiple of the vector $\mathrm{v}=(2,3)$ is also an eigenvector with eigenvalue equal to 4 . If we substitute $\lambda=-1$ in $\left(^{*}\right)$, then we similarly find $y=-x$, and hence every multiple of the vector $\mathrm{v}=(1,-1)$ is an eigenvector with eigenvalue equal to -1 . //

We will generalize this approach in the next section. However, let us first take a brief look at some of the relationships between the eigenvalues of an operator and the roots of its minimal polynomial.

Theorem 7.7 Let $\lambda$ be an eigenvalue of $T \in L(V)$. Then $p(\lambda)$ is an eigenvalue of $\mathrm{p}(\mathrm{T})$ for any $\mathrm{p} \in \mathcal{F}[\mathrm{x}]$.

Proof If $\lambda$ is an eigenvalue of T , then there exists a nonzero $\mathrm{v} \in \mathrm{V}$ such that $\mathrm{Tv}=\lambda \mathrm{v}$. But then

$$
\mathrm{T}^{2}(\mathrm{v})=\mathrm{T}(\mathrm{Tv})=\mathrm{T}(\lambda \mathrm{v})=\lambda \mathrm{T}(\mathrm{v})=\lambda^{2} \mathrm{v}
$$

and by induction, it is clear that $T^{k}(v)=\lambda^{k} v$ for any $k=1,2, \ldots$. If we define $\mathrm{p}=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{m}} \mathrm{x}^{\mathrm{m}}$, then we have

$$
\mathrm{p}(\mathrm{~T})=\mathrm{a}_{0} 1+\mathrm{a}_{1} \mathrm{~T}+\cdots+\mathrm{a}_{\mathrm{m}} \mathrm{~T}^{\mathrm{m}}
$$

and hence

$$
\begin{aligned}
p(T) v & =a_{0} v+a_{1} \lambda v+\cdots+a_{m} \lambda^{m} v \\
& =\left(a_{0}+a_{1} \lambda+\cdots+a_{m} \lambda^{m}\right) v \\
& =p(\lambda) v .
\end{aligned}
$$

Corollary Let $\lambda$ be an eigenvalue of $\mathrm{T} \in \mathrm{L}(\mathrm{V})$. Then $\lambda$ is a root of the minimal polynomial for T .

Proof If $\mathrm{m}(\mathrm{x})$ is the minimal polynomial for T , then $\mathrm{m}(\mathrm{T})=0$ by definition. From Theorem 7.7, we have $m(\lambda) v=m(T) v=0$ where $v \neq 0$ is an eigenvector corresponding to $\lambda$. But then $\mathrm{m}(\lambda)=0$ (see Theorem 2.1(b)) so that $\lambda$ is a root of $m(x)$.

Since any eigenvalue of T is a root of the minimal polynomial for T , it is natural to ask about the number of eigenvalues that exist for a given $T \in L(V)$. Recall from the corollary to Theorem 6.4 that if $\mathrm{c} \in \mathcal{F}$ is a root of $\mathrm{f} \in \mathcal{F}[\mathrm{x}]$, then $(x-c) \mid f$. If $c$ is such that $(x-c)^{m} \mid f$ but no higher power of $x-c$ divides $f$, then we say that $c$ is a root of multiplicity $m$. (The context should make it
clear whether we mean the multiplicity $m$ or the minimal polynomial $m(x)$.) In counting the number of roots that a polynomial has, we shall always count a root of multiplicity m as m roots. A root of multiplicity 1 is frequently called a simple root.

If $\operatorname{dim} V=n$ then, since the minimal polynomial $m$ for $T \in L(V)$ is of degree at most $n^{2}$ (Corollary to Theorem 7.3), there can be at most $n^{2}$ roots of $m$ (Theorem 6.14). In particular, we see that any $T \in L(V)$ has a finite number of distinct eigenvalues. Moreover, if the field over which V is defined is algebraically closed, then T will in fact have at least as many (not necessarily distinct) eigenvalues as is the degree of its minimal polynomial.

Theorem 7.8 If $\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{r}}$ are eigenvectors belonging to the distinct eigenvalues $\lambda_{1}, \ldots, \lambda_{r}$ of $\mathrm{T} \in \mathrm{L}(\mathrm{V})$, then the set $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{r}}\right\}$ is linearly independent.

Proof If $\mathrm{r}=1$ there is nothing to prove, so we proceed by induction on r . In other words, we assume that the theorem is valid for sets of less than $r$ eigenvectors and show that in fact it is valid for sets of size $r$. Suppose that

$$
\begin{equation*}
a_{1} v_{1}+\cdots+a_{r} v_{r}=0 \tag{1}
\end{equation*}
$$

for some set of scalars $a_{i} \in \mathcal{F}$. We apply $T$ to this relation to obtain

$$
\begin{equation*}
a_{1} T\left(v_{1}\right)+\cdots+a_{r} T\left(v_{r}\right)=a_{1} \lambda_{1} v_{1}+\cdots+a_{r} \lambda_{r} v_{r}=0 . \tag{2}
\end{equation*}
$$

On the other hand, if we multiply (1) by $\lambda_{r}$ and subtract this from (2), we find (since $\mathrm{Tv}_{\mathrm{i}}=\lambda_{\mathrm{i}} \mathrm{v}_{\mathrm{i}}$ )

$$
\mathrm{a}_{1}\left(\lambda_{1}-\lambda_{\mathrm{r}}\right) \mathrm{v}_{1}+\cdots+\mathrm{a}_{\mathrm{r}-1}\left(\lambda_{\mathrm{r}-1}-\lambda_{\mathrm{r}}\right) \mathrm{v}_{\mathrm{r}-1}=0 .
$$

By our induction hypothesis, the set $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{r}-1}\right\}$ is linearly independent, and hence $\mathrm{a}_{\mathrm{i}}\left(\lambda_{\mathrm{i}}-\lambda_{\mathrm{r}}\right)=0$ for each $\mathrm{i}=1, \ldots, r-1$. But the $\lambda_{i}$ are distinct so that $\lambda_{i}-\lambda_{r} \neq 0$ for $\mathrm{i} \neq \mathrm{r}$, and therefore $\mathrm{a}_{\mathrm{i}}=0$ for each $\mathrm{i}=1, \ldots, \mathrm{r}-1$. Using this result in (1) shows that $a_{r}=0$ (since $v_{r} \neq 0$ by definition), and therefore $\mathrm{a}_{1}=\cdots=\mathrm{a}_{\mathrm{r}}=0$. This shows that the entire collection $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{r}}\right\}$ is independent.

Corollary 1 Suppose $T \in L(V)$ and $\operatorname{dim} V=n$. Then $T$ can have at most $n$ distinct eigenvalues in $\mathcal{F}$.

Proof Since $\operatorname{dim} \mathrm{V}=\mathrm{n}$, there can be at most n independent vectors in V. Since n distinct eigenvalues result in n independent eigenvectors, this corollary follows directly from Theorem 7.8.

Corollary 2 Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and $\operatorname{dim} \mathrm{V}=\mathrm{n}$. If T has n distinct eigenvalues, then there exists a basis for V which consists of eigenvectors of T .

Proof If T has n distinct eigenvalues, then (by Theorem 7.8) T must have n linearly independent eigenvectors. But n is the number of elements in any basis for V , and hence these n linearly independent eigenvectors in fact form a basis for V .

It should be remarked that one eigenvalue can belong to more than one linearly independent eigenvector. In fact, if $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and $\lambda$ is an eigenvalue of T , then the set $\mathrm{V}_{\lambda}$ of all eigenvectors of T belonging to $\lambda$ is a subspace of V called the eigenspace of $\lambda$. It is also easy to see that $\operatorname{V} \lambda=\operatorname{Ker}(\lambda 1-T)$ (see Exercise 7.2.1).

## Exercises

1. (a) If $T \in L(V)$ and $\lambda$ is an eigenvalue of $T$, show that the set $V_{\lambda}$ of all eigenvectors of T belonging to $\lambda$ is a T -invariant subspace of V (i.e., a subspace with the property that $\mathrm{T}(\mathrm{v}) \in \mathrm{V}_{\lambda}$ for all $\mathrm{v} \in \mathrm{V}_{\lambda}$ ).
(b) Show that $V_{\lambda}=\operatorname{Ker}(\lambda 1-T)$.
2. An operator $T \in L(V)$ with the property that $T^{n}=0$ for some $n \in \mathbb{Z}^{+}$is said to be nilpotent. Show that the only eigenvalue of a nilpotent operator is 0 .
3. If $\mathrm{S}, \mathrm{T} \in \mathrm{L}(\mathrm{V})$, show that ST and TS have the same eigenvalues. [Hint: First use Theorems 5.16 and 7.6 to show that 0 is an eigenvalue of ST if and only if 0 is an eigenvalue of TS. Now assume $\lambda \neq 0$, and let $\mathrm{ST}(\mathrm{v})=$ $\lambda \mathrm{v}$. Show that Tv is an eigenvector of TS.]
4. (a) Consider the rotation operator $R(\alpha) \in L\left(\mathbb{R}^{2}\right)$ defined in Example 1.2. Does $R(\alpha)$ have any eigenvectors? Explain.
(b) Repeat part (a) but now consider rotations in $\mathbb{R}^{3}$.
5. For each of the following matrices, find all eigenvalues and linearly independent eigenvectors:
(a) $\left(\begin{array}{ll}2 & 2 \\ 1 & 3\end{array}\right)$
(b) $\left(\begin{array}{ll}4 & 2 \\ 3 & 3\end{array}\right)$
(c) $\left(\begin{array}{rr}5 & -1 \\ 1 & 3\end{array}\right)$
6. Consider the spaces $\mathrm{D}[\mathrm{R}]$ and $\mathrm{F}[\mathrm{R}]$ defined in Exercise 2.1.6, and let d: $\mathrm{D}[\mathrm{R}] \rightarrow \mathrm{F}[\mathbb{R}]$ be the usual derivative operator.
(a) Show that the eigenfunctions (i.e., eigenvectors) of $d$ are of the form $\exp (\lambda x)$ where $\lambda$ is the corresponding eigenvalue.
(b) Suppose $\lambda_{1}, \ldots, \lambda_{r} \in \mathbb{R}$ are distinct. Show that the set

$$
S=\left\{\exp \left(\lambda_{1} x\right), \ldots, \exp \left(\lambda_{r} x\right)\right\}
$$

is linearly independent. [Hint: Consider the linear span of S.]
7. Suppose $T \in L(V)$ is invertible. Show that $\lambda$ is an eigenvalue of $T$ if and only if $\lambda \neq 0$ and $\lambda^{-1}$ is an eigenvalue of $\mathrm{T}^{-1}$.
8. Suppose $T \in L(V)$ and $\operatorname{dim} V=n$. If $T$ has $n$ linearly independent eigenvectors, what can you say about the matrix representation of T ?
9. Let V be a two-dimensional space over $\mathbb{R}$, and let $\left\{\mathrm{e}_{1}, \mathrm{e}_{2}\right\}$ be a basis for $V$. Find the eigenvalues and eigenvectors of the operator $T \in L(V)$ defined by:
(a) $\mathrm{Te}_{1}=\mathrm{e}_{1}+\mathrm{e}_{2}$
$\mathrm{Te}_{2}=\mathrm{e}_{1}-\mathrm{e}_{2}$.
(b) $\mathrm{Te}_{1}=5 \mathrm{e}_{1}+6 \mathrm{e}_{2}$ $T e_{2}=-7 \mathrm{e}_{2}$.
(c) $\mathrm{Te}_{1}=\mathrm{e}_{1}+2 \mathrm{e}_{2}$
$T e_{2}=3 \mathrm{e}_{1}+6 \mathrm{e}_{2}$.
10. Suppose $A \in M_{n}(\mathbb{C})$ and define $R_{i}=\sum_{j=1}^{n}\left|a_{i j}\right|$ and $P_{i}=R_{i}-\left|a_{i i}\right|$.
(a) Show that if $\mathrm{Ax}=0$ for some nonzero $\mathrm{x}=\left(\mathrm{x}_{1}, \ldots, \mathrm{x}_{\mathrm{n}}\right)$, then for any r such that $x_{r} \neq 0$ we have

$$
\left|a_{r r}\right|\left|x_{r}\right|=\left|\sum_{j \neq r} a_{r j} x_{j}\right| .
$$

(b) Show that part (a) implies that for some $r$ we have $\left|a_{r r}\right| \leq P_{r}$.
(c) Prove that if $\left|\mathrm{a}_{\mathrm{i}}\right|>\mathrm{P}_{\mathrm{i}}$ for all $\mathrm{i}=1, \ldots, \mathrm{n}$, then all eigenvalues of A are nonzero (or, equivalently, that $\operatorname{det} \mathrm{A} \neq 0$ ).
11. (a) Suppose $A \in M_{n}(\mathbb{C})$ and let $\lambda$ be an eigenvalue of $A$. Using the previous exercise prove Gershgorin's Theorem: $\left|\lambda-a_{r r}\right| \leq P_{r}$ for some $r$ with $1 \leq r \leq n$.
(b) Use this result to show that every eigenvalue $\lambda$ of the matrix

$$
A=\left(\begin{array}{rrrrr}
4 & 1 & 1 & 0 & 1 \\
1 & 3 & 1 & 0 & 0 \\
1 & 2 & 3 & 1 & 0 \\
1 & 1 & -1 & 4 & 0 \\
1 & 1 & 1 & 1 & 5
\end{array}\right)
$$

satisfies $1 \leq|\lambda| \leq 9$.

### 7.3 CHARACTERISTIC POLYNOMIALS

So far our discussion has dealt only theoretically with the existence of eigenvalues of an operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$. From a practical standpoint (as we saw in Example 7.1), it is much more convenient to deal with the matrix representation of an operator. Recall that the definition of an eigenvalue $\lambda \in \mathcal{F}$ and eigenvector $\mathrm{v}=\sum \mathrm{v}_{\mathrm{i}} \mathrm{e}_{\mathrm{i}}$ of a matrix $\mathrm{A}=\left(\mathrm{a}_{\mathrm{ij}}\right) \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ is given in terms of components by $\sum_{\mathrm{j}} \mathrm{a}_{\mathrm{ij}} \mathrm{v}_{\mathrm{j}}=\lambda \mathrm{v}_{\mathrm{i}}$ for each $\mathrm{i}=1, \ldots, \mathrm{n}$. This may be written in the form

$$
\sum_{j=1}^{n} a_{i j} v_{j}=\lambda \sum_{j=1}^{n} \delta_{i j} v_{j}
$$

or, alternatively, as

$$
\sum_{j=1}^{n}\left(\lambda \delta_{i j}-a_{i j}\right) v_{j}=0
$$

In matrix notation, this is

$$
(\lambda I-A) v=0 .
$$

By the corollary to Theorem 4.13, this set of homogeneous equations has a nontrivial solution if and only if $\operatorname{det}(\lambda I-A)=0$.

Another way to see this is to note that by Theorem $7.6, \lambda$ is an eigenvalue of the operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ if and only if $\lambda 1-\mathrm{T}$ is singular. But according to Theorem 5.16, this means that $\operatorname{det}(\lambda 1-\mathrm{T})=0$ (recall that the determinant of a linear transformation T is defined to be the determinant of any matrix representation of T ). In other words, $\lambda$ is an eigenvalue of T if and only if $\operatorname{det}(\lambda 1-\mathrm{T})=0$. This proves the following important result.

Theorem 7.9 Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and $\lambda \in \mathcal{F}$. Then $\lambda$ is an eigenvalue of T if and only if $\operatorname{det}(\lambda 1-T)=0$.

Let $[T]$ be a matrix representation of $T$. The matrix $x I-[T]$ is called the characteristic matrix of $[\mathrm{T}]$, and the expression $\operatorname{det}(\mathrm{x} 1-\mathrm{T})=0$ is called the characteristic (or secular) equation of $T$. The determinant $\operatorname{det}(x 1-T)$ is frequently denoted by $\Delta_{\mathrm{T}}(\mathrm{x})$. Writing out the determinant in a particular basis, we see that $\operatorname{det}(\mathrm{x} 1-\mathrm{T})$ is of the form

$$
\Delta_{T}(x)=\left|\begin{array}{cccc}
x-a_{11} & -a_{12} & \cdots & -a_{1 n} \\
-a_{21} & x-a_{22} & \cdots & -a_{2 n} \\
\vdots & \vdots & & \vdots \\
-a_{n 1} & -a_{n 2} & \cdots & x-a_{n n}
\end{array}\right|
$$

where $A=\left(a_{i j}\right)$ is the matrix representation of $T$ in the chosen basis. Since the expansion of a determinant contains exactly one element from each row and each column, we see that (see Exercise 7.3.1)

$$
\left.\left.\begin{array}{rl}
\operatorname{det}(x 1-T)=( & x
\end{array}\right) a_{11}\right)\left(x-a_{22}\right) \cdots\left(x-a_{n n}\right) .
$$

This polynomial is called the characteristic polynomial of T.
From the discussion following Theorem 5.18, we see that if $\mathrm{A}^{\prime}=\mathrm{P}^{-1} \mathrm{AP}$ is similar to A , then

$$
\operatorname{det}\left(x I-A^{\prime}\right)=\operatorname{det}\left(x I-P^{-1} A P\right)=\operatorname{det}\left[P^{-1}(x I-A) P\right]=\operatorname{det}(x I-A)
$$

(since $\operatorname{det} \mathrm{P}^{-1}=(\operatorname{det} \mathrm{P})^{-1}$ ). We thus see that similar matrices have the same characteristic polynomial (the converse of this statement is not true), and hence also the same eigenvalues. Therefore the eigenvalues (not eigenvectors) of an operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ do not depend on the basis chosen for V. Note also that according to Exercise 4.2.14, we may as well write $\operatorname{Tr} \mathrm{T}$ and $\operatorname{det} \mathrm{T}$ (rather than $\operatorname{Tr} \mathrm{A}$ and $\operatorname{det} \mathrm{A}$ ) since these are independent of the particular basis chosen. Using this terminology, we may rephrase Theorem 7.9 as follows.

Theorem 7.9́ A scalar $\lambda \in \mathcal{F}$ is an eigenvalue of $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ if and only if $\lambda$ is a root of the characteristic polynomial $\Delta_{\mathrm{T}}(\mathrm{x})$.

Since the characteristic polynomial is of degree $n$ in $x$, the corollary to Theorem 6.14 tells us that if we are in an algebraically closed field (such as
$\mathbb{C}$ ), then there must exist n roots. In this case, the characteristic polynomial may be factored into the form

$$
\operatorname{det}(x 1-T)=\left(x-\lambda_{1}\right)\left(x-\lambda_{2}\right) \cdots\left(x-\lambda_{n}\right)
$$

where the eigenvalues $\lambda_{\mathrm{i}}$ are not necessarily distinct. Expanding this expression we have

$$
\operatorname{det}(\mathrm{x} 1-\mathrm{T})=\mathrm{x}^{\mathrm{n}}-\left(\sum \lambda_{\mathrm{i}}\right) \mathrm{x}^{\mathrm{n}-1}+\cdots+(-1)^{\mathrm{n}} \lambda_{1} \lambda_{2} \cdots \lambda_{\mathrm{n}} .
$$

Comparing this with the above general expression for the characteristic polynomial, we see that

$$
\operatorname{Tr} T=\sum_{i=1}^{n} \lambda_{i}
$$

and

$$
\operatorname{det} T=\prod_{i=1}^{n} \lambda_{i}
$$

It should be remembered that this result only applies to an algebraically closed field (or to any other field $\mathcal{F}$ as long as all n roots of the characteristic polynomial lie in $\mathcal{F}$ ).

Example 7.2 Let us find the eigenvalues and eigenvectors of the matrix

$$
A=\left(\begin{array}{ll}
1 & 4 \\
2 & 3
\end{array}\right)
$$

The characteristic polynomial of A is given by

$$
\Delta_{A}(x)=\left|\begin{array}{cc}
x-1 & -4 \\
-2 & x-3
\end{array}\right|=x^{2}-4 x-5=(x-5)(x+1)
$$

and hence the eigenvalues of A are $\lambda=5,-1$. To find the eigenvectors corresponding to each eigenvalue, we must solve $A v=\lambda v$ or $(\lambda I-A) v=0$. Written out for $\lambda=5$ this is

$$
\left(\begin{array}{rr}
4 & -4 \\
-2 & 2
\end{array}\right)\binom{x}{y}=\binom{0}{0} .
$$

We must therefore solve the set of homogeneous linear equations

$$
\begin{array}{r}
4 x-4 y=0 \\
-2 x+2 y=0
\end{array}
$$

which is clearly equivalent to the single equation $x-y=0$, or $x=y$. This means that every eigenvector corresponding to the eigenvalue $\lambda=5$ is a multiple of the vector $(1,1)$, and thus the corresponding eigenspace is onedimensional.

For $\lambda=-1$ we have

$$
\left(\begin{array}{ll}
-2 & -4 \\
-2 & -4
\end{array}\right)\binom{x}{y}=\binom{0}{0}
$$

and the equation to be solved is (since both are the same) $-2 x-4 y=0$. The solution is thus $-\mathrm{x}=2 \mathrm{y}$ so that the eigenvector is a multiple of $(2,-1)$.

We now note that

$$
\operatorname{Tr} \mathrm{A}=1+3=4=\Sigma \lambda_{\mathrm{i}}
$$

and

$$
\operatorname{det} \mathrm{A}=3-8=-5=\Pi \lambda_{\mathrm{i}}
$$

It is also easy to see that these relationships hold for the matrix given in Example 7.1. //

It is worth remarking that the existence of eigenvalues of a given operator (or matrix) depends on the particular field we are working with. For example, the matrix

$$
A=\left(\begin{array}{rr}
0 & -1 \\
1 & 0
\end{array}\right)
$$

has characteristic polynomial $x^{2}+1$ which has no real roots, but does have the complex roots $\pm i$. In other words, A has no eigenvalues in $\mathbb{R}$, but does have the eigenvalues $\pm i$ in $\mathbb{C}$ (see Exercise 7.3.6).

Returning to the general case of an arbitrary field, it is clear that letting $\lambda=0$ in $\Delta_{\mathrm{T}}(\lambda)=\operatorname{det}(\lambda 1-\mathrm{T})=0$ shows that the constant term in the characteristic polynomial of A is given by $\Delta_{\mathrm{T}}(0)=(-1)^{\mathrm{n}}$ det A . In view of Theorems 4.6, 7.5 and the corollary to Theorem 7.7, we wonder if there are any relationships between the characteristic and minimal polynomials of T. There are indeed, and the first step in showing some of these relationships is to prove that every $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ satisfies some nontrivial polynomial. This is the essential content of our next result, the famous Cayley-Hamilton theorem.

Before we prove this theorem, we should point out that we will be dealing with matrices that have polynomial entries, rather than entries that are simply elements of the field $\mathcal{F}$. However, if we regard the polynomials as elements of
the field of quotients (see Section 6.5), then all of our previous results (for example, those dealing with determinants) remain valid. We shall elaborate on this problem in detail in the next chapter. Furthermore, the proof we are about to give is the standard one at this level. We shall find several other methods of proof throughout this text, including a remarkably simple one in the next chapter (see the discussion of matrices over the ring of polynomials).

Theorem 7.10 (Cayley-Hamilton Theorem) Every matrix $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ satisfies its characteristic polynomial.

Proof First recall from Theorem 4.11 that any matrix $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ obeys the relation

$$
\mathrm{A}(\operatorname{adj} \mathrm{~A})=(\operatorname{det} \mathrm{A}) \mathrm{I}_{\mathrm{n}}
$$

where $\operatorname{adj} \mathrm{A}$ is the matrix whose elements are the determinants of the minor matrices of A. In particular, the characteristic matrix xI - A obeys

$$
(x I-A) B(x)=\operatorname{det}(x I-A) I
$$

where we let

$$
B(x)=\operatorname{adj}(x I-A) .
$$

Thus the entries of the $n x n$ matrix $B(x)$ are polynomials in $x$ of degree $\leq n-$ 1. For example, if

$$
B(x)=\left(\begin{array}{ccc}
x^{2}+2 & x & 3 \\
-x+1 & 1 & 0 \\
0 & 4 & x^{2}
\end{array}\right)
$$

then

$$
B(x)=\left(\begin{array}{lll}
1 & 0 & 0 \\
0 & 0 & 0 \\
0 & 0 & 1
\end{array}\right) x^{2}+\left(\begin{array}{rrr}
0 & 1 & 0 \\
-1 & 0 & 0 \\
0 & 0 & 0
\end{array}\right) x+\left(\begin{array}{lll}
2 & 0 & 3 \\
1 & 1 & 0 \\
0 & 4 & 0
\end{array}\right)
$$

Hence in general, we may write $B(x)$ in the form

$$
\mathrm{B}(\mathrm{x})=\mathrm{B}_{0}+\mathrm{B}_{1} \mathrm{x}+\cdots+\mathrm{B}_{\mathrm{n}-1} \mathrm{x}^{\mathrm{n}-1}
$$

where each $\mathrm{B}_{\mathrm{i}} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$.
Now write

$$
\Delta_{\mathrm{A}}(\mathrm{x})=\operatorname{det}(\mathrm{xI}-\mathrm{A})=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{n}-1} \mathrm{x}^{\mathrm{n}-1}+\mathrm{x}^{\mathrm{n}}
$$

Then Theorem 4.11 becomes

$$
(x I-A)\left(B_{0}+B_{1} x+\cdots+B_{n-1} x^{n-1}\right)=\left(a_{0}+a_{1} x+\cdots+a_{n-1} x^{n-1}+x^{n}\right) I
$$

Equating powers of x in this equation yields

$$
\begin{aligned}
-A B_{0} & =a_{0} I \\
B_{0}-A B_{1} & =a_{1} I \\
B_{1}-A B_{2} & =a_{2} I \\
& \vdots \\
B_{n-2}-A B_{n-1} & =a_{n-1} I \\
B_{n-1} & =I
\end{aligned}
$$

We now multiply the first of these equations by $\mathrm{A}^{0}=\mathrm{I}$, the second by $\mathrm{A}^{1}=\mathrm{A}$, the third by $\mathrm{A}^{2}, \ldots$, the n th by $\mathrm{A}^{\mathrm{n}-1}$, and the last by $\mathrm{A}^{\mathrm{n}}$ to obtain

$$
\begin{aligned}
-A B_{0} & =a_{0} I \\
A B_{0}-A^{2} B_{1} & =a_{1} A \\
A^{2} B_{1}-A^{3} B_{2} & =a_{2} A^{2} \\
& \vdots \\
A^{n-1} B_{n-2}-A^{n} B_{n-1} & =a_{n-1} A^{n-1} \\
A^{n} B_{n-1} & =A^{n}
\end{aligned}
$$

Adding this last group of matrix equations, we see that the left side cancels out and we are left with

$$
0=\mathrm{a}_{0} \mathrm{I}+\mathrm{a}_{1} \mathrm{~A}+\mathrm{a}_{2} \mathrm{~A}^{2}+\cdots+\mathrm{a}_{\mathrm{n}-1} \mathrm{~A}^{\mathrm{n}-1}+\mathrm{A}^{\mathrm{n}}
$$

This shows that A satisfies its characteristic polynomial.
In view of this theorem, we see that there exists a nonempty set of nonzero polynomials $\mathrm{p}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$ such that $\mathrm{p}(\mathrm{A})=0$ for any $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$. (Alternatively, Theorem 7.3 and its corollary apply equally well to the algebra of matrices, although the degree is bounded by $\mathrm{n}^{2}$ rather than by n .) As we did for linear transformations, we may define the minimal polynomial for A as that polynomial $\mathrm{p}(\mathrm{x})$ of least degree for which $\mathrm{p}(\mathrm{A})=0$. We also noted following Theorem 7.4 that any $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and its corresponding representation $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ satisfy the same minimal polynomial. Theorem 7.4 thus applies equally well to matrices, and hence there exists a unique monic
minimal polynomial $\mathrm{m}(\mathrm{x})$ for A such that $\mathrm{m}(\mathrm{A})=0$. In addition, $\mathrm{m}(\mathrm{x})$ divides every other polynomial which has A as a zero. In particular, since A satisfies $\Delta_{\mathrm{A}}(\mathrm{x})$, we must have $\mathrm{m}(\mathrm{x}) \mid \Delta_{\mathrm{A}}(\mathrm{x})$.

Theorem 7.11 Suppose $A \in M_{n}(\mathcal{F})$ and $m(x)$ is the minimal polynomial for A. Then $m(x) \mid \Delta_{A}(x)$ and $\Delta_{A}(x) \mid[m(x)]^{n}$.

Proof That $\mathrm{m}(\mathrm{x}) \mid \Delta_{\mathrm{A}}(\mathrm{x})$ was proved in the previous paragraph. Let $\mathrm{m}(\mathrm{x})=$ $\mathrm{x}^{\mathrm{k}}+\mathrm{m}_{1} \mathrm{x}^{\mathrm{k}-1}+\cdots+\mathrm{m}_{\mathrm{k}-1} \mathrm{x}+\mathrm{m}_{\mathrm{k}}$. We define the matrices $\mathrm{B}_{\mathrm{i}} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ by

$$
\begin{aligned}
B_{0} & =I \\
B_{1} & =A+m_{1} I \\
B_{2} & =A^{2}+m_{1} A+m_{2} I \\
& \vdots \\
B_{k-1} & =A^{k-1}+m_{1} A^{k-2}+\cdots+m_{k-1} I
\end{aligned}
$$

where $I=I_{n}$. Working our way successively down this set of equations, we may rewrite them in the form

$$
\begin{aligned}
B_{0} & =I \\
B_{1}-A B_{0} & =m_{1} I \\
B_{2}-A B_{1} & =m_{2} I \\
& \vdots \\
B_{k-1}-A B_{k-2} & =m_{k-1} I
\end{aligned}
$$

From the previous expression for $\mathrm{B}_{\mathrm{k}-1}$, we multiply by -A and then add and subtract $\mathrm{m}_{\mathrm{k}} \mathrm{I}$ to obtain (using $\mathrm{m}(\mathrm{A})=0$ )

$$
\begin{aligned}
-A B_{k-1} & =m_{k} I-\left(A^{k}+m_{1} A^{k-1}+\cdots+m_{k-1} A+m_{k} I\right) \\
& =m_{k} I-m(A) \\
& =m_{k} I .
\end{aligned}
$$

Now define $B(x)=x^{k-1} B_{0}+x^{k-2} B_{1}+\cdots+x B_{k-2}+B_{k-1} \quad$ (which may be viewed either as a polynomial with matrix coefficients or as a matrix with polynomial entries). Then, using our previous results, we find that

$$
\begin{align*}
(x I-A) B(x)= & x B(x)-A B(x) \\
= & \left(x^{k} B_{0}+x^{k-1} B_{1}+\cdots+x^{2} B_{k-2}+x B_{k-1}\right) \\
& \quad-\left(x^{k-1} A B_{0}+x^{k-2} A B_{1}+\cdots+x A B_{k-2}+A B_{k-1}\right) \\
= & x^{k} B_{0}+x^{k-1}\left(B_{1}-A B_{0}\right)+x^{k-2}\left(B_{2}-A B_{1}\right) \\
& \quad+\cdots+x\left(B_{k-1}-A B_{k-2}\right)-A B_{k-1} \\
= & x^{k} I+m_{1} x^{k-1} I+m_{2} x^{k-2} I+\cdots+m_{k-1} x I+m_{k} I \\
= & m(x) I . \tag{*}
\end{align*}
$$

Since the determinant of a diagonal matrix is the product of its (diagonal) elements (see the corollary to Theorem 4.5), we see that

$$
\operatorname{det}[\mathrm{m}(\mathrm{x}) \mathrm{I}]=[\mathrm{m}(\mathrm{x})]^{\mathrm{n}}
$$

Therefore, taking the determinant of both sides of (*) and using Theorem 4.8 we find that

$$
[\operatorname{det}(\mathrm{xI}-\mathrm{A})][\operatorname{det} \mathrm{B}(\mathrm{x})]=\operatorname{det}[\mathrm{m}(\mathrm{x}) \mathrm{I}]=[\mathrm{m}(\mathrm{x})]^{\mathrm{n}}
$$

But det $B(x)$ is just some polynomial in $x$, so this equation shows that $[m(x)]^{n}$ is some multiple of $\Delta_{\mathrm{A}}(\mathrm{x})=\operatorname{det}(\mathrm{xI}-\mathrm{A})$. In other words, $\Delta_{\mathrm{A}}(\mathrm{x}) \mid[\mathrm{m}(\mathrm{x})]^{\mathrm{n}}$.

Theorem 7.12 The characteristic polynomial $\Delta_{\mathrm{A}}(\mathrm{x})$ and minimal polynomial $\mathrm{m}(\mathrm{x})$ of a matrix $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ have the same prime factors.

Proof Let $\mathrm{m}(\mathrm{x})$ have the prime factor $\mathrm{f}(\mathrm{x})$ so that $\mathrm{f}(\mathrm{x}) \mid \mathrm{m}(\mathrm{x})$. Since we showed in the above discussion that $\mathrm{m}(\mathrm{x}) \mid \Delta_{\mathrm{A}}(\mathrm{x})$, it follows that $\mathrm{f}(\mathrm{x}) \mid \Delta_{\mathrm{A}}(\mathrm{x})$ and hence $f(x)$ is a factor of $\Delta_{A}(x)$ also. Now suppose that $f(x) \mid \Delta_{A}(x)$. Theorem 7.11 shows that $\Delta_{A}(x) \mid[m(x)]^{n}$, and therefore $f(x) \mid[m(x)]^{n}$. However, since $f(x)$ is prime, Corollary $2^{\prime}$ of Theorem 6.5 tells us that $\mathrm{f}(\mathrm{x}) \mid \mathrm{m}(\mathrm{x})$.

It is important to realize that this theorem does not say that $\Delta_{A}(x)=m(x)$, but only that $\Delta_{\mathrm{A}}(\mathrm{x})$ and $\mathrm{m}(\mathrm{x})$ have the same prime factors. However, each factor can be of a different multiplicity in $\mathrm{m}(\mathrm{x})$ from what it is in $\Delta_{\mathrm{A}}(\mathrm{x})$. In particular, since $\mathrm{m}(\mathrm{x}) \mid \Delta_{\mathrm{A}}(\mathrm{x})$, the multiplicity of any factor in $\Delta_{\mathrm{A}}(\mathrm{x})$ must be greater than or equal to the multiplicity of the same factor in $m(x)$. Since a linear factor (i.e., a factor of the form $x-\lambda$ ) is prime, it then follows that $\Delta_{\mathrm{A}}(\mathrm{x})$ and $\mathrm{m}(\mathrm{x})$ have the same roots (although of different multiplicities).

Theorem 7.13 Suppose $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ and $\lambda \in \mathcal{F}$. Then $\lambda$ is an eigenvalue of A if and only if $\lambda$ is a root of the minimal polynomial for A .

Proof By Theorem 7.9, $\lambda$ is an eigenvalue of A if and only if $\Delta_{\mathrm{A}}(\lambda)=0$. But from the above remarks, $\lambda$ is a root of $\Delta_{\mathrm{A}}(\mathrm{x})$ if and only if $\lambda$ is a root of the minimal polynomial for A.

An alternative proof of Theorem 7.13 is to note that since $m(x) \mid \Delta_{A}(x)$, we may write $\Delta_{A}(x)=m(x) p(x)$ for some polynomial $p(x)$. If $\lambda$ is a root of $m(x)$, then $\Delta_{\mathrm{A}}(\lambda)=\mathrm{m}(\lambda) \mathrm{p}(\lambda)=0$ so that $\lambda$ is also a root of $\Delta_{\mathrm{A}}(\mathrm{x})$. In other words, $\lambda$ is an eigenvalue of $A$. The converse is just the corollary to Theorem 7.7. If we use this proof, then Theorem 7.12 is essentially just a corollary of Theorem 7.13.

Using Theorem 7.13, we can give another proof of Theorem 7.5 which also applies to the characteristic polynomial of any $\mathrm{T} \in \mathrm{L}(\mathrm{V})$. In particular, from Theorem 5.10 we see that T is invertible if and only if T is nonsingular if and only if 0 is not an eigenvalue of T (because this would mean that $\mathrm{Tv}=$ $0 \mathrm{v}=0$ for some $\mathrm{v} \neq 0$ ). But from Theorem 7.13, this is true if and only if 0 is not a root of the minimal polynomial for T . Writing the minimal polynomial as $m(x)=a_{0}+a_{1} x+\cdots+a_{k-1} x^{k-1}+x^{k}$, we then see that $a_{0}=m(0) \neq 0$ as claimed.

Example 7.3 Consider the matrix A given by

$$
\left(\begin{array}{llll}
2 & 1 & 0 & 0 \\
0 & 2 & 0 & 0 \\
0 & 0 & 2 & 0 \\
0 & 0 & 0 & 5
\end{array}\right)
$$

The characteristic polynomial is given by

$$
\Delta_{\mathrm{A}}(\mathrm{x})=\operatorname{det}(\mathrm{xI}-\mathrm{A})=(\mathrm{x}-2)^{3}(\mathrm{x}-5)
$$

and hence Theorem 7.12 tells us that both $x-2$ and $x-5$ must be factors of $\mathrm{m}(\mathrm{x})$. Furthermore, it follows from Theorems 7.4 and 7.10 that $\mathrm{m}(\mathrm{x}) \mid \Delta_{\mathrm{A}}(\mathrm{x})$, and thus the minimal polynomial must be either $\mathrm{m}_{1}(\mathrm{x}), \mathrm{m}_{2}(\mathrm{x})$ or $\mathrm{m}_{3}(\mathrm{x})$ where

$$
m_{j}(x)=(x-2)^{j}(x-5)
$$

From the Cayley-Hamilton theorem we know that $m_{3}(A)=\Delta_{A}(A)=0$, and it is easy to show that $\mathrm{m}_{2}(\mathrm{~A})=0$ also while $\mathrm{m}_{1}(\mathrm{~A}) \neq 0$. Therefore the minimal polynomial for A must be $\mathrm{m}_{2}(\mathrm{x})$. //

We now turn our attention to one of the most important aspects of the existence of eigenvalues. Suppose that $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ with $\operatorname{dim} \mathrm{V}=\mathrm{n}$. If V has a basis $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{n}}\right\}$ that consists entirely of eigenvectors of T , then the matrix representation of T in this basis is defined by

$$
T\left(v_{i}\right)=\sum_{j=1}^{n} v_{j} a_{j i}=\lambda_{i} v_{i}=\sum_{j=1}^{n} \delta_{j i} \lambda_{j} v_{j}
$$

and therefore $\mathrm{a}_{\mathrm{ji}}=\delta_{\mathrm{ji}} \lambda_{\mathrm{j}}$. In other words, T is represented by a diagonal matrix in a basis of eigenvectors. Conversely, if T is represented by a diagonal matrix $\mathrm{a}_{\mathrm{ji}}=\delta_{\mathrm{j} i} \lambda_{\mathrm{j}}$ relative to some basis $\left\{\mathrm{v}_{\mathrm{i}}\right\}$, then reversing the argument shows that each $v_{i}$ is an eigenvector of T. This proves the following theorem.

Theorem 7.14 A linear operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ can be represented by a diagonal matrix if and only if V has a basis consisting of eigenvectors of T . If this is the case, then the diagonal elements of the matrix representation are precisely the eigenvalues of T . (Note however, that the eigenvalues need not necessarily be distinct.)

If $T \in L(V)$ is represented in some basis $\left\{e_{i}\right\}$ by a matrix $A$, and in the basis of eigenvectors $v_{i}$ by a diagonal matrix D , then Theorem 5.18 tells us that A and D must be similar matrices. This proves the following version of Theorem 7.14, which we state as a corollary.

Corollary 1 A matrix $A \in M_{n}(\mathcal{F})$ is similar to a diagonal matrix $D$ if and only if A has n linearly independent eigenvectors.

Corollary 2 A linear operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ can be represented by a diagonal matrix if T has $\mathrm{n}=\operatorname{dim} \mathrm{V}$ distinct eigenvalues.

Proof This follows from Corollary 2 of Theorem 7.8.
Note that the existence of $n=\operatorname{dim} V$ distinct eigenvalues of $T \in L(V)$ is a sufficient but not necessary condition for T to have a diagonal representation. For example, the identity operator has the usual diagonal representation, but its only eigenvalues are $\lambda=1$. In general, if any eigenvalue has multiplicity greater than 1, then there will be fewer distinct eigenvalues than the dimension of V. However, in this case we may be able to choose an appropriate linear combination of eigenvectors in each eigenspace so that the matrix of T will still be diagonal. We shall have more to say about this in Section 7.7.

We say that a matrix A is diagonalizable if it is similar to a diagonal matrix $D$. If $P$ is a nonsingular matrix such that $D=P^{-1} A P$, then we say that $P$
diagonalizes $A$. It should be noted that if $\lambda$ is an eigenvalue of a matrix $A$ with eigenvector v (i.e., $\mathrm{Av}=\lambda \mathrm{v}$ ), then for any nonsingular matrix P we have

$$
\left(\mathrm{P}^{-1} \mathrm{AP}\right)\left(\mathrm{P}^{-1} \mathrm{v}\right)=\mathrm{P}^{-1} \mathrm{Av}=\mathrm{P}^{-1} \lambda \mathrm{v}=\lambda\left(\mathrm{P}^{-1} \mathrm{v}\right) .
$$

In other words, $\mathrm{P}^{-1} \mathrm{v}$ is an eigenvector of $\mathrm{P}^{-1} \mathrm{AP}$. Similarly, we say that $\mathrm{T} \in$ $\mathrm{L}(\mathrm{V})$ is diagonalizable if there exists a basis for V that consists entirely of eigenvectors of $T$.

While all of this sounds well and good, the reader might wonder exactly how the transition matrix P is to be constructed. Actually, the method has already been given in Section 5.4. If $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and A is the matrix representation of T in a basis $\left\{\mathrm{e}_{\mathrm{i}}\right\}$, then P is defined to be the transformation that takes the basis $\left\{\mathrm{e}_{\mathrm{i}}\right\}$ into the basis $\left\{\mathrm{v}_{\mathrm{i}}\right\}$ of eigenvectors. In other words, $\mathrm{v}_{\mathrm{i}}=\mathrm{Pe}_{\mathrm{i}}=\sum_{\mathrm{j}} \mathrm{e}_{\mathrm{j}} \mathrm{p}_{\mathrm{ji}}$. This means that the i th column of $\left(\mathrm{p}_{\mathrm{ij}}\right)$ is just the i th eigenvector of A . The fact that P must be nonsingular coincides with the requirement that $T$ (or A) have $n$ linearly independent eigenvectors $\mathrm{v}_{\mathrm{i}}$.

Example 7.4 Referring to Example 7.2, we found the eigenvectors $\mathrm{v}_{1}=(1,1)$ and $\mathrm{v}_{2}=(2,-1)$ belonging to the matrix

$$
A=\left(\begin{array}{ll}
1 & 4 \\
2 & 3
\end{array}\right) .
$$

Then P and $\mathrm{P}^{-1}$ are given by

$$
P=\left(\begin{array}{rr}
1 & 2 \\
1 & -1
\end{array}\right)
$$

and

$$
P^{-1}=\frac{\operatorname{adj} P}{\operatorname{det} P}=\left(\begin{array}{rr}
1 / 3 & 2 / 3 \\
1 / 3 & -1 / 3
\end{array}\right)
$$

and therefore

$$
D=P^{-1} A P=\left(\begin{array}{rr}
1 / 3 & 2 / 3 \\
1 / 3 & -1 / 3
\end{array}\right)\left(\begin{array}{rr}
1 & 4 \\
2 & 3
\end{array}\right)\left(\begin{array}{rr}
1 & 2 \\
1 & -1
\end{array}\right)=\left(\begin{array}{rr}
5 & 0 \\
0 & -1
\end{array}\right) .
$$

We see that D is a diagonal matrix, and that the diagonal elements are just the eigenvalues of A. Note also that

$$
D\left(P^{-1} v_{1}\right)=\left(\begin{array}{rr}
5 & 0 \\
0 & -1
\end{array}\right)\left(\begin{array}{rr}
1 / 3 & 2 / 3 \\
1 / 3 & -1 / 3
\end{array}\right)\binom{1}{1}
$$

$$
\begin{aligned}
& =\left(\begin{array}{rr}
5 & 0 \\
0 & -1
\end{array}\right)\binom{1}{0}=5\binom{1}{0} \\
& =\lambda_{1}\left(P^{-1} v_{1}\right)
\end{aligned}
$$

with a similar result holding for $\mathrm{P}^{-1} \mathrm{v}_{2}$. //

Example 7.5 Let us show that the matrix

$$
A=\left(\begin{array}{ll}
1 & 2 \\
0 & 1
\end{array}\right)
$$

is not diagonalizable. The characteristic equation is $(x-1)^{2}=0$, and hence there are two identical roots $\lambda=1$. If there existed an eigenvector $v=(x, y)$, it would have to satisfy the equation $(\lambda I-A) v=0$ or

$$
\left(\begin{array}{rr}
0 & -2 \\
0 & 0
\end{array}\right)\binom{x}{y}=\binom{0}{0} .
$$

Since this yields $-2 y=0$, the eigenvectors must be of the form ( $x, 0$ ), and hence it is impossible to find two linearly independent such eigenvectors.

Note that the minimal polynomial for A is either $\mathrm{x}-1$ or $(\mathrm{x}-1)^{2}$. But since $A-1 I \neq 0, m(x)$ must be $(x-1)^{2}$. We will see later (Theorem 7.24) that a matrix is diagonalizable if and only if its minimal polynomial is a product of distinct linear factors. //

## Exercises

1. Suppose $T \in L(V)$ has matrix representation $A=\left(a_{i j}\right)$, and $\operatorname{dim} V=n$. Prove that

$$
\begin{aligned}
\operatorname{det}(\mathrm{x} 1 & -\mathrm{T}) \\
& =\mathrm{x}^{\mathrm{n}}-(\operatorname{Tr} \mathrm{A}) \mathrm{x}^{\mathrm{n}-1}+\text { terms of lower degree in } \mathrm{x}+(-1)^{\mathrm{n}} \operatorname{det} \mathrm{~A} .
\end{aligned}
$$

[Hint: Use the definition of determinant.]
2. Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is diagonalizable. Show that the minimal polynomial $\mathrm{m}(\mathrm{x})$ for T must consist of distinct linear factors. [Hint: Let T have distinct eigenvalues $\lambda_{1}, \ldots, \lambda_{\mathrm{r}}$ and consider the polynomial

$$
f(x)=\left(x-\lambda_{1}\right) \cdots\left(x-\lambda_{r}\right) .
$$

Show that $m(x)=f(x)$.]
3. Prove by direct substitution that $\Delta_{\mathrm{A}}(\mathrm{A})=0$ if $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ is diagonal.
4. Find, in the form $a_{0}+a_{1} x+a_{2} x^{2}+a_{3} x^{3}$, the characteristic polynomial of

$$
A=\left(\begin{array}{rrr}
1 & 2 & -1 \\
0 & 3 & 1 \\
2 & 0 & -2
\end{array}\right)
$$

Show by direct substitution that A satisfies its characteristic polynomial.
5. If $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and $\Delta_{\mathrm{T}}(\mathrm{x})$ is a product of distinct linear factors, prove that T is diagonalizable.
6. Consider the following matrices:

$$
A=\left(\begin{array}{rr}
2 & -1 \\
1 & 4
\end{array}\right) \quad B=\left(\begin{array}{rr}
3 & -1 \\
13 & -3
\end{array}\right)
$$

(a) Find all eigenvalues and linearly independent eigenvectors over $\mathbb{R}$.
(b) Find all eigenvalues and linearly independent eigenvectors over $\mathbb{C}$.
7. For each of the following matrices, find all eigenvalues, a basis for each eigenspace, and determine whether or not the matrix is diagonalizable:
(a) $\left(\begin{array}{lll}1 & -3 & 3 \\ 3 & -5 & 3 \\ 6 & -6 & 4\end{array}\right)$
(b) $\left(\begin{array}{lll}-3 & 1 & -1 \\ -7 & 5 & -1 \\ -6 & 6 & -2\end{array}\right)$
8. Consider the operator $T \in L\left(\mathbb{R}^{3}\right)$ defined by

$$
\mathrm{T}(\mathrm{x}, \mathrm{y}, \mathrm{z})=(2 \mathrm{x}+\mathrm{y}, \mathrm{y}-\mathrm{z}, 2 \mathrm{y}+4 \mathrm{z})
$$

Find all eigenvalues and a basis for each eigenspace.
9. Let $A=\left(\mathrm{a}_{\mathrm{ij}}\right)$ be a triangular matrix, and assume that all of the diagonal entries of A are distinct. Is A diagonalizable? Explain.
10. Suppose $A \in M_{3}(\mathbb{R})$. Show that $A$ can not be a zero of the polynomial $f=$ $x^{2}+1$.
11. If $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$, show that A and $\mathrm{A}^{\mathrm{T}}$ have the same eigenvalues.
12. Suppose $A$ is a block triangular matrix with square matrices $A_{i i}$ on the diagonal. Show that the characteristic polynomial of A is the product of the characteristic polynomials of the $\mathrm{A}_{\mathrm{ii}}$.
13. Find the minimal polynomial of

$$
A=\left(\begin{array}{rrrr}
2 & 1 & 0 & 0 \\
0 & 2 & 0 & 0 \\
0 & 0 & 1 & 1 \\
0 & 0 & -2 & 4
\end{array}\right)
$$

14. For each of the following matrices A , find a nonsingular matrix P (if it exists) such that $\mathrm{P}^{-1} \mathrm{AP}$ is diagonal:
(a) $A=\left(\begin{array}{lll}3 & 1 & 1 \\ 2 & 4 & 2 \\ 1 & 1 & 3\end{array}\right)$
(b) $\left(\begin{array}{rrr}1 & 2 & 2 \\ 1 & 2 & -1 \\ -1 & 1 & 4\end{array}\right)$
(c) $\left(\begin{array}{lll}1 & 1 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right)$
15. Consider the following real matrix:

$$
A=\left(\begin{array}{ll}
a & b \\
c & d
\end{array}\right)
$$

Find necessary and sufficient conditions on $\mathrm{a}, \mathrm{b}, \mathrm{c}$ and d so that A is diagonalizable.
16. Let $A$ be an idempotent matrix (i.e., $A^{2}=A$ ) of rank $r$. Show that $A$ is similar to the matrix

$$
B=\left(\begin{array}{cc}
I_{r} & 0 \\
0 & 0
\end{array}\right) \text {. }
$$

17. Let $V$ be the space of all real polynomials $f \in \mathbb{R}[x]$ of degree at most 2 , and define $T \in L(V)$ by $T f=f+f^{\prime}+x f^{\prime}$ where $f^{\prime}$ denotes the usual derivative with respect to x .
(a) Write down the most obvious basis $\left\{\mathrm{e}_{1}, \mathrm{e}_{2}, \mathrm{e}_{3}\right\}$ for V you can think of, and then write down $[\mathrm{T}]_{\mathrm{e}}$.
(b) Find all eigenvalues of T , and then find a nonsingular matrix P such that $\mathrm{P}^{-1}[\mathrm{~T}]_{\mathrm{e}} \mathrm{P}$ is diagonal.
18. Prove that any real symmetric $2 \times 2$ matrix is diagonalizable.
19. (a) Let $\mathrm{C} \in \mathrm{M}_{2}(\mathbb{C})$ be such that $\mathrm{C}^{2}=0$. Prove that either $\mathrm{C}=0$, or else C is similar over $\mathbb{C}$ to the matrix

$$
\left(\begin{array}{ll}
0 & 0 \\
1 & 0
\end{array}\right) .
$$

(b) Prove that $\mathrm{A} \in \mathrm{M}_{2}(\mathbb{C})$ is similar over $\mathbb{C}$ to one of the following two types of matrices:

$$
\left(\begin{array}{ll}
a & 0 \\
0 & b
\end{array}\right) \quad \text { or } \quad\left(\begin{array}{ll}
\mathrm{a} & 0 \\
1 & \mathrm{a}
\end{array}\right)
$$

20. Find a matrix $A \in M_{3}(\mathbb{R})$ that has the eigenvalues 3,2 and 2 with corresponding eigenvectors $(2,1,1),(1,0,1)$ and $(0,0,4)$.
21. Is it possible for the matrix

$$
A=\left(\begin{array}{rrrr}
3 & 1 & 0 & 0 \\
-1 & 0 & 0 & 0 \\
0 & 0 & 2 & 1 \\
0 & 0 & 3 & 7
\end{array}\right)
$$

to have the eigenvalues $-1,2,3$ and 5 ?

### 7.4 ANNIHILATORS

The purpose of this section is to repeat our description of minimal polynomials using the formalism of ideals developed in the previous chapter. Our reasons for this are twofold. First, we will gain additional insight into the meaning and action of minimal polynomials. And second, these results will be of use in the next chapter when we discuss cyclic subspaces.

If V is a vector space of dimension n over $\mathcal{F}$, then for any $\mathrm{v} \in \mathrm{V}$ and any $\mathrm{T} \in \mathrm{L}(\mathrm{V})$, the $\mathrm{n}+1$ vectors $\mathrm{v}, \mathrm{T}(\mathrm{v}), \mathrm{T}^{2}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{n}}(\mathrm{v})$ must be linearly
dependent. This means there exist scalars $\mathrm{a}_{0}, \ldots, \mathrm{a}_{\mathrm{n}} \in \mathcal{F}$ not all equal to zero, such that

$$
\sum_{i=1}^{n} a_{i} T^{i}(v)=0
$$

If we define the polynomial $f(x)=a_{0}+a_{1} x+\cdots+a_{n} x^{n} \in \mathcal{F}[x]$, we see that our relation may be written as $f(T)(v)=0$. In other words, given any (fixed) $\mathrm{v} \in \mathrm{V}$ and $\mathrm{T} \in \mathrm{L}(\mathrm{V})$, then there exists a polynomial f of degree $\leq \mathrm{n}$ such that $f(T)(v)=0$.

Now, for any fixed $v \in V$ and $T \in L(V)$, we define the set $\mathrm{N}_{\mathrm{T}}(\mathrm{v})$ by

$$
\mathrm{N}_{\mathrm{T}}(\mathrm{v})=\{\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]: \mathrm{f}(\mathrm{~T})(\mathrm{v})=0\}
$$

This set is called the $T$-annihilator of $v$. If $f_{1}, f_{2} \in N_{T}(v)$, then we have

$$
\left[\mathrm{f}_{1}(\mathrm{~T}) \pm \mathrm{f}_{2}(\mathrm{~T})\right](\mathrm{v})=\mathrm{f}_{1}(\mathrm{~T})(\mathrm{v}) \pm \mathrm{f}_{2}(\mathrm{~T})(\mathrm{v})=0
$$

and for any $\mathrm{g}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$, we also have

$$
\left[\mathrm{g}(\mathrm{~T}) \mathrm{f}_{\mathrm{l}}(\mathrm{~T})\right](\mathrm{v})=\mathrm{g}(\mathrm{~T})\left[\mathrm{f}_{\mathrm{l}}(\mathrm{~T})(\mathrm{v})\right]=0 .
$$

This shows that $\mathrm{N}_{\mathrm{T}}(\mathrm{v})$ is actually an ideal of $\mathcal{F}[\mathrm{x}]$. Moreover, from Theorem 6.8 and its corollary, we see that $\mathrm{N}_{\mathrm{T}}(\mathrm{v})$ is a principal ideal, and hence has a unique monic generator, which we denote by $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$. By definition, this means that $\mathrm{N}_{\mathrm{T}}(\mathrm{v})=\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mathcal{F}[\mathrm{x}]$, and since we showed above that $\mathrm{N}_{\mathrm{T}}(\mathrm{v})$ contains at least one polynomial of degree less than or equal to $n$, it follows from Theorem 6.2(b) that deg $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \leq \mathrm{n}$. We call $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ the minimal polynomial of the vector $\mathbf{v}$ corresponding to the given transformation T. (Many authors also refer to $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ as the $\mathbf{T}$-annihilator of v , or the order of $v$.) It is thus the unique monic polynomial of least degree such that $m(T)(v)=0$.

Theorem 7.15 Suppose $T \in L(V)$, and let $v \in V$ have minimal polynomial $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$. Assume $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ is reducible so that $\mathrm{m}_{\mathrm{v}}(\mathrm{x})=\mathrm{m}_{1}(\mathrm{x}) \mathrm{m}_{2}(\mathrm{x})$ where $\mathrm{m}_{1}(\mathrm{x})$ and $m_{2}(x)$ are both monic polynomials of degree $\geq 1$. Then the vector $w=$ $\mathrm{m}_{1}(\mathrm{~T})(\mathrm{v}) \in \mathrm{V}$ has minimal polynomial $\mathrm{m}_{2}(\mathrm{x})$. In other words, every factor of the minimal polynomial of a vector is also the minimal polynomial of some other vector.

Proof First note that

$$
\mathrm{m}_{2}(\mathrm{~T})(\mathrm{w})=\mathrm{m}_{2}(\mathrm{~T})\left[\mathrm{m}_{1}(\mathrm{~T})(\mathrm{v})\right]=\mathrm{m}_{\mathrm{v}}(\mathrm{~T})(\mathrm{v})=0
$$

and thus $\mathrm{m}_{2}(\mathrm{x}) \in \mathrm{N}_{\mathrm{T}}(\mathrm{w})$. Then for any nonzero polynomial $\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$ with $\mathrm{f}(\mathrm{T})(\mathrm{w})=0$, we have $\mathrm{f}(\mathrm{T})\left[\mathrm{m}_{1}(\mathrm{~T})(\mathrm{v})\right]=0$, and hence $\mathrm{f}(\mathrm{x}) \mathrm{m}_{1}(\mathrm{x}) \in \mathrm{N}_{\mathrm{T}}(\mathrm{v})=$ $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mathcal{F}[\mathrm{x}]$. Using this result along with Theorem 6.2(b), we see that

$$
\operatorname{deg} m_{1}+\operatorname{deg} f=\operatorname{deg} m_{1} f \geq \operatorname{deg} m_{v}=\operatorname{deg} m_{1}+\operatorname{deg} m_{2}
$$

and therefore deg $\mathrm{m}_{2} \leq \operatorname{deg} \mathrm{f}$. This shows that $\mathrm{m}_{2}(\mathrm{x})$ is the monic polynomial of least degree such that $\mathrm{m}_{2}(\mathrm{~T})(\mathrm{w})=0$.

Theorem 7.16 Suppose $T \in L(V)$, and let $\mathrm{m}_{\mathrm{u}}(\mathrm{x})$ and $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ be the minimal polynomials of $u \in V$ and $v \in V$ respectively. Then the least common multiple $\mathrm{m}(\mathrm{x})$ of $\mathrm{m}_{\mathrm{u}}(\mathrm{x})$ and $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ is the minimal polynomial of some vector in V.

Proof We first assume that $\mathrm{m}_{\mathrm{u}}(\mathrm{x})$ and $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ are relatively prime so that their greatest common divisor is 1 . By Theorem 6.9 we then have $\mathrm{m}_{\mathrm{u}}(\mathrm{x}) \mathrm{m}_{\mathrm{v}}(\mathrm{x})=$ $m(x)$. From Theorem 6.5 we may write $m_{u}(x) k(x)+m_{v}(x) h(x)=1$ for some polynomials $\mathrm{h}, \mathrm{k} \in \mathcal{F}[\mathrm{x}]$, and therefore

$$
\mathrm{m}_{\mathrm{u}}(\mathrm{~T}) \mathrm{k}(\mathrm{~T})+\mathrm{m}_{\mathrm{v}}(\mathrm{~T}) \mathrm{h}(\mathrm{~T})=1
$$

Now define the vector $w \in V$ by

$$
\mathrm{w}=\mathrm{h}(\mathrm{~T})(\mathrm{u})+\mathrm{k}(\mathrm{~T})(\mathrm{v}) .
$$

Then, using $\mathrm{m}_{\mathrm{u}}(\mathrm{T})(\mathrm{u})=0=\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{v})$ we have

$$
\begin{aligned}
m_{u}(T)(w) & =m_{u}(T) h(T)(u)+m_{u}(T) k(T)(v) \\
& =m_{u}(T) k(T)(v) \\
& =\left[1-m_{v}(T) h(T)\right](v) \\
& =v
\end{aligned}
$$

and similarly, we find that $\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{w})=\mathrm{u}$. This means $\mathrm{m}_{\mathrm{u}}(\mathrm{T}) \mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{w})=0$ so that $\mathrm{m}_{\mathrm{u}}(\mathrm{x}) \mathrm{m}_{\mathrm{v}}(\mathrm{x}) \in \mathrm{N}_{\mathrm{T}}(\mathrm{w})$.

Now observe that $\mathrm{N}_{\mathrm{T}}(\mathrm{w})=\mathrm{m}_{\mathrm{w}}(\mathrm{x}) \mathcal{F}[\mathrm{x}]$ where $\mathrm{m}_{\mathrm{w}}(\mathrm{x})$ is the minimal polynomial of $w$. Then

$$
\mathrm{m}_{\mathrm{w}}(\mathrm{~T})(\mathrm{u})=\mathrm{m}_{\mathrm{w}}(\mathrm{~T}) \mathrm{m}_{\mathrm{v}}(\mathrm{~T})(\mathrm{w})=0
$$

and

$$
\mathrm{m}_{\mathrm{w}}(\mathrm{~T})(\mathrm{v})=\mathrm{m}_{\mathrm{w}}(\mathrm{~T}) \mathrm{m}_{\mathrm{u}}(\mathrm{~T})(\mathrm{w})=0
$$

so that $\mathrm{m}_{\mathrm{w}}(\mathrm{x}) \in \mathrm{N}_{\mathrm{T}}(\mathrm{u}) \cap \mathrm{N}_{\mathrm{T}}(\mathrm{v})$. Since $\mathrm{N}_{\mathrm{T}}(\mathrm{u})=\mathrm{m}_{\mathrm{u}}(\mathrm{x}) \mathcal{F}[\mathrm{x}]$ and $\mathrm{N}_{\mathrm{T}}(\mathrm{v})=$ $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mathcal{F}[\mathrm{x}]$, we see from Example 6.9 (along with the fact that $\mathrm{m}_{\mathrm{u}}(\mathrm{x})$ and $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ are relatively prime) that

$$
\mathrm{m}_{\mathrm{w}}(\mathrm{x}) \in \mathrm{m}(\mathrm{x}) \mathcal{F}[\mathrm{x}]=\mathrm{m}_{\mathrm{u}}(\mathrm{x}) \mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mathcal{F}[\mathrm{x}]
$$

and hence $m_{u}(x) m_{v}(x) \mid m_{w}(x)$. On the other hand, since

$$
\mathrm{m}_{\mathrm{u}}(\mathrm{x}) \mathrm{m}_{\mathrm{v}}(\mathrm{x}) \in \mathrm{N}_{\mathrm{T}}(\mathrm{w})=\mathrm{m}_{\mathrm{w}}(\mathrm{x}) \mathcal{F}[\mathrm{x}]
$$

we have $\mathrm{m}_{\mathrm{w}}(\mathrm{x}) \mid \mathrm{m}_{\mathrm{u}}(\mathrm{x}) \mathrm{m}_{\mathrm{v}}(\mathrm{x})$. Since $\mathrm{m}_{\mathrm{w}}(\mathrm{x}), \mathrm{m}_{\mathrm{u}}(\mathrm{x})$ and $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ are monic, it follows that $\mathrm{m}_{\mathrm{w}}(\mathrm{x})=\mathrm{m}_{\mathrm{u}}(\mathrm{x}) \mathrm{m}_{\mathrm{v}}(\mathrm{x})=\mathrm{m}(\mathrm{x})$. This shows that in the case where $\mathrm{m}_{\mathrm{u}}(\mathrm{x})$ and $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ are relatively prime, then $\mathrm{m}(\mathrm{x})=\mathrm{m}_{\mathrm{u}}(\mathrm{x}) \mathrm{m}_{\mathrm{v}}(\mathrm{x})$ is the minimal polynomial of $w$.

Now let $\mathrm{d}(\mathrm{x})$ be the greatest common divisor of $\mathrm{m}_{\mathrm{u}}(\mathrm{x})$ and $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$, and consider the general case where (see Theorem 6.9) $\mathrm{m}_{\mathrm{u}}(\mathrm{x}) \mathrm{m}_{\mathrm{v}}(\mathrm{x})=\mathrm{m}(\mathrm{x}) \mathrm{d}(\mathrm{x})$. Using the notation of Theorem 6.10, we write $m_{u}=\alpha \beta$ and $m_{v}=\gamma \delta$. Since $m_{u}$ and $\mathrm{m}_{\mathrm{v}}$ are minimal polynomials by hypothesis, Theorem 7.15 tells us that $\alpha$ and $\delta$ are each also the minimal polynomial of some vector. However, by their construction, $\alpha$ and $\delta$ are relatively prime since they have no factors in common. This means that we may apply the first part of this proof to conclude that $\alpha \delta$ is the minimal polynomial of some vector. To finish the proof, we simply note that (according to Theorem 6.10) $\alpha \delta$ is just $m(x)$.

A straightforward induction argument gives the following result.
Corollary For each $\mathrm{i}=1, \ldots, \mathrm{k}$ let $\mathrm{m}_{\mathrm{v}_{\mathrm{i}}}(\mathrm{x})$ be the minimal polynomial of a vector $\mathrm{v}_{\mathrm{i}} \in \mathrm{V}$. Then there exists a vector $\mathrm{w} \in \mathrm{V}$ whose minimal polynomial $\mathrm{m}(\mathrm{x})$ is the least common multiple of the $\mathrm{m}_{\mathrm{v}_{\mathrm{i}}}(\mathrm{x})$.

Now suppose that $T \in L(V)$ and $V$ has a basis $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{n}}\right\}$. If $\mathrm{m}_{\mathrm{v}_{\mathrm{i}}}(\mathrm{x})$ is the minimal polynomial of $v_{i}$, then by the corollary to Theorem 7.16, the least common multiple $\mathrm{m}(\mathrm{x})$ of the $\mathrm{m}_{\mathrm{v}_{\mathrm{i}}}(\mathrm{x})$ is the minimal polynomial of some vector $\mathrm{w} \in \mathrm{V}$, and therefore $\operatorname{deg} \mathrm{m}(\mathrm{x}) \leq \operatorname{dim} \mathrm{V}=\mathrm{n}$. But $\mathrm{m}(\mathrm{x})$ is the least common multiple, so that for each $i=1, \ldots, n$ we have $m(x)=f_{i}(x) m_{v_{i}}(x)$ for some $\mathrm{f}_{\mathrm{i}}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$. This means that

$$
\mathrm{m}(\mathrm{~T})\left(\mathrm{v}_{\mathrm{i}}\right)=\left[\mathrm{f}_{\mathrm{i}}(\mathrm{~T}) \mathrm{m}_{\mathrm{v}_{\mathrm{i}}}(\mathrm{~T})\right]\left(\mathrm{v}_{\mathrm{i}}\right)=\mathrm{f}_{\mathrm{i}}(\mathrm{~T})\left[\mathrm{m}_{\mathrm{v}_{\mathrm{i}}}(\mathrm{~T})\left(\mathrm{v}_{\mathrm{i}}\right)\right]=0
$$

for each $v_{i}$, and hence $m(T)=0$. In other words, every $T \in L(V)$ satisfies some monic polynomial $\mathrm{m}(\mathrm{x})$ with $\operatorname{deg} \mathrm{m}(\mathrm{x}) \leq \operatorname{dim} \mathrm{V}=\mathrm{n}$.

We now define the (nonempty) set

$$
\mathrm{N}_{\mathrm{T}}=\{\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]: \mathrm{f}(\mathrm{~T})=0\}
$$

As was the case with the T -annihilator, it is easy to prove that $\mathrm{N}_{\mathrm{T}}$ is an ideal of $\mathcal{F}[\mathrm{x}]$. Since $\mathrm{N}_{\mathrm{T}}$ consists of those polynomials in T that annihilate every vector in V , it must be the same as the intersection of all T-annihilators $\mathrm{N}_{\mathrm{T}}(\mathrm{v})$ in V , i.e.,

$$
\mathrm{N}_{\mathrm{T}}=\cap_{\mathrm{v} \in \mathrm{~V}} \mathrm{~N}_{\mathrm{T}}(\mathrm{v}) .
$$

By Theorem 6.8 the ideal $\mathrm{N}_{\mathrm{T}}$ is principal, and we define the minimal polynomial for $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ to be the unique monic generator of $\mathrm{N}_{\mathrm{T}}$. We claim that the minimal polynomial for T is precisely the polynomial $\mathrm{m}(\mathrm{x})$ defined in the previous paragraph.

To see this, note first that $\operatorname{deg} \mathrm{m}(\mathrm{x}) \leq \operatorname{dim} \mathrm{V}=\mathrm{n}$, and since $\mathrm{m}(\mathrm{x})$ is the minimal polynomial of some $w \in V$, it follows directly from the definition of the minimal polynomial of $w$ as the unique monic generator of $\mathrm{N}_{\mathrm{T}}(\mathrm{w})$ that $\mathrm{N}_{\mathrm{T}}(\mathrm{w})=\mathrm{m}(\mathrm{x}) \mathcal{F}[\mathrm{x}]$. Next, the fact that $\mathrm{m}(\mathrm{T})=0$ means that $\mathrm{m}(\mathrm{T})(\mathrm{v})=0$ for every $\mathrm{v} \in \mathrm{V}$, and therefore $\mathrm{m}(\mathrm{x}) \in \cap_{\mathrm{v} \in \mathrm{V}} \mathrm{N}_{\mathrm{T}}(\mathrm{v})=\mathrm{N}_{\mathrm{T}}$. Since any polynomial in $N_{T}(w)$ is a multiple of $m(x)$ and hence annihilates every $v \in V$, we see that $N_{T}(w) \subset N_{T}$. Conversely, any element of $N_{T}$ is automatically an element of $\mathrm{N}_{\mathrm{T}}(\mathrm{w})$, and thus $\mathrm{N}_{\mathrm{T}}=\mathrm{N}_{\mathrm{T}}(\mathrm{w})=\mathrm{m}(\mathrm{x}) \mathcal{F}[\mathrm{x}]$. This shows that $\mathrm{m}(\mathrm{x})$ is the minimal polynomial for $T$, and since $m(x)$ generates $N_{T}$, it is the polynomial of least degree such that $\mathrm{m}(\mathrm{T})=0$.

Example 7.6 Let $V=\mathbb{R}^{4}$ have basis $\left\{\mathrm{e}_{1}, \mathrm{e}_{2}, \mathrm{e}_{3}, \mathrm{e}_{4}\right\}$ and define the operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ by

$$
\begin{array}{ll}
T\left(e_{1}\right)=e_{1}+e_{3} & T\left(e_{2}\right)=3 e_{2}-e_{4} \\
T\left(e_{3}\right)=3 e_{1}-e_{3} & T\left(e_{4}\right)=3 e_{2}-e_{4}
\end{array}
$$

Note that since $T\left(e_{2}\right)=T\left(e_{4}\right)$, the matrix representation of $T$ has zero determinant, and hence T is singular (either by Theorem 5.9 or Theorem 5.16). Alternatively, we have $T\left(e_{2}-e_{4}\right)=0$ so that $T$ must be singular since $e_{2}-e_{4}$ $\neq 0$. In any case, we now have

$$
\mathrm{T}^{2}\left(\mathrm{e}_{1}\right)=\mathrm{T}\left(\mathrm{e}_{1}+\mathrm{e}_{3}\right)=\mathrm{T}\left(\mathrm{e}_{1}\right)+\mathrm{T}\left(\mathrm{e}_{3}\right)=4 \mathrm{e}_{1}
$$

so that

$$
\left(\mathrm{T}^{2}-4\right)\left(\mathrm{e}_{1}\right)=(\mathrm{T}-2)(\mathrm{T}+2)=0 .
$$

Similarly

$$
\mathrm{T}^{2}\left(\mathrm{e}_{2}\right)=\mathrm{T}\left(3 \mathrm{e}_{2}-\mathrm{e}_{4}\right)=3 \mathrm{~T}\left(\mathrm{e}_{2}\right)-\mathrm{T}\left(\mathrm{e}_{4}\right)=6 \mathrm{e}_{2}-2 \mathrm{e}_{4}=2 \mathrm{~T}\left(\mathrm{e}_{2}\right)
$$

so that

$$
\left(\mathrm{T}^{2}-2 \mathrm{~T}\right)\left(\mathrm{e}_{2}\right)=\mathrm{T}(\mathrm{~T}-2)\left(\mathrm{e}_{2}\right)=0 .
$$

Thus the minimal polynomial of $e_{1}$ is given by

$$
m_{1}(x)=x^{2}-4=(x-2)(x+2)
$$

and the minimal polynomial of $\mathrm{e}_{2}$ is given by

$$
\mathrm{m}_{2}(\mathrm{x})=\mathrm{x}(\mathrm{x}-2)
$$

That these are indeed minimal polynomials is clear if we note that in neither case will a linear expression in $T$ annihilate either $e_{1}$ or $e_{2}$ (just look at the definition of T).

It should be obvious that the least common multiple of $m_{1}$ and $m_{2}$ is

$$
x(x-2)(x+2)=x\left(x^{2}-4\right)
$$

and hence (by Theorem 7.16) this is the minimal polynomial of some vector $w \in \mathbb{R}^{4}$ which we now try to find. We know that $m_{1}(x)=x^{2}-4$ is the minimal polynomial of $e_{1}$, but is $x$ the minimal polynomial of some vector $u$ ? Since $m_{2}(x)=x(x-2)$ is the minimal polynomial of $e_{2}$, we see from Theorem 7.15 that the vector $u=(T-2)\left(e_{2}\right)=e_{2}-e_{4}$ has minimal polynomial $x$. Now, $x$ and $x^{2}-4$ are relatively prime so, as was done in the first part of the proof of Theorem 7.16, we define the polynomials $\mathrm{h}_{1}(\mathrm{x})=\mathrm{x} / 4$ and $\mathrm{k}_{1}(\mathrm{x})=-1 / 4$ by the requirement that $\mathrm{xh}_{1}+\left(\mathrm{x}^{2}-4\right) \mathrm{k}_{1}=1$. Hence

$$
\mathrm{w}=(\mathrm{T} / 4)\left(\mathrm{e}_{1}\right)+(-1 / 4)(\mathrm{u})=(1 / 4)\left(\mathrm{e}_{1}-\mathrm{e}_{2}+\mathrm{e}_{3}+\mathrm{e}_{4}\right)
$$

is the vector with minimal polynomial $\mathrm{x}\left(\mathrm{x}^{2}-4\right)$.
We leave it to the reader to show that $\mathrm{T}^{2}\left(\mathrm{e}_{3}\right)=4 \mathrm{e}_{3}$ and $\mathrm{T}^{2}\left(\mathrm{e}_{4}\right)=2 \mathrm{~T}\left(\mathrm{e}_{4}\right)$, and thus $e_{3}$ has minimal polynomial $m_{3}(x)=x^{2}-4$ and $e_{4}$ has minimal polynomial $m_{4}(x)=x(x-2)$. It is now easy to see that $m(x)=x\left(x^{2}-4\right)$ is the minimal polynomial for $T$ since $m(x)$ has the property that $m(T)\left(e_{i}\right)=0$ for each $\mathrm{i}=1, \ldots, 4$ and it is the least common multiple of the $\mathrm{m}_{\mathrm{i}}(\mathrm{x})$. We also note that the constant term in $\mathrm{m}(\mathrm{x})$ is zero, and hence T is singular by Theorem 7.5. //

Example 7.7 Relative to the standard basis for $\mathbb{R}^{3}$, the matrix

$$
A=\left(\begin{array}{rrr}
2 & -1 & 0 \\
1 & 0 & -2 \\
0 & 1 & 1
\end{array}\right)
$$

represents the operator $T \in L\left(\mathbb{R}^{3}\right)$ defined by $T e_{1}=2 e_{1}+e_{2}, T e_{2}=-e_{1}+e_{3}$ and $T e_{3}=-2 e_{2}+e_{3}$. It is easy to see that $T^{2} e_{1}=3 e_{1}+2 e_{2}+e_{3}$ and $T^{3} e_{1}=$ $4 e_{1}+e_{2}+3 e_{3}$, and hence the set $\left\{e_{1}, T e_{1}, T^{2} e_{1}\right\}$ is linearly independent, while the four vectors $\left\{\mathrm{e}_{1}, \mathrm{Te}_{1}, \mathrm{~T}^{2} \mathrm{e}_{1}, \mathrm{~T}^{3} \mathrm{e}_{1}\right\}$ must be linearly dependent. This means there exists $\mathrm{a}, \mathrm{b}, \mathrm{c}, \mathrm{d} \in \mathbb{R}$ such that

$$
\mathrm{ae}_{1}+\mathrm{bTe}_{2}+\mathrm{cT}^{2} \mathrm{e}_{1}+\mathrm{dT}^{3} \mathrm{e}_{1}=0
$$

This is equivalent to the system

$$
\begin{aligned}
a+2 b+3 c+4 d & =0 \\
b+2 c+d & =0 \\
c+3 d & =0
\end{aligned}
$$

Since there is one free variable, choosing $d=1$ we find $a=-5, b=5$ and $c=$ -3 . Therefore the minimal polynomial of $e_{1}$ is $m_{1}(x)=x^{3}-3 x^{2}+5 x-5$. Since the lcm of the minimal polynomials of $e_{1}, e_{2}$ and $e_{3}$ must be of degree $\leq$ 3 , it follows that $m_{1}(x)$ is in fact the minimal polynomial $m(x)$ of $T$ (and hence also of A). Note that by Theorem 7.5 we have $\mathrm{A}\left(\mathrm{A}^{2}-3 \mathrm{~A}+5 \mathrm{I}\right)=5 \mathrm{I}$, and hence $A^{-1}=(1 / 5)\left(A^{2}-3 A+5 I\right)$ or

$$
A^{-1}=\frac{1}{5}\left(\begin{array}{rrr}
2 & 1 & 2 \\
-1 & 2 & 4 \\
1 & -2 & 1
\end{array}\right)
$$

## Exercises

1. Show that $\mathrm{N}_{\mathrm{T}}$ is an ideal of $\mathcal{F}[\mathrm{x}]$.
2. For each of the following linear operators $T \in L(V)$, find the minimal polynomial of each of the (standard) basis vectors for V , find the minimal polynomial of T , and find a vector whose minimal polynomial is the same as that of T :
(a) $V=\mathbb{R}^{2}: \mathrm{Te}_{1}=\mathrm{e}_{2}, \mathrm{Te}_{2}=\mathrm{e}_{1}+\mathrm{e}_{2}$.
(b) $V=\mathbb{R}^{2}: T e_{1}=2 \mathrm{e}_{1}-3 \mathrm{e}_{2}, \mathrm{Te}_{2}=\mathrm{e}_{1}+5 \mathrm{e}_{2}$.
(c) $V=\mathbb{R}^{3}: T e_{1}=e_{1}-\mathrm{e}_{2}+\mathrm{e}_{3}, \mathrm{Te}_{2}=-2 \mathrm{e}_{2}+5 \mathrm{e}_{3}, \mathrm{Te}_{3}=2 \mathrm{e}_{1}+3 \mathrm{e}_{2}$.
(d) $V=\mathbb{R}^{3}: T e_{1}=2 \mathrm{e}_{2}, T e_{2}=2 \mathrm{e}_{1}, \mathrm{Te}_{3}=2 \mathrm{e}_{3}$.
(e) $V=\mathbb{R}^{4}: T e_{1}=e_{1}+e_{3}, T e_{2}=3 e_{4}, T e_{3}=e_{1}-e_{3}, T e_{4}=e_{2}$.
(f) $V=\mathbb{R}^{4}: T e_{1}=e_{1}+e_{2}, T e_{2}=e_{2}-e_{3}, T e_{3}=e_{3}+e_{4}, T e_{4}=e_{1}-e_{4}$.
3. For each of the following matrices, find its minimal polynomial over $\mathbb{R}$, and then find its inverse if it exists:
(a) $\left(\begin{array}{rr}2 & -1 \\ 1 & 3\end{array}\right)$
(b) $\left(\begin{array}{rr}4 & -1 \\ -4 & 1\end{array}\right)$
(c) $\left(\begin{array}{rrr}0 & 0 & 1 \\ 1 & 0 & -2 \\ 0 & 1 & 3\end{array}\right)$
(d) $\left(\begin{array}{rrr}2 & 0 & 0 \\ -1 & 3 & 0 \\ 5 & 2 & 1\end{array}\right)$
(d) $\left(\begin{array}{rrrr}1 & 1 & 0 & 0 \\ 4 & 1 & 0 & 0 \\ 0 & 0 & 0 & -3 \\ 0 & 0 & 1 & 4\end{array}\right)$
(e) $\left(\begin{array}{rrrr}1 & 0 & -1 & 0 \\ 0 & 1 & 0 & 1 \\ -1 & 1 & 0 & 0 \\ 0 & 0 & 1 & 1\end{array}\right)$

### 7.5 INVARIANT SUBSPACES

Recall that two matrices $\mathrm{A}, \mathrm{B} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ are said to be similar if there exists a nonsingular matrix $\mathrm{P} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ such that $\mathrm{B}=\mathrm{P}^{-1} \mathrm{AP}$. As was shown in Exercise 5.4.1, this defines an equivalence relation on $\mathrm{M}_{\mathrm{n}}(\mathcal{F})$. Since $\mathrm{L}(\mathrm{V})$ and $\mathrm{M}_{\mathrm{n}}(\mathcal{F})$ are isomorphic (Theorem 5.13), this definition applies equally well to linear operators. We call the equivalence class of a matrix (or linear operator) defined by this similarity relation its similarity class. While Theorem 7.14 gave us a condition under which a matrix may be diagonalized, this form is not possible to achieve in general. The approach we shall now take is to look for a basis in which the matrix of some linear operator in a given similarity class has a particularly simple standard form. As mentioned at the beginning of this chapter, these representations are called canonical forms. Of the many possible canonical forms, we shall consider only several of the more useful and important forms in this book. We begin with a discussion of some additional types of subspaces. A complete discussion of canonical forms under similarity is given in Chapter 8.

Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and let W be a subspace of V . Then W is said to be invariant under $\mathbf{T}$ (or simply $\mathbf{T}$-invariant) if $\mathrm{T}(\mathrm{w}) \in \mathrm{W}$ for every $\mathrm{w} \in \mathrm{W}$. For example, if $\mathrm{V}=\mathbb{R}^{3}$ then the xy-plane is invariant under the linear transformation that rotates every vector in $\mathbb{R}^{3}$ about the z -axis. As another example, note that if $\mathrm{v} \in \mathrm{V}$ is an eigenvector of T , then $\mathrm{T}(\mathrm{v})=\lambda \mathrm{v}$ for some $\lambda \in \mathcal{F}$, and hence v generates a one-dimensional subspace of V that is invariant under T (this is not necessarily the same as the eigenspace of $\lambda$ ).

Another way to describe the invariance of W under T is to say that $\mathrm{T}(\mathrm{W}) \subset \mathrm{W}$. Then clearly $\mathrm{T}^{2}(\mathrm{~W})=\mathrm{T}(\mathrm{T}(\mathrm{W})) \subset \mathrm{W}$, and in general $\mathrm{T}^{\mathrm{n}}(\mathrm{W}) \subset \mathrm{W}$ for every $\mathrm{n}=1,2, \ldots$. Since W is a subspace of V , this means that $\mathrm{f}(\mathrm{T})(\mathrm{W}) \subset \mathrm{W}$ for any $\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$. In other words, if W is invariant under T , then W is also invariant under any polynomial in T (over the same field as W ).

If $\mathrm{W} \subset \mathrm{V}$ is T -invariant, we may define the restriction of T to W in the usual way as that operator $\mathrm{T} \mid \mathrm{W}: \mathrm{W} \rightarrow \mathrm{W}$ defined by $(\mathrm{T} \mid \mathrm{W})(\mathrm{w})=\mathrm{T}(\mathrm{w})$ for every $\mathrm{w} \in \mathrm{W}$. We will frequently write $\mathrm{T}_{\mathrm{W}}$ instead of $\mathrm{T} \mid \mathrm{W}$.

Theorem 7.17 Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and let W be a T -invariant subspace of V . Then
(a) $\mathrm{f}\left(\mathrm{T}_{\mathrm{W}}\right)(\mathrm{w})=\mathrm{f}(\mathrm{T})(\mathrm{w})$ for any $\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$ and $\mathrm{w} \in \mathrm{W}$.
(b) The minimal polynomial $\mathrm{m}^{\prime}(\mathrm{x})$ for $\mathrm{T}_{\mathrm{W}}$ divides the minimal polynomial $m(x)$ for $T$.

## Proof This is Exercise 7.5.2.

If $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and $\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$, then $\mathrm{f}(\mathrm{T})$ is also a linear operator on V , and hence we may define the kernel (or null space) of $f(T)$ in the usual way by

$$
\operatorname{Ker} f(T)=\{v \in V: f(T)(v)=0\}
$$

Theorem 7.18 If $T \in L(V)$ and $f(x) \in \mathcal{F}[x]$, then $\operatorname{Ker} f(T)$ is a T-invariant subspace of V.

Proof Recall from Section 5.2 that $\operatorname{Ker} \mathrm{f}(\mathrm{T})$ is a subspace of V. To show that Ker $f(T)$ is T-invariant, we must show that $\operatorname{Tv} \in \operatorname{Ker} f(T)$ for any $v \in$ Ker $f(T)$, i.e., $f(T)(T v)=0$. But using Theorem 7.2(a) we see that

$$
\mathrm{f}(\mathrm{~T})(\mathrm{Tv})=\mathrm{T}(\mathrm{f}(\mathrm{~T})(\mathrm{v}))=\mathrm{T}(0)=0
$$

as desired.

Now suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and let $\mathrm{W} \subset \mathrm{V}$ be a T -invariant subspace. Furthermore let $\left\{v_{1}, v_{2}, \ldots, v_{n}\right\}$ be a basis for $V$, where the first $m<n$ vectors form a basis for W . If $\mathrm{A}=\left(\mathrm{a}_{\mathrm{ij}}\right)$ is the matrix representation of T relative to
this basis for V , then a little thought should convince you that A must be of the block matrix form

$$
A=\left(\begin{array}{ll}
B & C \\
0 & D
\end{array}\right)
$$

where $\mathrm{a}_{\mathrm{ij}}=0$ for $\mathrm{j} \leq \mathrm{m}$ and $\mathrm{i}>\mathrm{m}$. This is because $\mathrm{T}(\mathrm{w}) \in \mathrm{W}$ and any $\mathrm{w} \in \mathrm{W}$ has components $\left(\mathrm{w}_{1}, \ldots, \mathrm{w}_{\mathrm{m}}, 0, \ldots, 0\right)$ relative to the above basis for V . The formal proof of this fact is given in the following theorem.

Theorem 7.19 Let W be a subspace of V and suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$. Then W is T -invariant if and only if T can be represented in the block matrix form

$$
A=\left(\begin{array}{ll}
B & C \\
0 & D
\end{array}\right)
$$

where $B$ is a matrix representation of $\mathrm{T}_{\mathrm{W}}$.

Proof First suppose that W is T-invariant. Choose a basis $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{m}}\right\}$ for W , and extend this to a basis $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{m}}, \mathrm{v}_{\mathrm{m}+1}, \ldots, \mathrm{v}_{\mathrm{n}}\right\}$ for V (see Theorem 2.10). Then, since $T\left(v_{i}\right) \in W$ for each $i=1, \ldots, m$, there exist scalars $b_{i j}$ such that

$$
\mathrm{T}_{\mathrm{w}}\left(\mathrm{v}_{\mathrm{i}}\right)=\mathrm{T}\left(\mathrm{v}_{\mathrm{i}}\right)=\mathrm{v}_{\mathrm{l}} \mathrm{~b}_{\mathrm{li}}+\cdots+\mathrm{v}_{\mathrm{m}} \mathrm{~b}_{\mathrm{mi}}
$$

for each $\mathrm{i}=1, \ldots, \mathrm{~m}$. In addition, since $\mathrm{T}\left(\mathrm{v}_{\mathrm{i}}\right) \in \mathrm{V}$ for each $\mathrm{i}=\mathrm{m}+1, \ldots, \mathrm{n}$, there also exist scalars $\mathrm{c}_{\mathrm{ij}}$ and $\mathrm{d}_{\mathrm{ij}}$ such that

$$
\mathrm{T}\left(\mathrm{v}_{\mathrm{i}}\right)=\mathrm{v}_{1} \mathrm{c}_{\mathrm{li}}+\cdots+\mathrm{v}_{\mathrm{m}} \mathrm{c}_{\mathrm{mi}}+\mathrm{v}_{\mathrm{m}+1} \mathrm{~d}_{\mathrm{m}+1 \mathrm{i}}+\cdots+\mathrm{v}_{\mathrm{n}} \mathrm{~d}_{\mathrm{ni}}
$$

for each $\mathrm{i}=\mathrm{m}+1, \ldots, \mathrm{n}$.
From Theorem 5.11, we see that the matrix representation of T is given by an nx n matrix A of the form

$$
A=\left(\begin{array}{ll}
B & C \\
0 & D
\end{array}\right)
$$

where $B$ is an $m \times m$ matrix that represents $T_{W}, C$ is an $m x(n-m)$ matrix, and $D$ is an $(n-m) x(n-m)$ matrix.

Conversely, if A has the stated form and $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{n}}\right\}$ is a basis for V , then the subspace W of V defined by vectors of the form

$$
w=\sum_{i=1}^{m} \alpha_{i} v_{i}
$$

where each $\alpha_{i} \in \mathcal{F}$ will be invariant under $T$. Indeed, for each $i=1, \ldots, m$ we have

$$
T\left(v_{i}\right)=\sum_{j=1}^{n} v_{j} a_{j i}=v_{1} b_{1 i}+\cdots+v_{m} b_{m i} \in W
$$

and hence $\mathrm{T}(\mathrm{w}) \in \mathrm{W}$.

Corollary Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and W is a T-invariant subspace of V. Then the characteristic polynomial of $\mathrm{T}_{\mathrm{W}}$ divides the characteristic polynomial of T .

## Proof See Exercise 7.5.3.

Recall from Theorem 2.18 that the orthogonal complement $\mathrm{W}^{\perp}$ of a set $\mathrm{W} \subset \mathrm{V}$ is a subspace of V . If W is a subspace of V and both W and $\mathrm{W}^{\perp}$ are T invariant, then since $\mathrm{V}=\mathrm{W} \oplus \mathrm{W}^{\perp}$ (Theorem 2.22), a little more thought should convince you that the matrix representation of T will now be of the block diagonal form

$$
A=\left(\begin{array}{ll}
B & 0 \\
0 & D
\end{array}\right)
$$

We now proceed to discuss a variation of Theorem 7.19 in which we take into account the case where V can be decomposed into a direct sum of subspaces.

Let us assume that $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ where each $\mathrm{W}_{\mathrm{i}}$ is a T -invariant subspace of V . Then we define the restriction of T to $\mathrm{W}_{\mathrm{i}}$ to be the operator $\mathrm{T}_{\mathrm{i}}=\mathrm{T}_{\mathrm{W}_{\mathrm{i}}}=\mathrm{T} \mid \mathrm{W}_{\mathrm{i}}$. In other words, $\mathrm{T}_{\mathrm{i}}\left(\mathrm{w}_{\mathrm{i}}\right)=\mathrm{T}\left(\mathrm{w}_{\mathrm{i}}\right) \in \mathrm{W}_{\mathrm{i}}$ for any $\mathrm{w}_{\mathrm{i}} \in \mathrm{W}_{\mathrm{i}}$. Given any $\mathrm{v} \in \mathrm{V}$ we have $\mathrm{v}=\mathrm{v}_{1}+\cdots+\mathrm{v}_{\mathrm{r}}$ where $\mathrm{v}_{\mathrm{i}} \in \mathrm{W}_{\mathrm{i}}$ for each $\mathrm{i}=1, \ldots, r$, and hence

$$
T(v)=\sum_{i=1}^{r} T\left(v_{i}\right)=\sum_{i=1}^{r} T_{i}\left(v_{i}\right) .
$$

This shows that $T$ is completely determined by the effect of each $T_{i}$ on $W_{i}$. In this case we call T the direct sum of the $\mathrm{T}_{\mathrm{i}}$ and we write

$$
\mathrm{T}=\mathrm{T}_{1} \oplus \cdots \oplus \mathrm{~T}_{\mathrm{r}} .
$$

We also say that T is reducible (or decomposable) into the operators $\mathrm{T}_{\mathrm{i}}$, and the spaces $W_{i}$ are said to reduce $T$, or to form a $T$-invariant direct sum decomposition of V . In other words, T is reducible if there exists a basis for V such that $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ and each $\mathrm{W}_{\mathrm{i}}$ is T -invariant.

For each $\mathrm{i}=1, \ldots, \mathrm{r}$ we let $\mathrm{B}_{\mathrm{i}}=\left\{\mathrm{w}_{\mathrm{i} 1}, \ldots, \mathrm{w}_{\mathrm{in}}{ }_{\mathrm{i}}\right\}$ be a basis for $\mathrm{W}_{\mathrm{i}}$ so that $\mathrm{B}=\cup_{\mathrm{i}}$ is a basis for $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ (Theorem 2.15). We also let $\mathrm{A}_{\mathrm{i}}=$ $\left(a_{i, k j}\right)$ be the matrix representation of $T_{i}$ with respect to the basis $B_{i}$ (where $k$
and $j$ label the rows and columns respectively of the matrix $A_{i}$ ). Therefore we see that

$$
T\left(w_{i j}\right)=T_{i}\left(w_{i j}\right)=\sum_{k=1}^{n_{i}} w_{i k} a_{i, k j}
$$

where $\mathrm{i}=1, \ldots, \mathrm{r}$ and $\mathrm{j}=1, \ldots, \mathrm{n}_{\mathrm{i}}$. If A is the matrix representation of T with respect to the basis $B=\left\{w_{11}, \ldots, w_{1 n_{1}}, \ldots, w_{r 1}, \ldots, w_{r_{r}}\right\}$ for $V$, then since the i th column of A is just the image of the i th basis vector under T , we see that A must be of the block diagonal form

$$
\left(\begin{array}{cccc}
A_{1} & 0 & \cdots & 0 \\
0 & A_{2} & \cdots & 0 \\
\vdots & \vdots & & \vdots \\
0 & 0 & \cdots & A_{r}
\end{array}\right)
$$

If this is not immediately clear, then a minute's thought should help, keeping in mind that each $A_{i}$ is an $n_{i} \times n_{i}$ matrix, and $A$ is an $n \times n$ matrix where $n=$ $\sum_{i=1}^{\mathrm{r}} \mathrm{n}_{\mathrm{i}}$. It is also helpful to think of the elements of B as being numbered from 1 to n rather than by the confusing double subscripts (also refer to the proof of Theorem 7.19).

The matrix $A$ is called the direct sum of the matrices $A_{1}, \ldots, A_{r}$ and we write

$$
\mathrm{A}=\mathrm{A}_{1} \oplus \cdots \oplus \mathrm{~A}_{\mathrm{r}}
$$

In this case we also say that the matrix A is reducible. Thus a representation [ T ] of T is reducible if there exists a basis for V in which [ T ] is block diagonal. (Some authors say that a representation is reducible if there exists a basis for V in which the matrix of T is triangular. In this case, if there exists a basis for V in which the matrix is block diagonal, then the representation is said to be completely reducible. We shall not follow this convention.) This discussion proves the following theorem.

Theorem 7.20 Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and assume that $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ where each $W_{i}$ is $T$-invariant. If $A_{i}$ is the matrix representation of $T_{i}=T \mid W_{i}$, then the matrix representation of $T$ is given by the matrix $A=A_{1} \oplus \cdots \oplus A_{r}$.

Corollary Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ where each $\mathrm{W}_{\mathrm{i}}$ is T invariant. If $\Delta_{\mathrm{T}}(\mathrm{x})$ is the characteristic polynomial for T and $\Delta_{\mathrm{i}}(\mathrm{x})$ is the characteristic polynomial for $\mathrm{T}_{\mathrm{i}}=\mathrm{T} \mid \mathrm{W}_{\mathrm{i}}$, then $\Delta_{\mathrm{T}}(\mathrm{x})=\Delta_{\mathrm{l}}(\mathrm{x}) \cdots \Delta_{\mathrm{r}}(\mathrm{x})$.

## Proof See Exercise 7.5.4.

Example 7.8 Referring to Example 2.8, consider the space $V=\mathbb{R}^{3}$. We write $\mathrm{V}=\mathrm{W}_{1} \oplus \mathrm{~W}_{2}$ where $\mathrm{W}_{1}=\mathbb{R}^{2}$ (the xy-plane) and $\mathrm{W}_{2}=\mathbb{R}^{1}$ (the z -axis). Note that $\mathrm{W}_{1}$ has basis vectors $\mathrm{W}_{11}=(1,0,0)$ and $\mathrm{w}_{12}=(0,1,0)$, and $\mathrm{W}_{2}$ has basis vector $W_{21}=(0,0,1)$.

Now let $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ be the linear operator that rotates any $\mathrm{v} \in \mathrm{V}$ counterclockwise by an angle $\theta$ about the z-axis. Then clearly both $\mathrm{W}_{1}$ and $\mathrm{W}_{2}$ are $T$-invariant. Letting $\left\{e_{i}\right\}$ be the standard basis for $\mathbb{R}^{3}$, we have $T_{i}=T \mid W_{i}$ and consequently (see Example 1.2),

$$
\begin{aligned}
& T_{1}\left(e_{1}\right)=T\left(e_{1}\right)=(\cos \theta) e_{1}+(\sin \theta) e_{2} \\
& T_{1}\left(e_{2}\right)=T\left(e_{2}\right)=(-\sin \theta) e_{1}+(\cos \theta) e_{2} \\
& T_{2}\left(e_{3}\right)=T\left(e_{3}\right)=e_{3}
\end{aligned}
$$

Thus $\mathrm{V}=\mathrm{W}_{1} \oplus \mathrm{~W}_{2}$ is a T-invariant direct sum decomposition of V , and T is the direct sum of $T_{1}$ and $T_{2}$. It should be clear that the matrix representation of T is given by

$$
\left(\begin{array}{ccc}
\cos \theta & -\sin \theta & 0 \\
\sin \theta & \cos \theta & 0 \\
0 & 0 & 1
\end{array}\right)
$$

which is just the direct sum of the matrix representations of $\mathrm{T}_{1}$ and $\mathrm{T}_{2}$. //

## Exercises

1. Suppose $\mathrm{V}=\mathrm{W}_{1} \oplus \mathrm{~W}_{2}$ and let $\mathrm{T}_{1}: \mathrm{W}_{1} \rightarrow \mathrm{~V}$ and $\mathrm{T}_{2}: \mathrm{W}_{2} \rightarrow \mathrm{~V}$ be linear. Show that $T=T_{1} \oplus T_{2}$ is linear.
2. Prove Theorem 7.17.
3. Prove the corollary to Theorem 7.19.
4. Prove the corollary to Theorem 7.20.
5. Let V be a finite-dimensional inner product space over $\mathbb{C}$, and let G be a finite group. If for each $g \in G$ there is a $U(g) \in L(V)$ such that

$$
\mathrm{U}\left(\mathrm{~g}_{1}\right) \mathrm{U}\left(\mathrm{~g}_{2}\right)=\mathrm{U}\left(\mathrm{~g}_{1} \mathrm{~g}_{2}\right)
$$

then the collection $\mathrm{U}(\mathrm{G})=\{\mathrm{U}(\mathrm{g})\}$ is said to form a representation of G . If $W$ is a subspace of $V$ with the property that $U(g)(W) \subset W$ for all $g \in$

G , then we will say that W is $\mathrm{U}(\mathrm{G})$-invariant (or simply invariant). Furthermore, we say that $U(G)$ is irreducible if there is no nontrivial $\mathrm{U}(\mathrm{G})$-invariant subspace (i.e., the only invariant subspaces are $\{0\}$ and V itself ).
(a) Prove Schur's lemma 1: Let $\mathrm{U}(\mathrm{G})$ be an irreducible representation of $G$ on $V$. If $A \in L(V)$ is such that $A U(g)=U(g) A$ for all $g \in G$, then $A=$ $\lambda 1$ where $\lambda \in \mathbb{C}$. [Hint: Let $\lambda$ be an eigenvalue of $A$ with corresponding eigenspace $\mathrm{V}_{\lambda}$. Show that $\mathrm{V}_{\lambda}$ is $\mathrm{U}(\mathrm{G})$-invariant.]
(b) If $S \in L(V)$ is nonsingular, show that $U^{\prime}(G)=S U(G) S^{-1}$ is also a representation of $G$ on $V$. (Two representations of $G$ related by such a similarity transformation are said to be equivalent.)
(c) Prove Schur's lemma 2: Let $\mathrm{U}(\mathrm{G})$ and $\mathrm{U}^{\prime}(\mathrm{G})$ be two irreducible representations of $G$ on $V$ and $V^{\prime}$ respectively, and suppose $A \in L\left(V^{\prime}, V\right)$ is such that $A U^{\prime}(g)=U(g) A$ for all $g \in G$. Then either $A=0$, or else $A$ is an isomorphism of $\mathrm{V}^{\prime}$ onto V so that $\mathrm{A}^{-1}$ exists and $\mathrm{U}(\mathrm{G})$ is equivalent to $\mathrm{U}^{\prime}(\mathrm{G})$. [Hint: Show that $\operatorname{Im} A$ is invariant under $\mathrm{U}(\mathrm{G})$, and that Ker A is invariant under $\mathrm{U}^{\prime}(\mathrm{G})$.]
6. Suppose $A \in M_{n}(\mathcal{F})$ has minimal polynomial $m_{A}$ and $B \in M_{m}(\mathcal{F})$ has minimal polynomial $\mathrm{m}_{\mathrm{B}}$. Let $\mathrm{m}_{\mathrm{A} \oplus \mathrm{B}}$ be the minimal polynomial for $\mathrm{A} \oplus$ $B$ and let $\mathrm{p}=\operatorname{lcm}\left\{\mathrm{m}_{\mathrm{A}}, \mathrm{m}_{\mathrm{B}}\right\}$. Prove $\mathrm{m}_{\mathrm{A} \oplus \mathrm{B}}=\mathrm{p}$.
7. Let W be a T -invariant subspace of a finite-dimensional vector space V over $\mathcal{F}$, and suppose $\mathrm{v} \in \mathrm{V}$. Define the set

$$
\mathrm{N}_{\mathrm{T}}(\mathrm{v}, \mathrm{~W})=\{\mathrm{f} \in \mathcal{F}[\mathrm{x}]: \mathrm{f}(\mathrm{~T}) \mathrm{v} \in \mathrm{~W}\} .
$$

(a) Show that $\mathrm{N}_{\mathrm{T}}(\mathrm{v}, \mathrm{W})$ is an ideal of $\mathcal{F}[\mathrm{x}]$. This means that $\mathrm{N}_{\mathrm{T}}(\mathrm{v}, \mathrm{W})$ has a unique monic generator $c_{v}(x)$ which is called the T-conductor of $v$ into W.
(b) Show that every T-conductor divides the minimal polynomial $\mathrm{m}(\mathrm{x})$ for T .
(c) Now suppose the minimal polynomial for T is of the form

$$
m(x)=\left(x-\lambda_{1}\right)^{n_{1}} \cdots\left(x-\lambda_{r}\right)^{n_{r}}
$$

and let W be a proper T -invariant subspace of V . Prove there exists a vector $\mathrm{v} \in \mathrm{V}$ with $\mathrm{v} \notin \mathrm{W}$ such that $(\mathrm{T}-\lambda 1) \mathrm{v} \in \mathrm{W}$ where $\lambda$ is an eigenvalue of $T$. [Hint: Suppose $v_{1} \in V$ with $v_{1} \notin \mathrm{~W}$. Show that the T-conductor $\mathrm{c}_{\mathrm{v}_{1}}$ of $\mathrm{v}_{1}$ into W is of the form $\mathrm{c}_{\mathrm{v}_{1}}(\mathrm{x})=(\mathrm{x}-\boldsymbol{\lambda}) \mathrm{d}(\mathrm{x})$. Now consider the vector $\mathrm{v}=\mathrm{d}(\mathrm{T}) \mathrm{v}_{1}$.]
8. Let V be finite-dimensional over $\mathcal{F}$ and suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$. Prove there exists a basis for V in which the matrix representation A of T is uppertriangular if and only if the minimal polynomial for T is of the form $m(x)=\left(x-\lambda_{1}\right)^{n_{1}} \cdots\left(x-\lambda_{r}\right)^{n_{r}}$ where each $n_{i} \in \mathbb{Z}^{+}$and the $\lambda_{i}$ are the eigenvalues of T. [Hint: Apply part (c) of the previous problem to the basis $\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{n}}$ in which A is upper-triangular. Start with $\mathrm{W}=\{0\}$ to get $\mathrm{v}_{1}$, then consider the span of $\mathrm{v}_{1}$ to get $\mathrm{v}_{2}$, and continue this process.]
9. Relative to the standard basis for $\mathbb{R}^{2}$, let $\mathrm{T} \in \mathrm{L}\left(\mathbb{R}^{2}\right)$ be represented by

$$
A=\left(\begin{array}{rr}
1 & -1 \\
2 & 2
\end{array}\right) .
$$

(a) Prove that the only T-invariant subspaces of $\mathbb{R}^{2}$ are $\{0\}$ and $\mathbb{R}^{2}$ itself.
(b) Suppose $U \in L\left(\mathbb{C}^{2}\right)$ is also represented by A. Show that there exist one-dimensional U-invariant subspaces.
10. Find all invariant subspaces over $\mathbb{R}$ of the operator represented by

$$
A=\left(\begin{array}{ll}
2 & -5 \\
1 & -2
\end{array}\right)
$$

11. (a) Suppose $T \in L(V)$, and let $v \in V$ be arbitrary. Define the set of vectors

$$
\mathrm{Z}(\mathrm{v}, \mathrm{~T})=\{\mathrm{f}(\mathrm{~T})(\mathrm{v}): \mathrm{f} \in \mathcal{F}[\mathrm{x}]\}
$$

Show that $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ is a T -invariant subspace of V . (This is called the T cyclic subspace generated by v.)
(b) Let $v$ have minimal polynomial $m_{v}(x)=x^{r}+a_{r-1} x^{r-1}+\cdots+a_{r} x+$ $a_{0}$. Prove that $Z(v, T)$ has a basis $\left\{v, T v, \ldots, T^{r-1} v\right\}$, and hence also that $\operatorname{dim} \mathrm{Z}(\mathrm{v}, \mathrm{T})=\operatorname{deg} \mathrm{m}_{\mathrm{v}}(\mathrm{x})$. [Hint: Show that $\mathrm{T}^{\mathrm{k}} \mathrm{v}$ is a linear combination of $\left\{v, T v, \ldots, T^{r-1} v\right\}$ for every integer $k \geq r$.]
(c) Let $\mathrm{T} \in \mathrm{L}\left(\mathbb{R}^{3}\right)$ be represented in the standard basis by

$$
A=\left(\begin{array}{rrr}
1 & 0 & 1 \\
1 & 1 & -1 \\
1 & -2 & 2
\end{array}\right)
$$

If $v=e_{1}-e_{2}$, find the minimal polynomial of $v$ and a basis for $Z(v, T)$. Extend this to a basis for $\mathbb{R}^{3}$, and show that the matrix of $T$ relative to this basis is block triangular.

### 7.6 THE PRIMARY DECOMPOSITION THEOREM

We now proceed to show that there exists an important relationship between the minimal polynomial of a linear operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and the conditions under which V can be written as a direct sum of T -invariant subspaces of V . In the present context, this theorem (the primary decomposition theorem) is best obtained by first proving two relatively simple preliminary results.

Theorem 7.21 Suppose $T \in L(V)$ and assume that $\mathrm{f} \in \mathcal{F}[x]$ is a polynomial such that $f(T)=0$ and $f=h_{1} h_{2}$ where $h_{1}$ and $h_{2}$ are relatively prime. Define $W_{1}$ $=\operatorname{Ker} \mathrm{h}_{1}(\mathrm{~T})$ and $\mathrm{W}_{2}=\operatorname{Ker} \mathrm{h}_{2}(\mathrm{~T})$. Then $\mathrm{W}_{1}$ and $\mathrm{W}_{2}$ are T-invariant subspaces, and $\mathrm{V}=\mathrm{W}_{1} \oplus \mathrm{~W}_{2}$.

Proof We first note that $\mathrm{W}_{1}$ and $\mathrm{W}_{2}$ are T-invariant subspaces according to Theorem 7.18. Next, since $h_{1}$ and $h_{2}$ are relatively prime, there exist polynomials $g_{1}$ and $g_{2}$ such that $g_{1} h_{1}+g_{2} h_{2}=1$ (Theorem 6.5), and hence

$$
\begin{equation*}
g_{1}(T) h_{1}(T)+g_{2}(T) h_{2}(T)=1 . \tag{*}
\end{equation*}
$$

Then for any $\mathrm{v} \in \mathrm{V}$ we have

$$
g_{1}(T) h_{1}(T)(v)+g_{2}(T) h_{2}(T)(v)=v .
$$

However

$$
\begin{aligned}
h_{2}(T) g_{1}(T) h_{1}(T)(v) & =g_{1}(T) h_{1}(T) h_{2}(T)(v) \\
& =g_{1}(T) f(T)(v) \\
& =g_{1}(T) 0(v) \\
& =0
\end{aligned}
$$

so that $g_{1}(T) h_{1}(T)(v) \in \operatorname{Ker} h_{2}(T)=W_{2}$. Similarly, it is easy to see that $\mathrm{g}_{2}(\mathrm{~T}) \mathrm{h}_{2}(\mathrm{~T})(\mathrm{v}) \in \mathrm{W}_{1}$, and hence $\mathrm{v} \in \mathrm{W}_{1}+\mathrm{W}_{2}$. This shows that $\mathrm{V}=\mathrm{W}_{1}+\mathrm{W}_{2}$, and it remains to be shown that this sum is direct.

To show that this sum is direct we use Theorem 2.14. In other words, if $\mathrm{v}=\mathrm{w}_{1}+\mathrm{w}_{2}$ where $\mathrm{w}_{\mathrm{i}} \in \mathrm{W}_{\mathrm{i}}$, then we must show that each $\mathrm{w}_{\mathrm{i}}$ is uniquely determined by $v$. Applying $g_{1}(T) h_{1}(T)$ to $v=w_{1}+w_{2}$ and using the fact that $h_{1}(T)\left(w_{1}\right)=0$ we obtain

$$
\mathrm{g}_{1}(\mathrm{~T}) \mathrm{h}_{1}(\mathrm{~T})(\mathrm{v})=\mathrm{g}_{1}(\mathrm{~T}) \mathrm{h}_{1}(\mathrm{~T})\left(\mathrm{w}_{2}\right) .
$$

Next, we apply $(*)$ to $w_{2}$ and use the fact that $h_{2}(T)\left(w_{2}\right)=0$ to obtain

$$
\mathrm{g}_{1}(\mathrm{~T}) \mathrm{h}_{1}(\mathrm{~T})\left(\mathrm{w}_{2}\right)=\mathrm{w}_{2} .
$$

Combining these last two equations shows that $\mathrm{w}_{2}=\mathrm{g}_{1}(\mathrm{~T}) \mathrm{h}_{1}(\mathrm{~T})(\mathrm{v})$, and thus $w_{2}$ is uniquely determined by $v$ (through the action of $\left.g_{1}(T) h_{1}(T)\right)$. We leave it to the reader to show in a similar manner that $\mathrm{w}_{1}=\mathrm{g}_{2}(\mathrm{~T}) \mathrm{h}_{2}(\mathrm{~T})(\mathrm{v})$. Therefore $\mathrm{v}=\mathrm{w}_{1}+\mathrm{w}_{2}$ is a unique decomposition, and hence $\mathrm{V}=\mathrm{W}_{1} \oplus \mathrm{~W}_{2}$.

Theorem 7.22 Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$, and let $\mathrm{f}=\mathrm{h}_{1} \mathrm{~h}_{2} \in \mathcal{F}[\mathrm{x}]$ be the (monic) minimal polynomial for $T$, where $h_{1}$ and $h_{2}$ are relatively prime. If $W_{i}=$ Ker $h_{i}(T)$, then $h_{i}$ is the (monic) minimal polynomial for $T_{i}=T \mid W_{i}$.

Proof For each $\mathrm{i}=1$, 2 let $\mathrm{m}_{\mathrm{i}}$ be the (monic) minimal polynomial for $\mathrm{T}_{\mathrm{i}}$. Since $W_{i}=\operatorname{Ker} h_{i}(T)$, Theorem 7.17(a) tells us that $h_{i}\left(T_{i}\right)=0$, and therefore (by Theorem 7.4) we must have $m_{i} \mid h_{i}$. This means that $m_{i} \mid f$ ( since $f=h_{1} h_{2}$ ), and hence f is a multiple of $\mathrm{m}_{1}$ and $\mathrm{m}_{2}$. From the definition of least common multiple, it follows that the lcm of $m_{1}$ and $m_{2}$ must divide $f$. Since $h_{1}$ and $h_{2}$ are relatively prime, $\mathrm{m}_{1}$ and $\mathrm{m}_{2}$ must also be relatively prime (because if $\mathrm{m}_{1}$ and $m_{2}$ had a common factor, then $h_{1}$ and $h_{2}$ would each also have this same common factor since $m_{i} \mid h_{i}$ ). But $m_{1}$ and $m_{2}$ are monic, and hence their greatest common divisor is 1 . Therefore the lcm of $\mathrm{m}_{1}$ and $\mathrm{m}_{2}$ is just $\mathrm{m}_{1} \mathrm{~m}_{2}$ (Theorem 6.9). This shows that $m_{1} m_{2} \mid f$.

On the other hand, since $\mathrm{V}=\mathrm{W}_{1} \oplus \mathrm{~W}_{2}$ (Theorem 7.21), we see that for any $\mathrm{v} \in \mathrm{V}$

$$
\begin{aligned}
{\left[m_{1}(T) m_{2}(T)\right](v) } & =\left[m_{1}(T) m_{2}(T)\right]\left(w_{1}+w_{2}\right) \\
& =m_{2}(T)\left[m_{1}(T)\left(w_{1}\right)\right]+m_{1}(T)\left[m_{2}(T)\left(w_{2}\right)\right] \\
& =m_{2}(T)\left[m_{1}\left(T_{1}\right)\left(w_{1}\right)\right]+m_{1}(T)\left[m_{2}\left(T_{2}\right)\left(w_{2}\right)\right] \\
& =0
\end{aligned}
$$

because $m_{i}$ is the minimal polynomial for $T_{i}$. Therefore $\left(m_{1} m_{2}\right)(T)=0$, and hence $\mathrm{f} \mid \mathrm{m}_{1} \mathrm{~m}_{2}$ (from Theorem 7.4 since, by hypothesis, f is the minimal polynomial for $T$ ). Combined with the fact that $m_{1} m_{2} \mid f$, this shows that $f=m_{1} m_{2}$ (since $m_{1}, m_{2}$ and $f$ are monic). This result, along with the definition $f=h_{1} h_{2}$, the fact that $h_{1}$ and $h_{2}$ are monic, and the fact that $m_{i} \mid h_{i}$ shows that $m_{i}=h_{i}$.

We are now in a position to prove the main result of this section.

Theorem 7.23 (Primary Decomposition Theorem) Suppose $T \in L(V)$ has minimal polynomial

$$
m(x)=f_{1}(x)^{n_{1}} f_{2}(x)^{n_{2}} \cdots f_{r}(x)^{n_{r}}
$$

where each $f_{i}(x)$ is a distinct monic prime polynomial and each $n_{i}$ is a positive integer. Let $W_{i}=\operatorname{Ker} f_{i}(T)^{n_{i}}$, and define $T_{i}=T \mid W_{i}$. Then $V$ is the direct sum of the T-invariant subspaces $W_{i}$, and $f_{i}(x)^{n_{i}}$ is the minimal polynomial for $T_{i}$.

Proof If $\mathrm{r}=1$ the theorem is trivial since $\mathrm{W}_{1}=\operatorname{Ker} \mathrm{f}_{1}(\mathrm{~T})^{\mathrm{n}_{1}}=\operatorname{Ker} \mathrm{m}(\mathrm{T})=\mathrm{V}$. We now assume that the theorem has been proved for some $\mathrm{r}-1 \geq 1$, and proceed by induction to show that it is true for $r$. We first remark that the $W_{i}$ are T-invariant subspaces by Theorem 7.18. Define the T-invariant subspace

$$
\mathrm{U}=\operatorname{Ker}\left[\mathrm{f}_{2}(\mathrm{~T})^{\mathrm{n}_{2}} \cdots \mathrm{f}_{\mathrm{r}}(\mathrm{~T})^{\mathrm{n}_{\mathrm{r}}}\right]
$$

Because the $f_{i}(x)^{n_{i}}$ are relatively prime (by Corollary 2 of Theorem 6.5, since the $\mathrm{f}_{\mathrm{i}}(\mathrm{x})$ are all distinct primes), we can apply Theorem 7.21 to write $\mathrm{V}=$ $\mathrm{W}_{1} \oplus \mathrm{U}$. In addition, since $\mathrm{m}(\mathrm{x})$ is the mini-mal polynomial for T , Theorem 7.22 tells us that $f_{1}(x)^{n_{1}}$ is the minimal polynomial for $T_{1}$, and $\left[f_{2}(x)^{n_{2}} \ldots\right.$ $\left.f_{r}(x)^{n_{r}}\right]$ is the minimal polynomial for $T_{U}=T \mid U$.

Applying our induction hypothesis, we find that $\mathrm{U}=\mathrm{W}_{2} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ where for each $i=2, \ldots, r$ we have $W_{i}=\operatorname{Ker} f_{i}\left(T_{U}\right)^{n_{i}}$, and $f_{i}(x)^{n_{i}}$ is the minimal polynomial for $T_{i}=T_{U} \mid W_{i}$. However, it is obvious that $f_{i}(x)^{n_{i}}$ divides $\left[\mathrm{f}_{2}(\mathrm{x})^{\mathrm{n}_{2}} \cdots \mathrm{f}_{\mathrm{r}}(\mathrm{x})^{\mathrm{n}_{\mathrm{r}}}\right]$ for each $\mathrm{i}=2, \ldots, \mathrm{r}$ and hence Ker $\mathrm{f}_{\mathrm{i}}(\mathrm{T})^{\mathrm{n}_{\mathrm{i}}} \subset \mathrm{U}$. Specifically, this means that the set of all vectors $v \in V$ with the property that $f_{i}(T)^{n_{i}}(v)=0$ are also in $U$, and therefore $\operatorname{Ker} f_{i}(T)^{n_{i}}=\operatorname{Ker} f_{i}\left(T_{U}\right)^{n_{i}}=W_{i}$. Furthermore, $T\left|W_{i}=T_{U}\right| W_{i}=T_{i}$ and thus $f_{i}(x)^{n_{i}}$ is also the minimal polynomial for $\mathrm{T} \mid \mathrm{W}_{\mathrm{i}}$.

Summarizing, we have shown that $\mathrm{V}=\mathrm{W}_{1} \oplus \mathrm{U}=\mathrm{W}_{1} \oplus \mathrm{~W}_{2} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ where $W_{i}=\operatorname{Ker} f_{i}(T)^{n_{i}}$ for each $i=1, \ldots, r$ and $f_{i}(x)^{n_{i}}$ is the minimal polynomial for $\mathrm{T} \mid \mathrm{W}_{\mathrm{i}}=\mathrm{T}_{\mathrm{i}}$. This completes the induction procedure and proves the theorem.

In order to make this result somewhat more transparent, as well as in aiding actual calculations, we go back and look carefully at what we have done in defining the spaces $W_{i}=\operatorname{Ker} f_{i}(T)^{n_{i}}$. For each $i=1, \ldots, r$ we define the polynomials $\mathrm{g}_{\mathrm{i}}(\mathrm{x})$ by

$$
\mathrm{m}(\mathrm{x})=\mathrm{f}_{\mathrm{i}}(\mathrm{x})^{\mathrm{n}_{\mathrm{i}}} \mathrm{~g}_{\mathrm{i}}(\mathrm{x})
$$

In other words, $g_{i}(x)$ is a product of the $r-1$ factors $f_{j}(x)^{n_{j}}$ with $j \neq i$. We claim that in fact

$$
\mathrm{W}_{\mathrm{i}}=\mathrm{g}_{\mathrm{i}}(\mathrm{~T})(\mathrm{V}) .
$$

It is easy to see that $\mathrm{g}_{\mathrm{i}}(\mathrm{T})(\mathrm{V}) \subset \mathrm{W}_{\mathrm{i}}$ because

$$
\mathrm{f}_{\mathrm{i}}(\mathrm{~T})^{\mathrm{n}_{\mathrm{i}}}\left[\mathrm{~g}_{\mathrm{i}}(\mathrm{~T})(\mathrm{v})\right]=\left[\mathrm{f}_{\mathrm{i}}(\mathrm{~T})^{\mathrm{n}_{\mathrm{i}}} \mathrm{~g}_{\mathrm{i}}(\mathrm{~T})\right](\mathrm{v})=\mathrm{m}(\mathrm{~T})(\mathrm{v})=0
$$

for every $\mathrm{v} \in \mathrm{V}$. On the other hand, $\mathrm{f}_{\mathrm{i}}(\mathrm{x})^{\mathrm{n}_{\mathrm{i}}}$ and $\mathrm{g}_{\mathrm{i}}(\mathrm{x})$ are monic relative primes, and hence (by Theorem 6.5) there exist polynomials $\mathrm{a}(\mathrm{x}), \mathrm{b}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$ such that $a(x) f_{i}(x)^{n_{i}}+b(x) g_{i}(x)=1$. Then for any $v_{i} \in W_{i}=\operatorname{Ker} f_{i}(T)^{n_{i}}$ we have $\mathrm{f}_{\mathrm{i}}(\mathrm{T})^{\mathrm{n}_{\mathrm{i}}}\left(\mathrm{v}_{\mathrm{i}}\right)=0$, and hence

$$
\mathrm{v}_{\mathrm{i}}=\mathrm{a}(\mathrm{~T})\left[\mathrm{f}_{\mathrm{i}}(\mathrm{~T})^{\mathrm{n}_{\mathrm{i}}}\left(\mathrm{v}_{\mathrm{i}}\right)\right]+\mathrm{g}_{\mathrm{i}}(\mathrm{~T})\left[\mathrm{b}(\mathrm{~T})\left(\mathrm{v}_{\mathrm{i}}\right)\right]=0+\mathrm{g}_{\mathrm{i}}(\mathrm{~T})\left[\mathrm{b}(\mathrm{~T})\left(\mathrm{v}_{\mathrm{i}}\right)\right] \in \mathrm{g}_{\mathrm{i}}(\mathrm{~T})(\mathrm{V}) .
$$

Hence $W_{i} \subset g_{i}(T)(V)$, and therefore $W_{i}=g_{i}(T)(V)$ as claimed. This gives us, at least conceptually, a practical method for computing the matrix of a linear transformation T with respect to the bases of the T -invariant subspaces.

As a final remark, note that for any $\mathrm{j} \neq \mathrm{i}$ we have $\mathrm{g}_{\mathrm{i}}(\mathrm{T})\left(\mathrm{W}_{\mathrm{j}}\right)=0$ because $g_{i}(T)$ contains $f_{j}(T)^{n_{j}}$ and $W_{j}=\operatorname{Ker} f_{j}(T)^{n_{j}}$. In addition, since $W_{i}$ is T-invariant we see that $\mathrm{g}_{\mathrm{i}}(\mathrm{T})\left(\mathrm{W}_{\mathrm{i}}\right) \subset \mathrm{W}_{\mathrm{i}}$. But we also have

$$
\begin{aligned}
W_{i} & =a(T)\left[f_{i}(T)^{n_{i}}\left(W_{i}\right)\right]+g_{i}(T)\left[b(T)\left(W_{i}\right)\right] \\
& =0+g_{i}(T)\left[b(T)\left(W_{i}\right)\right] \\
& \subset g_{i}(T)\left(W_{i}\right)
\end{aligned}
$$

and hence $g_{i}(T)\left(W_{i}\right)=W_{i}$. This should not be surprising for the following reason. If we write

$$
\mathrm{V}_{\mathrm{i}}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{i}-1} \oplus \mathrm{~W}_{\mathrm{i}+1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}
$$

then $\mathrm{g}_{\mathrm{i}}(\mathrm{T})\left(\mathrm{V}_{\mathrm{i}}\right)=0$, and therefore

$$
\mathrm{W}_{\mathrm{i}}=\mathrm{g}_{\mathrm{i}}(\mathrm{~T})(\mathrm{V})=\mathrm{g}_{\mathrm{i}}(\mathrm{~T})\left(\mathrm{W}_{\mathrm{i}} \oplus \mathrm{~V}_{\mathrm{i}}\right)=\mathrm{g}_{\mathrm{i}}(\mathrm{~T})\left(\mathrm{W}_{\mathrm{i}}\right) .
$$

Example 7.9 Consider the space $V=\mathbb{R}^{3}$ with basis $\left\{\mathrm{u}_{1}, \mathrm{u}_{2}, \mathrm{u}_{3}\right\}$ and define the linear transformation $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ by

$$
\begin{aligned}
& T\left(u_{1}\right)=u_{2} \\
& T\left(u_{2}\right)=u_{3} \\
& T\left(u_{3}\right)=-2 u_{1}+3 u_{2} .
\end{aligned}
$$

Then

$$
\mathrm{T}^{2}\left(\mathrm{u}_{1}\right)=\mathrm{T}\left(\mathrm{u}_{2}\right)=\mathrm{u}_{3}
$$

and hence

$$
\mathrm{T}^{3}\left(\mathrm{u}_{1}\right)=\mathrm{T}\left(\mathrm{u}_{3}\right)=-2 \mathrm{u}_{1}+3 \mathrm{u}_{2}=-2 \mathrm{u}_{1}+3 \mathrm{~T}\left(\mathrm{u}_{1}\right) .
$$

Therefore $T^{3}\left(u_{1}\right)-3 T\left(u_{1}\right)+2 u_{1}=0$ so that the minimal polynomial of $u_{1}$ is given by

$$
m_{1}(x)=x^{3}-3 x+2=(x-1)^{2}(x+2)
$$

Now recall that the minimal polynomial $m(x)$ for $T$ is just the least common multiple of the minimal polynomials $\mathrm{m}_{\mathrm{i}}(\mathrm{x})$ of the basis vectors for V , and $\operatorname{deg} \mathrm{m}(\mathrm{x}) \leq \operatorname{dim} \mathrm{V}=\mathrm{n}$ (see the discussion prior to Example 7.6). Since $\mathrm{m}_{1}(\mathrm{x})$ is written as a product of prime polynomials with $\operatorname{deg} \mathrm{m}_{1}=3=\operatorname{dim} \mathrm{V}$, it follows that $m(x)=m_{1}(x)$. We thus have $f_{1}(x)^{n_{1}}=(x-1)^{2}$ and $f_{2}(x)^{n_{2}}=(x+2)$.

We now define

$$
\mathrm{W}_{1}=\mathrm{g}_{1}(\mathrm{~T})(\mathrm{V})=(\mathrm{T}+2)(\mathrm{V})
$$

and

$$
\mathrm{W}_{2}=\mathrm{g}_{2}(\mathrm{~T})(\mathrm{V})=(\mathrm{T}-1)^{2}(\mathrm{~V})
$$

A simple calculation shows that

$$
\begin{aligned}
& (T+2) u_{1}=2 u_{1}+u_{2} \\
& (T+2) u_{2}=2 u_{2}+u_{3} \\
& (T+2) u_{3}=-2 u_{1}+3 u_{2}+2 u_{3}=(T+2)\left(-u_{1}+2 u_{2}\right)
\end{aligned}
$$

Therefore $W_{1}$ is spanned by the basis vectors $\left\{2 u_{1}+u_{2}, 2 u_{2}+u_{3}\right\}$. Similarly, it is easy to show that $(T-1)^{2} u_{1}=u_{1}-2 u_{2}+u_{3}$ and that both $(T-1)^{2} u_{2}$ and $(T-1)^{2} u_{3}$ are multiples of this. Hence $\left\{u_{1}-2 u_{2}+u_{3}\right\}$ is the basis vector for $\mathrm{W}_{2}$.

We now see that $T_{1}=T \mid W_{1}$ and $T_{2}=T \mid W_{2}$ yield the transformations

$$
\begin{array}{ll}
T_{1}\left(2 u_{1}+u_{2}\right) & =2 u_{2}+u_{3} \\
T_{1}\left(2 u_{2}+u_{3}\right) & =-2 u_{1}+3 u_{2}+2 u_{3}=-\left(2 u_{1}+u_{2}\right)+2\left(2 u_{2}+u_{3}\right) \\
T_{2}\left(u_{1}-2 u_{2}+u_{3}\right) & =-2\left(u_{1}-2 u_{2}+u_{3}\right)
\end{array}
$$

and hence $T_{1}$ and $T_{2}$ are represented by the matrices $A_{1}$ and $A_{2}$, respectively, given by

$$
A_{1}=\left(\begin{array}{rr}
0 & -1 \\
1 & 2
\end{array}\right) \quad A_{2}=(-2)
$$

Therefore $\mathrm{T}=\mathrm{T}_{1} \oplus \mathrm{~T}_{2}$ is represented by the matrix $\mathrm{A}=\mathrm{A}_{1} \oplus \mathrm{~A}_{2}$ where

$$
A=\left(\begin{array}{ccc}
\begin{array}{|rr|}
0 & -1 \\
1 & 2 \\
\hline
\end{array} & 0 \\
0 & 0 & \boxed{-2}
\end{array}\right)
$$

From Corollary 1 of Theorem 7.14 we know that a matrix $A \in M_{n}(\mathcal{F})$ is diagonalizable if and only if A has n linearly independent eigenvectors, and from Theorem 7.13, the corresponding distinct eigenvalues $\lambda_{1}, \ldots, \lambda_{\mathrm{r}}$ (where $\mathrm{r} \leq \mathrm{n}$ ) must be roots of the minimal polynomial for A. The factor theorem (corollary to Theorem 6.4) then says that $x-\lambda_{i}$ is a factor of the minimal polynomial for A, and hence the minimal polynomial for A must contain at least r distinct linear factors if A is to be diagonalizable.

We now show that the minimal polynomial of a diagonalizable linear transformation consists precisely of distinct linear factors.

Theorem 7.24 A linear transformation $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is diagonalizable if and only if the minimal polynomial $\mathrm{m}(\mathrm{x})$ for T is of the form

$$
\mathrm{m}(\mathrm{x})=\left(\mathrm{x}-\lambda_{1}\right) \cdots\left(\mathrm{x}-\lambda_{\mathrm{r}}\right)
$$

where $\lambda_{1}, \ldots, \lambda_{r}$ are the distinct eigenvalues of T .

Proof Suppose that $\mathrm{m}(\mathrm{x})=\left(\mathrm{x}-\lambda_{1}\right) \cdots\left(\mathrm{x}-\lambda_{\mathrm{r}}\right)$ where $\lambda_{1}, \ldots, \lambda_{\mathrm{r}} \in \mathcal{F}$ are distinct. Then, according to Theorem 7.23, $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ where $\mathrm{W}_{\mathrm{i}}=$ $\operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)$. But then for any $\mathrm{w} \in \mathrm{W}_{\mathrm{i}}$ we have

$$
0=\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)(\mathrm{w})=\mathrm{T}(\mathrm{w})-\lambda_{\mathrm{i}} \mathrm{w}
$$

and hence any $\mathrm{w} \in \mathrm{W}_{\mathrm{i}}$ is an eigenvector of T with eigenvalue $\lambda_{\mathrm{i}}$. It should be clear that any eigenvector of T with eigenvalue $\lambda_{\mathrm{i}}$ is also in $\mathrm{W}_{\mathrm{i}}$. In particular, this means that any basis vector of $W_{i}$ is also an eigenvector of $T$. By Theorem 2.15, the union of the bases of all the $\mathrm{W}_{\mathrm{i}}$ is a basis for V , and hence V has a basis consisting entirely of eigenvectors. This means that T is diagonalizable (Theorem 7.14).

On the other hand, assume that T is diagonalizable, and hence V has a basis $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{n}}\right\}$ of eigenvectors of T that correspond to the (not necessarily
distinct) eigenvalues $\lambda_{1}, \ldots, \lambda_{\mathrm{n}}$. If the $\mathrm{v}_{\mathrm{i}}$ are numbered so that $\lambda_{1}, \ldots, \lambda_{\mathrm{r}}$ are the distinct eigenvalues of T , then the operator

$$
\mathrm{f}(\mathrm{~T})=\left(\mathrm{T}-\lambda_{1} 1\right) \cdots\left(\mathrm{T}-\lambda_{\mathrm{r}} 1\right)
$$

has the property that $f(T)\left(v_{i}\right)=0$ for each of the basis eigenvectors $v_{1}, \ldots, v_{n}$. Thus $f(T)=0$ and the minimal polynomial $m(x)$ for $T$ must divide $f(x)$ (Theorem 7.4). While this shows that $\mathrm{m}(\mathrm{x})$ must consist of linear factors, it is in fact also true that $f(x)=m(x)$. To see this, suppose that we delete any factor $T-\lambda_{\alpha} 1$ from $f(T)$ to obtain a new linear operator $f^{\prime}(T)$. But the $\lambda_{i}$ are all distinct so that $\mathrm{f}^{\prime}(\mathrm{T})\left(\mathrm{v}_{\alpha}\right) \neq 0$, and hence $\mathrm{f}^{\prime}(\mathrm{T}) \neq 0$. Therefore $\mathrm{f}^{\prime}(\mathrm{x})$ cannot be the minimal polynomial for $T$, and we must have $f(x)=m(x)$.

In a manner similar to that used in Corollary 1 of Theorem 7.14, we can rephrase Theorem 7.24 in terms of matrices as follows.

Corollary 1 A matrix $A \in M_{n}(\mathcal{F})$ is similar to a diagonal matrix $D$ if and only if the minimal polynomial for A has the form

$$
\mathrm{m}(\mathrm{x})=\left(\mathrm{x}-\lambda_{\mathrm{I}}\right) \cdots\left(\mathrm{x}-\lambda_{\mathrm{r}}\right)
$$

where $\lambda_{1}, \ldots, \lambda_{\mathrm{r}} \in \mathcal{F}$ are all distinct. If this is the case, then $\mathrm{D}=\mathrm{P}^{-1} \mathrm{AP}$ where P is the invertible matrix whose columns are any set of n linearly independent eigenvectors $\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{n}}$ of A corresponding to the eigenvalues $\boldsymbol{\lambda}_{1}, \ldots, \boldsymbol{\lambda}_{\mathrm{n}}$. (If $\mathrm{r}<\mathrm{n}$, then some of the eigenvalues will be repeated.) In addition, the diagonal elements of $D$ are just the eigenvalues $\lambda_{1}, \ldots, \lambda_{n}$.

Corollary 2 A linear transformation $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is diagonalizable if and only if $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ where $\mathrm{W}_{\mathrm{i}}=\operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)=\mathrm{V}_{\lambda_{\mathrm{i}}}$.

Proof Recall that $\mathrm{V}_{\lambda_{\mathrm{i}}}$ is the eigenspace corresponding to the eigenvalue $\lambda_{\mathrm{i}}$, and the fact that $\mathrm{V}_{\lambda_{\mathrm{i}}}=\operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)$ was shown in the proof of Theorem 7.24. If T is diagonalizable, then the conclusion that $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ follows directly from Theorems 7.24 and 7.23. On the other hand, if $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus$ $\mathrm{W}_{\mathrm{r}}$ then each $\mathrm{W}_{\mathrm{i}}$ has a basis of eigenvectors and hence so does V (Theorem 2.15).

It is important to realize that one eigenvalue can correspond to more than one linearly independent eigenvector (recall the comment following Theorem 7.8). This is why the spaces $\mathrm{W}_{\mathrm{i}}$ in the first part of the proof of Theorem 7.24 can have bases consisting of more than one eigenvector. In particular, any eigenvalue of multiplicity greater than one can result in an eigenspace of
dimension greater than one, a result that we treat in more detail in the next section.

## Exercises

1. Write each of the following linear transformations $T \in L(V)$ as a direct sum of linear transformations whose minimal polynomials are powers of prime polynomials:
(a) $\mathrm{V}=\mathbb{R}^{2}: \mathrm{Te}_{1}=\mathrm{e}_{2}, \mathrm{Te}_{2}=3 \mathrm{e}_{1}+2 \mathrm{e}_{2}$.
(b) $V=\mathbb{R}^{2}: T e_{1}=-4 e_{1}+4 e_{2}, T e_{2}=e_{2}$.
(c) $V=\mathbb{R}^{3}: T e_{1}=e_{2}, T e_{2}=e_{3}, T e_{3}=2 e_{1}-e_{2}+2 e_{3}$.
(d) $V=\mathbb{R}^{3}: T e_{1}=3 e_{1}, T e_{2}=e_{2}-e_{3}, T e_{3}=e_{1}+3 e_{2}$.
(e) $V=\mathbb{R}^{3}: T e_{1}=3 e_{1}+e_{2}, T e_{2}=e_{2}-5 e_{3}, T e_{3}=2 e_{1}+2 \mathrm{e}_{2}+2 \mathrm{e}_{3}$.
2. Let V be finite-dimensional and suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ has minimal polynomial $m(x)=f_{1}(x)^{n_{1}} \cdots f_{r}(x)^{n_{r}}$ where the $f_{i}(x)$ are distinct monic primes and each $n_{i} \in \mathbb{Z}^{+}$. Show that the characteristic polynomial is of the form

$$
\Delta(\mathrm{x})=\mathrm{f}_{1}(\mathrm{x})^{\mathrm{d}_{1}} \cdots \mathrm{f}_{\mathrm{r}}(\mathrm{x})^{\mathrm{d}_{\mathrm{r}}}
$$

where

$$
d_{i}=\operatorname{dim}\left(\operatorname{Ker} f_{i}(T)^{n_{i}}\right) / \operatorname{deg} f_{i}
$$

3. Let $\mathcal{D}=\left\{\mathrm{T}_{\mathrm{i}}\right\}$ be a collection of mutually commuting (i.e., $\mathrm{T}_{\mathrm{i}} \mathrm{T}_{\mathrm{j}}=\mathrm{T}_{\mathrm{j}} \mathrm{T}_{\mathrm{i}}$ for all $\mathrm{i}, \mathrm{j}$ ) diagonalizable linear operators on a finite-dimensional space V . Prove that there exists a basis for V relative to which the matrix representation of each $\mathrm{T}_{\mathrm{i}}$ is diagonal. [Hint: Proceed by induction on dim V. Let $\mathrm{T} \in \mathcal{D}(\mathrm{T} \neq \mathrm{c} 1)$ have distinct eigenvalues $\lambda_{1}, \ldots, \lambda_{\mathrm{r}}$ and for each i define $\mathrm{W}_{\mathrm{i}}=\operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)$. Show that $\mathrm{W}_{\mathrm{i}}$ is invariant under each operator that commutes with T. Define $\mathcal{D}_{\mathrm{i}}=\left\{\mathrm{T}_{\mathrm{j}} \mid \mathrm{W}_{\mathrm{i}}: \mathrm{T}_{\mathrm{j}} \in \mathcal{D}\right\}$ and show that every member of $\mathcal{D}_{\mathrm{i}}$ is diagonalizable.]

### 7.7 MORE ON DIAGONALIZATION

If an operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is diagonalizable, then in a (suitably numbered) basis of eigenvectors, its matrix representation A will take the form

$$
A=\left(\begin{array}{cccccc}
\lambda_{1} I_{m_{1}} & 0 & 0 & \cdots & 0 & 0 \\
0 & \lambda_{2} I_{m_{2}} & 0 & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & \cdots & 0 & \lambda_{r} I_{m_{r}}
\end{array}\right)
$$

where each $\lambda_{i}$ is repeated $m_{i}$ times and $I_{m_{i}}$ is the $m_{i} \times m_{i}$ identity matrix. Note that $m_{1}+\cdots+m_{r}$ must be equal to dim $V$. Thus the characteristic polynomial for T has the form

$$
\Delta_{T}(x)=\operatorname{det}(x I-A)=\left(x-\lambda_{1}\right)^{m_{1}} \cdots\left(x-\lambda_{r}\right)^{m_{r}}
$$

which is a product of (possibly repeated) linear factors. That the characteristic polynomial of a diagonalizable operator is of this form also follows directly from Theorems 7.24 and 7.12. However, we stress that just because the characteristic polynomial factors into a product of linear terms does not mean that the operator is diagonalizable. We now investigate the conditions that determine just when an operator will be diagonalizable.

Let us assume that T is diagonalizable, and hence that the characteristic polynomial factors into linear terms. For each distinct eigenvalue $\lambda_{i}$, we have seen that the corresponding eigenspace $\mathrm{V}_{\lambda_{\mathrm{i}}}$ is just $\operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)$. Relative to a basis of eigenvectors, the matrix [ $\mathrm{T}-\lambda_{\mathrm{i}} 1$ ] is diagonal with precisely $\mathrm{m}_{\mathrm{i}}$ zeros along its main diagonal (just look at the matrix A shown above and subtract off $\lambda_{i}$ I). From Theorem 5.15 we know that the rank of a linear transformation is the same as the rank of its matrix representation, and hence $r\left(T-\lambda_{i} 1\right)$ is just the number of remaining nonzero rows in $\left[\mathrm{T}-\lambda_{\mathrm{i}} 1\right.$ ] which is $\operatorname{dim} \mathrm{V}-\mathrm{m}_{\mathrm{i}}$ (see Theorem 3.9). But from Theorem 5.6 we then see that

$$
\begin{aligned}
\operatorname{dim} V_{\lambda_{i}} & =\operatorname{dim} \operatorname{Ker}\left(t-\lambda_{i} 1\right)=\operatorname{nul}\left(T-\lambda_{i} 1\right)=\operatorname{dim} V-\mathrm{r}\left(T-\lambda_{i} 1\right) \\
& =m_{i} .
\end{aligned}
$$

In other words, if T is diagonalizable, then the dimension of each eigenspace $\mathrm{V}_{\lambda_{\mathrm{i}}}$ is just the multiplicity of the eigenvalue $\lambda_{\mathrm{i}}$. Let us clarify this in terms of some common terminology. In so doing, we will also repeat this conclusion from a slightly different viewpoint.

Given a linear operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$, what we have called the multiplicity of an eigenvalue $\lambda$ is the largest positive integer $m$ such that $(x-\lambda)^{m}$ divides the characteristic polynomial $\Delta_{T}(x)$. This is properly called the algebraic multiplicity of $\lambda$, in contrast to the geometric multiplicity which is the number of linearly independent eigenvectors belonging to that eigenvalue. In other words, the geometric multiplicity of $\lambda$ is the dimension of $V_{\lambda}$. In general, we will use the word "multiplicity" to mean the algebraic multiplicity. The set of
all eigenvalues of a linear operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is called the spectrum of T . If some eigenvalue in the spectrum of T is of algebraic multiplicity $>1$, then the spectrum is said to be degenerate.

If $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ has an eigenvalue $\lambda$ of algebraic multiplicity m , then it is not hard for us to show that the dimension of the eigenspace $V_{\lambda}$ must be less than or equal to $m$. Note that since every element of $V_{\lambda}$ is an eigenvector of $T$ with eigenvalue $\lambda$, the space $V_{\lambda}$ must be a $T$-invariant subspace of V . Furthermore, every basis for $\mathrm{V}_{\lambda}$ will obviously consist of eigenvectors corresponding to $\lambda$.

Theorem 7.25 Let $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ have eigenvalue $\lambda$. Then the geometric multiplicity of $\lambda$ is always less than or equal to its algebraic multiplicity. In other words, if $\lambda$ has algebraic multiplicity m , then $\operatorname{dim} \mathrm{V}_{\lambda} \leq \mathrm{m}$.

Proof Suppose $\operatorname{dim} \mathrm{V}_{\lambda}=\mathrm{r}$ and let $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{r}}\right\}$ be a basis for $\mathrm{V}_{\lambda}$. By Theorem 2.10, we extend this to a basis $\left\{v_{1}, \ldots, v_{n}\right\}$ for V. Relative to this basis, T must have the matrix representation (see Theorem 7.19)

$$
\left(\begin{array}{cc}
\lambda I_{r} & C \\
0 & D
\end{array}\right)
$$

Applying Theorem 4.14 and the fact that the determinant of a diagonal matrix is just the product of its (diagonal) elements, we see that the characteristic polynomial $\Delta_{T}(x)$ of $T$ is given by

$$
\begin{aligned}
\Delta_{T}(x) & =\left|\begin{array}{cc}
(x-\lambda) I_{r} & -C \\
0 & x I_{n-r}-D
\end{array}\right| \\
& =\operatorname{det}\left[(x-\lambda) I_{r}\right] \operatorname{det}\left(x I_{n-r}-D\right) \\
& =(x-\lambda)^{r} \operatorname{det}\left(x I_{n-r}-D\right)
\end{aligned}
$$

which shows that $(x-\lambda)^{r}$ divides $\Delta_{T}(x)$. Since by definition $m$ is the largest positive integer such that $(x-\lambda)^{m} \mid \Delta_{T}(x)$, it follows that $r \leq m$.

Note that a special case of this theorem arises when an eigenvalue is of (algebraic) multiplicity 1 . In this case, it then follows that the geometric and algebraic multiplicities are necessarily equal. We now proceed to show just when this will be true in general. Recall that any polynomial over an algebraically closed field will factor into linear terms (Theorem 6.13).

Theorem 7.26 Assume that $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ has a characteristic polynomial that factors into (not necessarily distinct) linear terms. Let T have distinct eigenvalues $\lambda_{1}, \ldots, \lambda_{r}$ with (algebraic) multiplicities $\mathrm{m}_{1}, \ldots, \mathrm{~m}_{\mathrm{r}}$ respectively, and let $\operatorname{dim} \mathrm{V}_{\lambda_{\mathrm{i}}}=\mathrm{d}_{\mathrm{i}}$. Then T is diagonalizable if and only if $\mathrm{m}_{\mathrm{i}}=\mathrm{d}_{\mathrm{i}}$ for each $\mathrm{i}=1, \ldots$, r .

Proof Let $\operatorname{dim} \mathrm{V}=\mathrm{n}$. We note that since the characteristic polynomial of T is of degree n and factors into linear terms, it follows that $\mathrm{m}_{1}+\cdots+\mathrm{m}_{\mathrm{r}}=\mathrm{n}$. We first assume that T is diagonalizable. By definition, this means that V has a basis consisting of $n$ linearly independent eigenvectors of T. Since each of these basis eigenvectors must belong to at least one of the eigenspaces $\mathrm{V}_{\lambda_{\mathrm{i}}}$, it follows that $\mathrm{V}=\mathrm{V}_{\lambda_{1}}+\cdots+\mathrm{V}_{\lambda_{\mathrm{r}}}$ and consequently $\mathrm{n} \leq \mathrm{d}_{1}+\cdots+\mathrm{d}_{\mathrm{r}}$. From Theorem 7.25 we know that $\mathrm{d}_{\mathrm{i}} \leq \mathrm{m}_{\mathrm{i}}$ for each $\mathrm{i}=1, \ldots, \mathrm{r}$ and hence

$$
\mathrm{n} \leq \mathrm{d}_{1}+\cdots+\mathrm{d}_{\mathrm{r}} \leq \mathrm{m}_{1}+\cdots+\mathrm{m}_{\mathrm{r}}=\mathrm{n}
$$

which implies that $d_{1}+\cdots+d_{r}=m_{1}+\cdots+m_{r}$ or

$$
\left(\mathrm{m}_{1}-\mathrm{d}_{1}\right)+\cdots+\left(\mathrm{m}_{\mathrm{r}}-\mathrm{d}_{\mathrm{r}}\right)=0 .
$$

But each term in this equation is nonnegative (by Theorem 7.25), and hence we must have $m_{i}=d_{i}$ for each $i$.

Conversely, suppose that $d_{i}=m_{i}$ for each $i=1, \ldots, r$. For each $i$, we know that any basis for $\mathrm{V}_{\lambda_{\mathrm{i}}}$ consists of linearly independent eigenvectors corresponding to the eigenvalue $\lambda_{\mathrm{i}}$, while by Theorem 7.8 , we know that eigenvectors corresponding to distinct eigenvalues are linearly independent. Therefore the union $\mathcal{B}$ of the bases of $\left\{\mathrm{V}_{\lambda_{\mathrm{i}}}\right\}$ forms a linearly independent set of $\mathrm{d}_{1}+\cdots+\mathrm{d}_{\mathrm{r}}=\mathrm{m}_{1}+\cdots+\mathrm{m}_{\mathrm{r}}$ vectors. But $\mathrm{m}_{1}+\cdots+\mathrm{m}_{\mathrm{r}}=\mathrm{n}=\operatorname{dim} \mathrm{V}$, and hence $\mathcal{B}$ forms a basis for V . Since this shows that V has a basis of eigenvectors of T , it follows by definition that T must be diagonalizable.

The following corollary is a repeat of Corollary 2 of Theorem 7.24. Its (very easy) proof may be based entirely on the material of this section.

Corollary 1 An operator $T \in L(V)$ is diagonalizable if and only if

$$
\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}
$$

where $\mathrm{W}_{1}, \ldots, \mathrm{~W}_{\mathrm{r}}$ are the eigenspaces corresponding to the distinct eigenvalues of T.

## Proof This is Exercise 7.7.1.

Using Theorem 5.6, we see that the geometric multiplicity of an eigenvalue $\lambda$ is given by

$$
\operatorname{dim} V_{\lambda}=\operatorname{dim}(\operatorname{Ker}(T-\lambda 1))=\operatorname{nul}(T-\lambda 1)=\operatorname{dim} V-r(T-\lambda 1) .
$$

This observation together with Theorem 7.26 proves the next corollary.
Corollary 2 An operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ whose characteristic polynomial factors into linear terms is diagonalizable if and only if the algebraic multiplicity of $\lambda$ is equal to $\operatorname{dim} \mathrm{V}-\mathrm{r}(\mathrm{T}-\lambda 1)$ for each eigenvalue $\lambda$.

Example 7.10 Consider the operator $\mathrm{T} \in \mathrm{L}\left(\mathbb{R}^{3}\right)$ defined by

$$
T(x, y, z)=(9 x+y, 9 y, 7 z)
$$

Relative to the standard basis for $\mathbb{R}^{3}$, the matrix representation of T is given by

$$
A=\left(\begin{array}{lll}
9 & 1 & 0 \\
0 & 9 & 0 \\
0 & 0 & 7
\end{array}\right)
$$

and hence the characteristic polynomial is

$$
\Delta_{\mathrm{A}}(\mathrm{x})=\operatorname{det}(\mathrm{A}-\lambda \mathrm{I})=(9-\lambda)^{2}(7-\lambda)
$$

which is a product of linear factors. However,

$$
A-9 I=\left(\begin{array}{rrr}
0 & 1 & 0 \\
0 & 0 & 0 \\
0 & 0 & -2
\end{array}\right)
$$

which clearly has rank 2 , and hence $\operatorname{nul}(\mathrm{T}-9)=3-2=1$ which is not the same as the algebraic multiplicity of $\lambda=9$. Thus T is not diagonalizable. //

Example 7.11 Consider the operator on $\mathbb{R}^{3}$ defined by the following matrix:

$$
A=\left(\begin{array}{rrr}
5 & -6 & -6 \\
-1 & 4 & 2 \\
3 & -6 & -4
\end{array}\right)
$$

In order to avoid factoring a cubic polynomial, we compute the characteristic polynomial $\Delta_{\mathrm{A}}(\mathrm{x})=\operatorname{det}(\mathrm{xI}-\mathrm{A})$ by applying Theorem 4.4 as follows (the reader should be able to see exactly what elementary row operations were performed in each step).

$$
\begin{aligned}
\left|\begin{array}{ccc}
x-5 & 6 & 6 \\
1 & x-4 & -2 \\
-3 & 6 & x+4
\end{array}\right| & =\left|\begin{array}{ccc}
x-2 & 0 & -x+2 \\
1 & x-4 & -2 \\
-3 & 6 & x+4
\end{array}\right| \\
& =(x-2)\left|\begin{array}{ccc}
1 & 0 & -1 \\
1 & x-4 & -2 \\
-3 & 6 & x+4
\end{array}\right| \\
& =(x-2)\left|\begin{array}{ccc}
1 & 0 & -1 \\
0 & x-4 & -1 \\
0 & 6 & x+1
\end{array}\right| \\
& =(x-2)\left|\begin{array}{cc}
x-4 & -1 \\
6 & x+1
\end{array}\right| \\
& =(x-2)^{2}(x-1) .
\end{aligned}
$$

We now see that A has eigenvalue $\lambda_{1}=1$ with (algebraic) multiplicity 1 , and eigenvalue $\lambda_{2}=2$ with (algebraic) multiplicity 2 . From Theorem 7.25 we know that the algebraic and geometric multiplicities of $\lambda_{1}$ are necessarily the same and equal to 1 , so we need only consider $\lambda_{2}$. Observing that

$$
A-2 I=\left(\begin{array}{rrr}
3 & -6 & -6 \\
-1 & 2 & 2 \\
3 & -6 & -6
\end{array}\right)
$$

it is obvious that $\mathrm{r}(\mathrm{A}-2 \mathrm{I})=1$, and hence $\operatorname{nul}(\mathrm{A}-2 \mathrm{I})=3-1=2$. This shows that A is indeed diagonalizable.

Let us now construct bases for the eigenspaces $W_{i}=V_{\lambda_{i}}$. This means that we seek vectors $\mathrm{v}=(\mathrm{x}, \mathrm{y}, \mathrm{z}) \in \mathbb{R}^{3}$ such that $\left(\mathrm{A}-\lambda_{\mathrm{i}} \mathrm{I}\right) \mathrm{v}=0$. This is easily solved by the usual row reduction techniques as follows. For $\lambda_{1}=1$ we have

$$
A-I=\left(\begin{array}{rrr}
4 & -6 & -6 \\
-1 & 3 & 2 \\
3 & -6 & -5
\end{array}\right) \rightarrow\left(\begin{array}{rrr}
1 & 0 & -1 \\
-1 & 3 & 2 \\
3 & -6 & -5
\end{array}\right) \rightarrow\left(\begin{array}{rrr}
1 & 0 & -1 \\
0 & 3 & 1 \\
0 & -6 & -2
\end{array}\right) \rightarrow\left(\begin{array}{rrr}
1 & 0 & -1 \\
0 & 3 & 1 \\
0 & 0 & 0
\end{array}\right)
$$

which has the solutions $x=z$ and $y=-z / 3=-x / 3$. Therefore $W_{1}$ is spanned by the single eigenvector $\mathrm{v}_{1}=(3,-1,3)$. As to $\lambda_{2}=2$, we proceed in a similar manner to obtain

$$
A-2 I=\left(\begin{array}{rrr}
3 & -6 & -6 \\
-1 & 2 & 2 \\
3 & -6 & -6
\end{array}\right) \rightarrow\left(\begin{array}{rrr}
1 & -2 & -2 \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right)
$$

which implies that any vector ( $x, y, z$ ) with $x=2 y+2 z$ will work. For example, we can let $x=0$ and $y=1$ to obtain $z=-1$, and hence one basis vector for $W_{2}$ is given by $v_{2}=(0,1,-1)$. If we let $x=1$ and $y=0$, then we have $z=1 / 2$ so that another independent basis vector for $W_{2}$ is given by $v_{3}=(2,0,1)$. In terms of these eigenvectors, the transformation matrix P that diagonalizes A is given by

$$
P=\left(\begin{array}{rrr}
3 & 0 & 2 \\
-1 & 1 & 0 \\
3 & -1 & 1
\end{array}\right)
$$

and we leave it to the reader to verify that $\mathrm{AP}=\mathrm{PD}$ (i.e., $\mathrm{P}^{-1} \mathrm{AP}=\mathrm{D}$ ) where D is the diagonal matrix with diagonal elements $d_{11}=1$ and $d_{22}=d_{33}=2$.

Finally, we note that since A is diagonalizable, Theorems 7.12 and 7.24 show that the minimal polynomial for A must be $(\mathrm{x}-1)(\mathrm{x}-2)$. //

## Exercises

1. Prove Corollary 1 of Theorem 7.26.
2. Show that two similar matrices $A$ and $B$ have the same eigenvalues, and these eigenvalues have the same geometric multiplicities.
3. Let $\lambda_{1}, \ldots, \lambda_{\mathrm{r}} \in \mathcal{F}$ be distinct, and let $\mathrm{D} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ be diagonal with a characteristic polynomial of the form

$$
\Delta_{D}(x)=\left(x-\lambda_{1}\right)^{d_{1}} \cdots\left(x-\lambda_{r}\right)^{d_{r}} .
$$

Let V be the space of all $\mathrm{n} x \mathrm{n}$ matrices B that commute with D , i.e., the set of all B such that $B D=D B$. Prove that $\operatorname{dim} V=d_{1}^{2}+\cdots+d_{r}{ }^{2}$.
4. Relative to the standard basis, let $\mathrm{T} \in \mathrm{L}\left(\mathbb{R}^{4}\right)$ be represented by

$$
A=\left(\begin{array}{llll}
0 & 0 & 0 & 0 \\
a & 0 & 0 & 0 \\
0 & b & 0 & 0 \\
0 & 0 & c & 0
\end{array}\right) .
$$

Find conditions on $\mathrm{a}, \mathrm{b}$ and c such that T is diagonalizable.
5. Determine whether or not each of the following matrices is diagonalizable. If it is, find a nonsingular matrix P and a diagonal matrix D such that $\mathrm{P}^{-1} \mathrm{AP}=\mathrm{D}$.
(a) $\left(\begin{array}{lll}1 & 1 & 0 \\ 0 & 2 & 2 \\ 0 & 0 & 3\end{array}\right)$
(b) $\left(\begin{array}{rrr}3 & -1 & -2 \\ 2 & 0 & -2 \\ 2 & -1 & -1\end{array}\right)$
(c) $\left(\begin{array}{rrr}-1 & 1 & 0 \\ 0 & 5 & 0 \\ 4 & -2 & 5\end{array}\right)$
(d) $\left(\begin{array}{rrr}-1 & -3 & -9 \\ 0 & 5 & 18 \\ 0 & -2 & -7\end{array}\right)$
(e) $\left(\begin{array}{lll}7 & -4 & 0 \\ 8 & -5 & 0 \\ 6 & -6 & 3\end{array}\right)$
(f) $\left(\begin{array}{rrr}0 & 0 & 1 \\ 1 & 0 & -1 \\ 0 & 1 & 1\end{array}\right)$
(g) $\left(\begin{array}{rrr}3 & 1 & 1 \\ 2 & 4 & 2 \\ -1 & -1 & 1\end{array}\right)$
6. Determine whether or not each of the following operators $T \in L\left(\mathbb{R}^{3}\right)$ is diagonalizable. If it is, find an eigenvector basis for $\mathbb{R}^{3}$ such that $[T]$ is diagonal.
(a) $T(x, y, z)=(-y, x, 3 z)$.
(b) $T(x, y, z)=(8 x+2 y-2 z, 3 x+3 y-z, 24 x+8 y-6 z)$.
(c) $T(x, y, z)=(4 x+z, 2 x+3 y+2 z, x+4 z)$.
(d) $T(x, y, z)=(-2 y-3 z, x+3 y+3 z, z)$.
7. Suppose a matrix $A$ is diagonalizable. Prove that $A^{m}$ is diagonalizable for any positive integer m .
8. Summarize several of our results by proving the following theorem:

Let V be finite-dimensional, suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ has distinct eigenvalues $\lambda_{1}, \ldots, \lambda_{\mathrm{r}}$, and let $\mathrm{W}_{\mathrm{i}}=\operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)$. Then the following are equivalent:
(a) T is diagonalizable.
(b) $\Delta_{T}(x)=\left(x-\lambda_{1}\right)^{m_{1}} \cdots\left(x-\lambda_{r}\right)^{m_{r}}$ and $W_{i}$ is of dimension $m_{i}$ for each $\mathrm{i}=1, \ldots, \mathrm{r}$.
(c) $\operatorname{dim} \mathrm{W}_{1}+\cdots+\operatorname{dim} \mathrm{W}_{\mathrm{r}}=\operatorname{dim} \mathrm{V}$.
9. Let $V_{3}$ be the space of real polynomials of degree at most 3, and let $f^{\prime}$ and $\mathrm{f}^{\prime \prime}$ denote the first and second derivatives of $\mathrm{f} \in \mathrm{V}$. Define $T \in L\left(\mathrm{~V}_{3}\right)$ by $T(f)=f^{\prime}+f^{\prime \prime}$. Decide whether or not $T$ is diagonalizable, and if it is, find a basis for $\mathrm{V}_{3}$ such that [ T ] is diagonal.
10. (a) Let $\mathrm{V}_{2}$ be the space of real polynomials of degree at most 2, and define $T \in L\left(V_{2}\right)$ by $T\left(a^{2}+b x+c\right)=c x^{2}+b x+a$. Decide whether or not T is diagonalizable, and if it is, find a basis for $\mathrm{V}_{2}$ such that [ T ] is diagonal.
(b) Repeat part (a) with $\mathrm{T}=(\mathrm{x}+1)(\mathrm{d} / \mathrm{dx})$. (See Exercise 7.3.17.)

### 7.8 PROJECTIONS

In this section we introduce the concept of projection operators and show how they may be related to direct sum decompositions where each of the subspaces in the direct sum is invariant under some linear operator.

Suppose that U and W are subspaces of a vector space V with the property that $\mathrm{V}=\mathrm{U} \oplus \mathrm{W}$. Then every $\mathrm{v} \in \mathrm{V}$ has a unique representation of the form v $=\mathrm{u}+\mathrm{w}$ where $\mathrm{u} \in \mathrm{U}$ and $\mathrm{w} \in \mathrm{W}$ (Theorem 2.14). We now define the mapping $\mathrm{E}: \mathrm{V} \rightarrow \mathrm{V}$ by $\mathrm{Ev}=\mathrm{u}$. Note that E is well-defined since the direct sum decomposition is unique. Moreover, given any other $\mathrm{v}^{\prime} \in \mathrm{V}=\mathrm{U} \oplus \mathrm{W}$ with $v^{\prime}=u^{\prime}+w^{\prime}$, we know that $v+v^{\prime}=\left(u+u^{\prime}\right)+\left(w+w^{\prime}\right)$ and $k v=k u+k w$, and hence it is easy to see that $E$ is in fact linear because

$$
\mathrm{E}\left(\mathrm{v}+\mathrm{v}^{\prime}\right)=\mathrm{u}+\mathrm{u}^{\prime}=\mathrm{Ev}+\mathrm{Ev}^{\prime}
$$

and

$$
\mathrm{E}(\mathrm{kv})=\mathrm{ku}=\mathrm{k}(\mathrm{Ev})
$$

The linear operator $E \in L(V)$ is called the projection of $V$ on $U$ in the direction of W . Furthermore, since any $\mathrm{u} \in \mathrm{U} \subset \mathrm{V}$ may be written in the form $\mathrm{u}=$ $\mathrm{u}+0$, we also see that $\mathrm{Eu}=\mathrm{u}$ and therefore

$$
\mathrm{E}^{2} \mathrm{v}=\mathrm{E}(\mathrm{Ev})=\mathrm{Eu}=\mathrm{u}=\mathrm{Ev}
$$

In other words, a projection operator $E$ has the property that $E^{2}=E$. By way of terminology, any operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ with the property that $\mathrm{T}^{2}=\mathrm{T}$ is said to be idempotent.

On the other hand, given a vector space V , suppose we have an operator $\mathrm{E} \in \mathrm{L}(\mathrm{V})$ with the property that $\mathrm{E}^{2}=\mathrm{E}$. We claim that $\mathrm{V}=\operatorname{Im} \mathrm{E} \oplus$ Ker E . Indeed, first note that if $u \in \operatorname{Im} E$, then by definition this means there exists $v \in V$ with the property that $E v=u$. It therefore follows that

$$
\mathrm{Eu}=\mathrm{E}(\mathrm{Ev})=\mathrm{E}^{2} \mathrm{v}=\mathrm{Ev}=\mathrm{u}
$$

and thus $\mathrm{Eu}=\mathrm{u}$ for any $\mathrm{u} \in \operatorname{Im} \mathrm{E}$. Conversely, the equation $\mathrm{Eu}=\mathrm{u}$ obviously says that $u \in \operatorname{Im} E$, and hence we see that $u \in \operatorname{Im} E$ if and only if $E u=u$.

Next, note that given any $\mathrm{v} \in \mathrm{V}$ we may clearly write

$$
v=E v+v-E v=E v+(1-E) v
$$

where by definition, $\mathrm{Ev} \in \operatorname{Im} \mathrm{E}$. Since

$$
E[(1-E) v]=\left(E-E^{2}\right) v=(E-E) v=0
$$

we see that $(1-E) v \in \operatorname{Ker} E$, and hence $V=\operatorname{Im} E+\operatorname{Ker} E$. We claim that this sum is in fact direct. To see this, let $w \in \operatorname{Im} E \cap$ Ker $E$. Since $w \in \operatorname{Im} E$ and $\mathrm{E}^{2}=\mathrm{E}$, we have seen that $\mathrm{Ew}=\mathrm{w}$, while the fact that $\mathrm{w} \in \operatorname{Ker} \mathrm{E}$ means that $\mathrm{Ew}=0$. Therefore $\mathrm{w}=0$ so that $\operatorname{Im} \mathrm{E} \cap \operatorname{Ker} \mathrm{E}=\{0\}$, and hence

$$
\mathrm{V}=\operatorname{Im} \mathrm{E} \oplus \operatorname{Ker} \mathrm{E}
$$

Since we have now shown that any $v \in V$ may be written in the unique form $v=u+w$ with $u \in \operatorname{Im} E$ and $w \in \operatorname{Ker} E$, it follows that $E v=E u+E w=u+0$ $=u$ so that $E$ is the projection of $V$ on $\operatorname{Im} E$ in the direction of Ker $E$.

It is also of use to note that

$$
\operatorname{Ker} \mathrm{E}=\operatorname{Im}(1-\mathrm{E})
$$

and

$$
\operatorname{Ker}(1-\mathrm{E})=\operatorname{Im} \mathrm{E} .
$$

To see this, suppose $w \in$ Ker E. Then

$$
\mathrm{w}=\mathrm{Ew}+(1-\mathrm{E}) \mathrm{w}=(1-\mathrm{E}) \mathrm{w}
$$

which implies that $w \in \operatorname{Im}(1-E)$, and hence $\operatorname{Ker} E \subset \operatorname{Im}(1-E)$. On the other hand, if $w \in \operatorname{Im}(1-E)$ then there exists $w^{\prime} \in V$ such that $w=(1-E) w^{\prime}$ and hence

$$
E w=\left(E-E^{2}\right) w^{\prime}=(E-E) w^{\prime}=0
$$

so that $\mathrm{w} \in$ Ker E . This shows that $\operatorname{Im}(1-\mathrm{E}) \subset$ Ker E , and therefore Ker $\mathrm{E}=$ $\operatorname{Im}(1-E)$. The similar proof that $\operatorname{Ker}(1-E)=\operatorname{Im} E$ is left as an exercise for the reader (Exercise 7.8.1).

Theorem 7.27 Let $V$ be a vector space with $\operatorname{dim} V=n$, and suppose $E \in$ $\mathrm{L}(\mathrm{V})$ has rank $\mathrm{k}=\operatorname{dim}(\operatorname{Im} \mathrm{E})$. Then E is idempotent (i.e., $\mathrm{E}^{2}=\mathrm{E}$ ) if and only if any one of the following statements is true:
(a) If $v \in \operatorname{Im} E$, then $E v=v$.
(b) $V=\operatorname{Im} E \oplus \operatorname{Ker} E$ and $E$ is the projection of $V$ on $\operatorname{Im} E$ in the direction of Ker E.
(c) $\operatorname{Im} \mathrm{E}=\operatorname{Ker}(1-\mathrm{E})$ and $\operatorname{Ker} \mathrm{E}=\operatorname{Im}(1-\mathrm{E})$.
(d) It is possible to choose a basis for $V$ such that $[E]=I_{k} \oplus 0_{n-k}$.

Proof Suppose $\mathrm{E}^{2}=\mathrm{E}$. In view of the above discussion, all that remains is to prove part (d). Applying part (b), we let $\left\{\mathrm{e}_{1}, \ldots, \mathrm{e}_{\mathrm{k}}\right\}$ be a basis for $\operatorname{Im} E$ and $\left\{\mathrm{e}_{\mathrm{k}+1}, \ldots, \mathrm{e}_{\mathrm{n}}\right\}$ be a basis for Ker E. By part (a), we know that $\mathrm{Ee}_{\mathrm{i}}=\mathrm{e}_{\mathrm{i}}$ for $\mathrm{i}=$ $1, \ldots, \mathrm{k}$, and by definition of Ker E, we have $\mathrm{Ee}_{\mathrm{i}}=0$ for $\mathrm{i}=\mathrm{k}+1, \ldots$, n. But then $[E]$ has the desired form since the ith column of $[E]$ is just $\mathrm{Ee}_{\mathrm{i}}$.

Conversely, suppose (a) is true and $v \in V$ is arbitrary. Then $E^{2} v=E(E v)=$ $E v$ implies that $E^{2}=E$. Now suppose that $(b)$ is true and $v \in V$. Then $v=u+$ w where $\mathrm{u} \in \operatorname{Im} \mathrm{E}$ and $\mathrm{w} \in \operatorname{Ker} \mathrm{E}$. Therefore $\mathrm{Ev}=\mathrm{Eu}+\mathrm{Ew}=\mathrm{Eu}=\mathrm{u}$ (by definition of projection) and $E^{2} v=E^{2} u=E u=u$ so that $E^{2} v=E v$ for all $v \in V$, and hence $\mathrm{E}^{2}=\mathrm{E}$. If (c) holds and $\mathrm{v} \in \mathrm{V}$, then $\mathrm{Ev} \in \operatorname{Im} \mathrm{E}=\operatorname{Ker}(1-\mathrm{E})$ so that $0=(1-E) E v=E v-E^{2} v$ and hence $E^{2} v=E v$ again. Similarly, $(1-E) v \in$ $\operatorname{Im}(1-E)=$ Ker $E$ so that $0=E(1-E) v=E v-E^{2} v$ and hence $E^{2} v=E v$. In either case, we have $E^{2}=E$. Finally, from the form of $[E]$ given in (d), it is obvious that $\mathrm{E}^{2}=\mathrm{E}$.

It is also worth making the following observation. If we are given a vector space V and a subspace $\mathrm{W} \subset \mathrm{V}$, then there may be many subspaces $\mathrm{U} \subset \mathrm{V}$ with the property that $V=U \oplus W$. For example, the space $\mathbb{R}^{3}$ is not necessarily represented by the usual orthogonal Cartesian coordinate system. Rather, it may be viewed as consisting of a line plus any (oblique) plane not containing the given line. However, in the particular case that $\mathrm{V}=\mathrm{W} \oplus \mathrm{W}^{\perp}$, then $\mathrm{W}^{\perp}$ is uniquely specified by W (see Section 2.5 ). In this case, the projection $\mathrm{E} \in$ $\mathrm{L}(\mathrm{V})$ defined by $\mathrm{Ev}=\mathrm{w}$ with $\mathrm{w} \in \mathrm{W}$ is called the orthogonal projection of V
on W. In other words, E is an orthogonal projection if $(\operatorname{Im} \mathrm{E})^{\perp}=\operatorname{Ker} \mathrm{E} . \mathrm{By}$ the corollary to Theorem 2.22, this is equivalent to the requirement that $(\text { Ker } \mathrm{E})^{\perp}=\operatorname{Im} \mathrm{E}$.

It is not hard to generalize these results to the direct sum of more than two subspaces. Indeed, suppose that we have a vector space V such that $\mathrm{V}=\mathrm{W}_{1} \oplus$ $\cdots \oplus \mathrm{W}_{\mathrm{r}}$. Since any $\mathrm{v} \in \mathrm{V}$ has the unique representation as $\mathrm{v}=\mathrm{w}_{1}+\cdots+\mathrm{w}_{\mathrm{r}}$ with $W_{i} \in W_{i}$, we may define for each $j=1, \ldots, r$ the operator $E_{j} \in L(V)$ by $E_{j} v=w_{j}$. That each $E_{j}$ is in fact linear is easily shown exactly as above for the simpler case. It should also be clear that $\operatorname{Im} \mathrm{E}_{\mathrm{j}}=\mathrm{W}_{\mathrm{j}}$ (see Exercise 7.8.2). If we write

$$
\mathrm{w}_{\mathrm{j}}=0+\cdots+0+\mathrm{w}_{\mathrm{j}}+0+\cdots+0
$$

as the unique expression for $\mathrm{w}_{\mathrm{j}} \in \mathrm{W}_{\mathrm{j}} \subset \mathrm{V}$, then we see that $\mathrm{E}_{\mathrm{j}} \mathrm{w}_{\mathrm{j}}=\mathrm{w}_{\mathrm{j}}$, and hence for any $\mathrm{v} \in \mathrm{V}$ we have

$$
\mathrm{E}_{\mathrm{j}}^{2} \mathrm{v}=\mathrm{E}_{\mathrm{j}}\left(\mathrm{E}_{\mathrm{j}} \mathrm{v}\right)=\mathrm{E}_{\mathrm{j}} \mathrm{w}_{\mathrm{j}}=\mathrm{w}_{\mathrm{j}}=\mathrm{E}_{\mathrm{j}} \mathrm{v}
$$

so that $E_{j}^{2}=E_{j}$.
The representation of each $w_{j}$ as $E_{j} v$ is very useful because we may write any $\mathrm{v} \in \mathrm{V}$ as

$$
\mathrm{v}=\mathrm{w}_{1}+\cdots+\mathrm{w}_{\mathrm{r}}=\mathrm{E}_{1} \mathrm{v}+\cdots+\mathrm{E}_{\mathrm{r}} \mathrm{v}=\left(\mathrm{E}_{1}+\cdots+\mathrm{E}_{\mathrm{r}}\right) \mathrm{v}
$$

and thus we see that $E_{1}+\cdots+E_{r}=1$. Furthermore, since the image of $E_{j}$ is $W_{j}$, it follows that if $\mathrm{E}_{\mathrm{j}} \mathrm{v}=0$ then $\mathrm{w}_{\mathrm{j}}=0$, and hence

$$
\text { Ker } \mathrm{E}_{\mathrm{j}}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{j}-1} \oplus \mathrm{~W}_{\mathrm{j}+1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}
$$

We then see that for any $j=1, \ldots, r$ we have $V=\operatorname{Im} E_{j} \oplus \operatorname{Ker} E_{j}$ exactly as before. It is also easy to see that $\mathrm{E}_{\mathrm{i}} \mathrm{E}_{\mathrm{j}}=0$ if $\mathrm{i} \neq \mathrm{j}$ because $\operatorname{Im} \mathrm{E}_{\mathrm{j}}=\mathrm{W}_{\mathrm{j}} \subset \operatorname{Ker} \mathrm{E}_{\mathrm{i}}$.

Theorem 7.28 Let V be a vector space, and suppose that $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus$ $W_{r}$. Then for each $j=1, \ldots, r$ there exists a linear operator $E_{j} \in L(V)$ with the following properties:
(a) $1=\mathrm{E}_{1}+\cdots+\mathrm{E}_{\mathrm{r}}$.
(b) $\mathrm{E}_{\mathrm{i}} \mathrm{E}_{\mathrm{j}}=0$ if $\mathrm{i} \neq \mathrm{j}$.
(c) $\mathrm{E}_{\mathrm{j}}^{2}=\mathrm{E}_{\mathrm{j}}$.
(d) $\operatorname{Im} E_{j}=W_{j}$.

Conversely, if $\left\{\mathrm{E}_{1}, \ldots, \mathrm{E}_{\mathrm{r}}\right\}$ are linear operators on V that obey properties (a) and (b), then each $\mathrm{E}_{\mathrm{j}}$ is idempotent and $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ where $\mathrm{W}_{\mathrm{j}}=\operatorname{Im} \mathrm{E}_{\mathrm{j}}$.

Proof In view of the previous discussion, we only need to prove the converse statement. From (a) and (b) we see that

$$
E_{j}=E_{j} 1=E_{j}\left(E_{1}+\cdots+E_{r}\right)=E_{j}^{2}+\Sigma_{i \neq j} E_{j} E_{i}=E_{j}^{2}
$$

which shows that each $E_{j}$ is idempotent. Next, property (a) shows us that for any $\mathrm{v} \in \mathrm{V}$ we have

$$
\mathrm{v}=1 \mathrm{v}=\mathrm{E}_{1} \mathrm{v}+\cdots+\mathrm{E}_{\mathrm{r}} \mathrm{v}
$$

and hence $\mathrm{V}=\mathrm{W}_{1}+\cdots+\mathrm{W}_{\mathrm{r}}$ where we have defined $\mathrm{W}_{\mathrm{j}}=\operatorname{Im} \mathrm{E}_{\mathrm{j}}$. Now suppose that $0=\mathrm{w}_{1}+\cdots+\mathrm{w}_{\mathrm{r}}$ where each $\mathrm{w}_{\mathrm{j}} \in \mathrm{W}_{\mathrm{j}}$. If we can show that this implies $\mathrm{w}_{1}=\cdots=\mathrm{w}_{\mathrm{r}}=0$, then any $\mathrm{v} \in \mathrm{V}$ will have a unique representation $\mathrm{v}=\mathrm{v}_{1}+\cdots+\mathrm{v}_{\mathrm{r}}$ with $\mathrm{v}_{\mathrm{i}} \in \mathrm{W}_{\mathrm{i}}$. This is because if

$$
\mathrm{v}=\mathrm{v}_{1}+\cdots+\mathrm{v}_{\mathrm{r}}=\mathrm{v}_{1}{ }^{\prime}+\cdots+\mathrm{v}_{\mathrm{r}}^{\prime}
$$

then

$$
\left(\mathrm{v}_{1}-\mathrm{v}_{1}^{\prime}\right)+\cdots+\left(\mathrm{v}_{\mathrm{r}}-\mathrm{v}_{\mathrm{r}}^{\prime}\right)=0
$$

would imply that $v_{i}-v_{i}{ }^{\prime}=0$ for each $i$, and thus $v_{i}=v_{i}{ }^{\prime}$. Hence it will follow that $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ (Theorem 2.14).

Since $w_{1}+\cdots+w_{r}=0$, it is obvious that $E_{j}\left(w_{1}+\cdots+w_{r}\right)=0$. However, note that $\mathrm{E}_{\mathrm{j}} \mathrm{w}_{\mathrm{i}}=0$ if $\mathrm{i} \neq \mathrm{j}$ (because $\mathrm{w}_{\mathrm{i}} \in \operatorname{Im} \mathrm{E}_{\mathrm{i}}$ and $\mathrm{E}_{\mathrm{j}} \mathrm{E}_{\mathrm{i}}=0$ ), while $\mathrm{E}_{\mathrm{j}} \mathrm{w}_{\mathrm{j}}=\mathrm{w}_{\mathrm{j}}$ (since $w_{j}=E_{j} w^{\prime}$ for some $w^{\prime} \in V$ and hence $E_{j} w_{j}=E_{j}^{2} w^{\prime}=E_{j} w^{\prime}=w_{j}$ ). This shows that $\mathrm{w}_{1}=\cdots=\mathrm{w}_{\mathrm{r}}=0$ as desired.

We now turn our attention to invariant direct sum decompositions, referring to Section 7.5 for notation. We saw in Corollary 1 of Theorem 7.26 that a diagonalizable operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ leads to a direct sum decomposition of V in terms of the eigenspaces of T. However, Theorem 7.28 shows us that such a decomposition should lead to a collection of projections on these eigenspaces. Our next theorem elaborates on this observation in detail. Before stating and proving this result however, let us take another look at a matrix that has been diagonalized. We observe that a diagonal matrix of the form

$$
A=\left(\begin{array}{cccccc}
\lambda_{1} I_{m_{1}} & 0 & 0 & \cdots & 0 & 0 \\
0 & \lambda_{2} I_{m_{2}} & 0 & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & \cdots & 0 & \lambda_{r} I_{m_{r}}
\end{array}\right)
$$

can also be written as

$$
A=\lambda_{1}\left(\begin{array}{cccc}
I_{m_{1}} & & & \\
& 0_{m_{2}} & & \\
& & \ddots & \\
& & & 0_{m_{r}}
\end{array}\right)+\lambda_{2}\left(\begin{array}{ccccc}
0_{m_{1}} & & & \\
& I_{m_{2}} & & \\
& & \ddots & \\
& & & 0_{m_{r}}
\end{array}\right)+\cdots+\lambda_{r}\left(\begin{array}{lllll}
0_{m_{1}} & & & \\
& 0_{m_{2}} & & \\
& & \ddots & \\
& & & I_{m_{r}}
\end{array}\right) .
$$

If we define $E_{i}$ to be the matrix obtained from $A$ by setting $\lambda_{i}=1$ and $\lambda_{j}=0$ for each $\mathrm{j} \neq \mathrm{i}$ (i.e., the i th matrix in the above expression), then this may be written in the simple form

$$
\mathrm{A}=\lambda_{1} \mathrm{E}_{1}+\lambda_{2} \mathrm{E}_{2}+\cdots+\lambda_{\mathrm{r}} \mathrm{E}_{\mathrm{r}}
$$

where clearly

$$
I=E_{1}+E_{2}+\cdots+E_{r}
$$

Furthermore, it is easy to see that the matrices $\mathrm{E}_{\mathrm{i}}$ have the property that

$$
\mathrm{E}_{\mathrm{i}} \mathrm{E}_{\mathrm{j}}=0 \text { if } \mathrm{i} \neq \mathrm{j}
$$

and

$$
\mathrm{E}_{\mathrm{i}}^{2}=\mathrm{E}_{\mathrm{i}} \neq 0
$$

With these observations in mind, we now prove this result in general.

Theorem 7.29 If $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is a diagonalizable operator with distinct eigenvalues $\lambda_{1}, \ldots, \lambda_{r}$, then there exist linear operators $E_{1}, \ldots, E_{r}$ in $L(V)$ such that:
(a) $1=\mathrm{E}_{1}+\cdots+\mathrm{E}_{\mathrm{r}}$.
(b) $\mathrm{E}_{\mathrm{i}} \mathrm{E}_{\mathrm{j}}=0$ if $\mathrm{i} \neq \mathrm{j}$.
(c) $\mathrm{T}=\lambda_{1} \mathrm{E}_{1}+\cdots+\lambda_{\mathrm{r}} \mathrm{E}_{\mathrm{r}}$.
(d) $\mathrm{E}_{\mathrm{j}}^{2}=\mathrm{E}_{\mathrm{j}}$.
(e) $\operatorname{Im} \mathrm{E}_{\mathrm{j}}=\mathrm{W}_{\mathrm{j}}$ where $\mathrm{W}_{\mathrm{j}}=\operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{j}} 1\right)$ is the eigenspace corresponding to $\lambda_{j}$.
Conversely, if there exist distinct scalars $\lambda_{1}, \ldots, \lambda_{r}$ and distinct nonzero linear operators $\mathrm{E}_{1}, \ldots, \mathrm{E}_{\mathrm{r}}$ satisfying properties (a), (b) and (c), then properties (d) and (e) are also satisfied, and T is diagonalizable with $\lambda_{1}, \ldots, \lambda_{\mathrm{r}}$ as its distinct eigenvalues.

Proof First assume T is diagonalizable with distinct eigenvalues $\lambda_{1}, \ldots, \lambda_{\mathrm{r}}$ and let $\mathrm{W}_{1}, \ldots, \mathrm{~W}_{\mathrm{r}}$ be the corresponding eigenspaces. By Corollary 1 of Theorem 7.26 we know that $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$. (Note that we do not base this result on Theorem 7.24, and hence the present theorem does not depend in any way on the primary decomposition theorem.) Then Theorem 7.28 shows the existence of the projection operators $\mathrm{E}_{1}, \ldots, \mathrm{E}_{\mathrm{r}}$ satisfying properties (a),
(b), (d) and (e). As to property (c), we see (by property (a)) that for any $\mathrm{v} \in \mathrm{V}$ we have $v=E_{1} v+\cdots+E_{r} v$. Since $E_{j} v \in W_{j}$, we know from the definition of eigenspace that $T\left(E_{j} v\right)=\lambda_{j}\left(E_{j} v\right)$, and therefore

$$
\begin{aligned}
T v & =T\left(E_{1} v\right)+\cdots+T\left(E_{r} v\right) \\
& =\lambda_{1}\left(E_{1} v\right)+\cdots+\lambda_{r}\left(E_{r} v\right) \\
& =\left(\lambda_{1} E_{1}+\cdots+\lambda_{r} E_{r}\right) v
\end{aligned}
$$

which verifies property (c).
Now suppose that we are given a linear operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ together with distinct scalars $\lambda_{1}, \ldots, \lambda_{r}$ and (nonzero) linear operators $E_{1}, \ldots, E_{r}$ that obey properties (a), (b) and (c). Multiplying (a) by $\mathrm{E}_{\mathrm{i}}$ and using (b) proves (d). Now multiply (c) from the right by $\mathrm{E}_{\mathrm{i}}$ and use property (b) to obtain $\mathrm{TE}_{\mathrm{i}}=\lambda_{\mathrm{i}} \mathrm{E}_{\mathrm{i}}$ or $\left(T-\lambda_{i} 1\right) E_{i}=0$. If $w_{i} \in \operatorname{Im} E_{i}$ is arbitrary, then $w_{i}=E_{i} w_{i}{ }^{\prime}$ for some $w_{i}{ }^{\prime} \in V$ and hence $\left(T-\lambda_{i} 1\right) w_{i}=\left(T-\lambda_{i} 1\right) \mathrm{E}_{\mathrm{i}} \mathrm{w}_{\mathrm{i}}{ }^{\prime}=0$ which shows that $\mathrm{w}_{\mathrm{i}} \in \operatorname{Ker}(\mathrm{T}-$ $\left.\lambda_{i} 1\right)$. Since $E_{i} \neq 0$, this shows the existence of a nonzero vector $w_{i} \in \operatorname{Ker}(T-$ $\lambda_{i} 1$ ) with the property that $\mathrm{Tw}_{\mathrm{i}}=\lambda_{\mathrm{i}} \mathrm{w}_{\mathrm{i}}$. This proves that each $\lambda_{\mathrm{i}}$ is an eigenvalue of T . We claim that there are no other eigenvalues of T other than $\left\{\lambda_{i}\right\}$. To see this, let $\alpha$ be any scalar and assume that $(T-\alpha 1) v=0$ for some nonzero $\mathrm{v} \in \mathrm{V}$. Using properties (a) and (c), we see that

$$
\mathrm{T}-\alpha 1=\left(\lambda_{1}-\alpha\right) \mathrm{E}_{1}+\cdots+\left(\lambda_{\mathrm{r}}-\alpha\right) \mathrm{E}_{\mathrm{r}}
$$

and hence letting both sides of this equation act on $v$ yields

$$
0=\left(\lambda_{1}-\alpha\right) \mathrm{E}_{1} \mathrm{v}+\cdots+\left(\lambda_{\mathrm{r}}-\alpha\right) \mathrm{E}_{\mathrm{r}} \mathrm{v}
$$

Multiplying this last equation from the left by $\mathrm{E}_{\mathrm{i}}$ and using properties (b) and (d), we then see that $\left(\lambda_{i}-\alpha\right) E_{i} v=0$ for every $i=1, \ldots, r$. Since $v \neq 0$ may be written as $v=E_{1} v+\cdots+E_{r} v$, it must be true that $E_{j} v \neq 0$ for some $j$, and hence in this case we have $\lambda_{j}-\alpha=0$ or $\alpha=\lambda_{j}$.

We must still show that $T$ is diagonalizable, and that $\operatorname{Im} \mathrm{E}_{\mathrm{i}}=\operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)$. It was shown in the previous paragraph that any nonzero $w_{j} \in \operatorname{Im} E_{j}$ satisfies $T w_{j}=\lambda_{j} w_{j}$, and hence any nonzero vector in the image of any $E_{i}$ is an eigenvector of $\mathrm{E}_{\mathrm{i}}$. Note this says that $\operatorname{Im} \mathrm{E}_{\mathrm{i}} \subset \operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)$. Using property (a), we see that any $\mathrm{w} \in \mathrm{V}$ may be written as $\mathrm{w}=\mathrm{E}_{1} \mathrm{w}+\cdots+\mathrm{E}_{\mathrm{r}} \mathrm{w}$ which shows that V is spanned by eigenvectors of T . But this is just what we mean when we say that $T$ is diagonalizable. Finally, suppose $\mathrm{w}_{\mathrm{i}} \in \operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)$ is arbitrary. Then ( $\mathrm{T}-\lambda_{\mathrm{i}} 1$ ) $\mathrm{w}_{\mathrm{i}}=0$ and hence (exactly as we showed above)

$$
0=\left(\lambda_{1}-\lambda_{\mathrm{i}}\right) \mathrm{E}_{1} \mathrm{~W}_{\mathrm{i}}+\cdots+\left(\lambda_{\mathrm{r}}-\lambda_{\mathrm{i}}\right) \mathrm{E}_{\mathrm{r}} \mathrm{~W}_{\mathrm{i}} .
$$

Thus for each $j=1, \ldots, r$ we have

$$
0=\left(\lambda_{j}-\lambda_{i}\right) \mathrm{E}_{j} \mathrm{w}_{\mathrm{i}}
$$

which implies $E_{j} w_{i}=0$ for $j \neq i$. Since $w_{i}=E_{1} w_{i}+\cdots+E_{r} w_{i}$ while $E_{j} w_{i}=0$ for $j \neq i$, we conclude that $w_{i}=E_{i} w_{i}$ which shows that $w_{i} \in \operatorname{Im} E_{i}$. In other words, we have also shown that $\operatorname{Ker}\left(T-\lambda_{i} 1\right) \subset \operatorname{Im} E_{i}$. Together with our earlier result, this proves that $\operatorname{Im} E_{i}=\operatorname{Ker}\left(T-\lambda_{i} 1\right)$.

## Exercises

1. (a) Let E be an idempotent linear operator. Show that $\operatorname{Ker}(1-\mathrm{E})=\operatorname{Im} \mathrm{E}$.
(b) If $\mathrm{E}^{2}=\mathrm{E}$, show that $(1-\mathrm{E})^{2}=1-\mathrm{E}$.
2. Let $\mathrm{V}=\mathrm{W}_{1} \oplus \cdots \oplus \mathrm{~W}_{\mathrm{r}}$ and suppose $\mathrm{v}=\mathrm{w}_{1}+\cdots+\mathrm{w}_{\mathrm{r}} \in \mathrm{V}$. For each $\mathrm{j}=$ $1, \ldots, r$ we define the operator $E_{j}$ on $V$ by $E_{j} v=w_{j}$.
(a) Show that $\mathrm{E}_{\mathrm{j}} \in \mathrm{L}(\mathrm{V})$.
(b) Show that $\operatorname{Im} \mathrm{E}_{\mathrm{j}}=\mathrm{W}_{\mathrm{j}}$.
3. Give a completely independent proof of Theorem 7.24 as follows:
(a) Let $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ be diagonalizable with decomposition $\mathrm{T}=\lambda_{1} \mathrm{E}_{1}+\cdots+$ $\lambda_{\mathrm{r}} \mathrm{E}_{\mathrm{r}}$. Show that $\mathrm{f}(\mathrm{T})=\mathrm{f}\left(\lambda_{1}\right) \mathrm{E}_{1}+\cdots+\mathrm{f}\left(\lambda_{\mathrm{r}}\right) \mathrm{E}_{\mathrm{r}}$ for any $\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$.
(b) Use part (a) to conclude that the minimal polynomial for T must be of the form $\mathrm{m}(\mathrm{x})=\left(\mathrm{x}-\lambda_{1}\right) \cdots\left(\mathrm{x}-\lambda_{\mathrm{r}}\right)$.
(c) Now suppose $T \in L(V)$ has minimal polynomial

$$
m(x)=\left(x-\lambda_{1}\right) \cdots\left(x-\lambda_{r}\right)
$$

where $\lambda_{1}, \ldots, \lambda_{r} \in \mathcal{F}$ are distinct. Define the polynomials

$$
p_{j}(x)=\prod_{i \neq j}\left(\frac{x-\lambda_{i}}{\lambda_{j}-\lambda_{i}}\right) .
$$

Note that $\operatorname{deg} p_{j}=r-1<\operatorname{deg} m$. By Exercise 6.4.2, any polynomial $f$ of degree $\leq r-1$ can be written as $f=\sum_{j} f\left(\lambda_{j}\right) p_{j}$. Defining $E_{j}=p_{j}(T)$, show that $\mathrm{E}_{\mathrm{j}} \neq 0$ and that

$$
1=E_{1}+\cdots+E_{r}
$$

and

$$
\mathrm{T}=\lambda_{1} \mathrm{E}_{1}+\cdots+\lambda_{\mathrm{r}} \mathrm{E}_{\mathrm{r}}
$$

(d) Show that $m \mid p_{i} p_{j}$ for $\mathrm{i} \neq \mathrm{j}$, and hence show that $\mathrm{E}_{\mathrm{i}} \mathrm{E}_{\mathrm{j}}=0$ for $\mathrm{i} \neq \mathrm{j}$.
(e) Conclude that T is diagonalizable.
4. Let $\mathrm{E}_{1}, \ldots, \mathrm{E}_{\mathrm{r}}$ and $\mathrm{W}_{1}, \ldots, \mathrm{~W}_{\mathrm{r}}$ be as defined in Theorem 7.28, and suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$.
(a) If $\mathrm{TE}_{\mathrm{i}}=\mathrm{E}_{\mathrm{i}} \mathrm{T}$ for every $\mathrm{E}_{\mathrm{i}}$, prove that every $\mathrm{W}_{\mathrm{j}}=\operatorname{Im} \mathrm{E}_{\mathrm{j}}$ is T-invariant.
(b) If every $\mathrm{W}_{\mathrm{j}}$ is T -invariant, prove that $\mathrm{TE}_{\mathrm{i}}=\mathrm{E}_{\mathrm{i}} \mathrm{T}$ for every $\mathrm{E}_{\mathrm{i}}$. [Hint: Let $\mathrm{v} \in \mathrm{V}$ be arbitrary. Show that property (a) of Theorem 7.28 implies $T\left(E_{i} v\right)=w_{i}$ for some $w_{i} \in W_{i}=\operatorname{Im} E_{i}$. Now show that $E_{j}\left(T E_{i}\right) v=$ $\left(\mathrm{E}_{\mathrm{i}} \mathrm{w}_{\mathrm{i}}\right) \delta_{\mathrm{ij}}$, and hence that $\mathrm{E}_{\mathrm{j}}(\mathrm{Tv})=\mathrm{T}\left(\mathrm{E}_{\mathrm{j}} \mathrm{v}\right)$.]
5. Prove that property (e) in Theorem 7.29 holds for the matrices $\mathrm{E}_{\mathrm{i}}$ given prior to the theorem.
6. Let W be a finite-dimensional subspace of an inner product space V .
(a) Show that there exists precisely one orthogonal projection on W.
(b) Let E be the orthogonal projection on W . Show that for any $\mathrm{v} \in \mathrm{V}$ we have $\|v-E v\| \leq\|v-w\|$ for every $w \in \mathrm{~W}$. In other words, show that Ev is the unique element of W that is "closest" to v .

### 7.9 QUOTIENT SPACES

Recall that in Section 1.5 we gave a brief description of normal subgroups and quotient groups (see Theorem 1.12). In this section we elaborate on and apply this concept to vector spaces, which are themselves abelian groups. In the next section we will apply this formalism to proving the triangular form theorem for linear operators.

Let V be a vector space and W a subspace of V. Since W is an abelian subgroup of V , it is easy to see that W is just a normal subgroup since for any $\mathrm{w} \in \mathrm{W}$ and $\mathrm{v} \in \mathrm{V}$ we have $\mathrm{v}+\mathrm{w}+(-\mathrm{v})=\mathrm{w} \in \mathrm{W}$ (remember that group multiplication in an abelian group is frequently denoted by the usual addition sign, and the inverse of an element $v$ is just -v ). We may therefore define the quotient group $\mathrm{V} / \mathrm{W}$ whose elements, $\mathrm{v}+\mathrm{W}$ for any $\mathrm{v} \in \mathrm{V}$, are just the cosets of W in V . It should be obvious that $\mathrm{V} / \mathrm{W}$ is also an abelian group. In fact, we will show below that V/W can easily be made into a vector space.

Example 7.12 Let $V=\mathbb{R}^{2}$ and suppose

$$
W=\left\{(x, y) \in \mathbb{R}^{2}: y=m x \text { for some fixed scalar } m\right\} .
$$

In other words, $W$ is just a line through the origin in the plane $\mathbb{R}^{2}$. The elements of $\mathrm{V} / \mathrm{W}$ are the cosets $\mathrm{v}+\mathrm{W}=\{\mathrm{v}+\mathrm{w}: \mathrm{w} \in \mathrm{W}\}$ where v is any vector in $V$.


Therefore, the set V/W consists of all lines in $\mathbb{R}^{2}$ that are parallel to W (i.e., that are displaced from W by the vector v). //

While we proved in Section 1.5 that cosets partition a group into disjoint subsets, let us repeat this proof in a different manner that should help familiarize us with some of the properties of V/W. We begin with several simple properties that are grouped together as a theorem for the sake of reference.

Theorem 7.30 Let W be a subspace of a vector space V . Then the following properties are equivalent:
(a) $u \in v+W$;
(b) $u-v \in W$;
(c) $v \in u+W$;
(d) $u+W=v+W$.

Proof $(a) \Rightarrow(b)$ : If $u \in v+W$, then there exists $w \in W$ such that $u=v+w$. But then $u-v=w \in W$.
(b) $\Rightarrow$ (c): $u-v \in W$ implies $v-u=-(u-v) \in W$, and hence there exists $\mathrm{w} \in \mathrm{W}$ such that $\mathrm{v}-\mathrm{u}=\mathrm{w}$. But then $\mathrm{v}=\mathrm{u}+\mathrm{w} \in \mathrm{u}+\mathrm{W}$.
$(c) \Rightarrow(d):$ If $v \in u+W$, then there exists $w \in W$ such that $v=u+w$. But then $v+W=u+w+W=u+W$.
$(\mathrm{d}) \Rightarrow(\mathrm{a}): 0 \in \mathrm{~W}$ implies $\mathrm{u}=\mathrm{u}+0 \in \mathrm{u}+\mathrm{W}=\mathrm{v}+\mathrm{W}$.

Theorem 7.31 Let W be a subspace of a vector space V . Then the cosets of W in V are distinct, and every element of V lies in some coset of W .

Proof It is easy to see that any $\mathrm{v} \in \mathrm{V}$ lies in $\mathrm{V} / \mathrm{W}$ since $\mathrm{v}=\mathrm{v}+0 \in \mathrm{v}+\mathrm{W}$. Now suppose that $v_{1} \neq v_{2}$ and that the cosets $v_{1}+W$ and $v_{2}+W$ have some element $u$ in common. Then $u \in v_{1}+W$ and $u \in v_{2}+W$, and hence by Theorem 7.30 we have $\mathrm{v}_{1}+\mathrm{W}=\mathrm{u}+\mathrm{W}=\mathrm{v}_{2}+\mathrm{W}$.

Let V be a vector space over $\mathcal{F}$, and let W be a subspace of V . We propose to make $\mathrm{V} / \mathrm{W}$ into a vector space. If $\alpha \in \mathcal{F}$ and $\mathrm{u}+\mathrm{W}, \mathrm{v}+\mathrm{W} \in \mathrm{V} / \mathrm{W}$, we define

$$
(u+W)+(v+W)=(u+v)+W
$$

and

$$
\alpha(u+W)=\alpha u+W .
$$

The first thing to do is show that these operations are well-defined. In other words, if we suppose that $u+W=u^{\prime}+W$ and $v+W=v^{\prime}+W$, then we must show that $(u+v)+W=\left(u^{\prime}+v^{\prime}\right)+W$ and $\alpha(u+W)=\alpha\left(u^{\prime}+W\right)$. Using $u+$ $\mathrm{W}=\mathrm{u}^{\prime}+\mathrm{W}$ and $\mathrm{v}+\mathrm{W}=\mathrm{v}^{\prime}+\mathrm{W}$, Theorem 7.30 tells us that $\mathrm{u}-\mathrm{u}^{\prime} \in \mathrm{W}$ and $\mathrm{v}-\mathrm{v}^{\prime} \in \mathrm{W}$. But then

$$
(\mathrm{u}+\mathrm{v})-\left(\mathrm{u}^{\prime}+\mathrm{v}^{\prime}\right)=\left(\mathrm{u}-\mathrm{u}^{\prime}\right)+\left(\mathrm{v}-\mathrm{v}^{\prime}\right) \in \mathrm{W}
$$

and hence $(u+v)+W=\left(u^{\prime}+v^{\prime}\right)+W$. Next, we see that $u-u^{\prime} \in W$ implies $\alpha\left(\mathrm{u}-\mathrm{u}^{\prime}\right) \in \mathrm{W}$ since W is a subspace. Then $\alpha u-\alpha u^{\prime} \in \mathrm{W}$ implies $\alpha u+\mathrm{W}=$ $\alpha u^{\prime}+W$, or $\alpha(u+W)=\alpha\left(u^{\prime}+W\right)$.

Theorem 7.32 Let V be a vector space over $\mathcal{F}$ and W a subspace of V . For any $\mathrm{u}+\mathrm{W}, \mathrm{v}+\mathrm{W} \in \mathrm{V} / \mathrm{W}$ and $\alpha \in \mathcal{F}$, we define the operations
(1) $(\mathrm{u}+\mathrm{W})+(\mathrm{v}+\mathrm{W})=(\mathrm{u}+\mathrm{v})+\mathrm{W}$
(2) $\alpha(u+W)=\alpha u+W$.

Then, with these operations, V/W becomes a vector space over $\mathcal{F}$.

Proof Since $(0+W)+(u+W)=(0+u)+W=u+W$, we see that $W$ is the zero element of V/W. Similarly, we see that $-(u+W)=-u+W$. In view of the above discussion, all that remains is to verify axioms (V1) - (V8) for a vector space given at the beginning of Section 2.1. We leave this as an exercise for the reader (see Exercise 7.9.1).

The vector space V/W defined in this theorem is called the quotient space of V by W . If V is finite-dimensional, then any subspace $\mathrm{W} \subset \mathrm{V}$ is also finite-
dimensional, where in fact $\operatorname{dim} \mathrm{W} \leq \operatorname{dim} \mathrm{V}$ (Theorem 2.9). It is then natural to ask about the dimension of V/W.

Theorem 7.33 Let V be finite-dimensional over $\mathcal{F}$ and W be a subspace of V. Then

$$
\operatorname{dim} \mathrm{V} / \mathrm{W}=\operatorname{dim} \mathrm{V}-\operatorname{dim} \mathrm{W}
$$

Proof Suppose $\left\{\mathrm{w}_{1}, \ldots, \mathrm{w}_{\mathrm{m}}\right\}$ is a basis for W. By Theorem 2.10 we can extend this to a basis $\left\{w_{1}, \ldots, w_{m}, v_{1}, \ldots, v_{r}\right\}$ for $V$, where $\operatorname{dim} V=m+r=$ $\operatorname{dim} \mathrm{W}+\mathrm{r}$. Then any $\mathrm{v} \in \mathrm{V}$ may be written as

$$
\mathrm{v}=\alpha_{1} \mathrm{w}_{1}+\cdots+\alpha_{\mathrm{m}} \mathrm{w}_{\mathrm{m}}+\beta_{1} \mathrm{v}_{1}+\cdots+\beta_{\mathrm{r}} \mathrm{v}_{\mathrm{r}}
$$

where $\left\{\alpha_{i}\right\},\left\{\beta_{j}\right\} \in \mathcal{F}$. For ease of notation, let us define $\overline{\mathrm{V}}=\mathrm{V} / \mathrm{W}$ and, for any $\mathrm{v} \in \mathrm{V}$, we let $\overline{\mathrm{v}}=\mathrm{v}+\mathrm{W} \in \overline{\mathrm{V}}$. Note that this association is linear because (by Theorem 7.32)

$$
\overline{\mathrm{v}+\mathrm{v}^{\prime}}=\mathrm{v}+\mathrm{v}^{\prime}+\mathrm{W}=\mathrm{v}+\mathrm{W}+\mathrm{v}^{\prime}+\mathrm{W}=\overline{\mathrm{v}}+\overline{\mathrm{v}}^{\prime}
$$

and

$$
\overline{\mathrm{kv}}=\mathrm{kv}+\mathrm{W}=\mathrm{k}(\mathrm{v}+\mathrm{W})=\mathrm{k} \overline{\mathrm{v}}
$$

Since $\mathrm{w}_{\mathrm{i}} \in \mathrm{W}$, we see that $\overline{\mathrm{w}}_{\mathrm{i}}=\mathrm{w}_{\mathrm{i}}+\mathrm{W}=\mathrm{W}$, and hence $\overline{\mathrm{v}}=\beta_{1} \overline{\mathrm{v}}_{1}+\cdots+\beta_{\mathrm{r}} \overline{\mathrm{v}}_{\mathrm{r}}$. Alternatively, any $\overline{\mathrm{v}} \in \overline{\mathrm{V}}$ may be written as

$$
\overline{\mathrm{v}}=\mathrm{v}+\mathrm{W}=\sum_{i} \alpha_{\mathrm{i}} \mathrm{~W}_{\mathrm{i}}+\sum_{j} \beta_{\mathrm{j}} \mathrm{v}_{\mathrm{j}}+\mathrm{W}=\sum_{\mathrm{j}} \beta_{\mathrm{j}} \mathrm{v}_{\mathrm{j}}+\mathrm{W}=\sum_{j} \beta_{\mathrm{j}} \overline{\mathrm{v}}_{\mathrm{j}}
$$

since $\sum \alpha_{\mathrm{i}} \mathrm{W}_{\mathrm{i}} \in \mathrm{W}$. In any case, this shows that the $\overline{\mathrm{v}}_{\mathrm{i}}$ span $\overline{\mathrm{V}}$. Now suppose that $\sum_{i} \gamma_{i} v_{i}=0$ for some scalars $\gamma_{i} \in \mathcal{F}$. Then

$$
\gamma_{1} \overline{\mathrm{v}}_{1}+\cdots+\gamma_{\mathrm{r}} \overline{\mathrm{v}}_{\mathrm{r}}=\overline{0}=\mathrm{W}
$$

Using $\bar{v}_{i}=v_{i}+W$, we then see that $\gamma_{1} v_{1}+\cdots+\gamma_{r} v_{r}+W=W$ which implies that $\sum_{i} \gamma_{i} v_{i} \in W$. But $\left\{\mathrm{w}_{\mathrm{i}}\right\}$ forms a basis for W , and hence there exist $\delta_{1}, \ldots$, $\delta_{\mathrm{m}} \in \mathcal{F}$ such that

$$
\gamma_{1} \mathrm{v}_{1}+\cdots+\gamma_{\mathrm{r}} \mathrm{v}_{\mathrm{r}}=\delta_{1} \mathrm{w}_{1}+\cdots+\delta_{\mathrm{m}} \mathrm{w}_{\mathrm{m}} .
$$

However, $\left\{\mathrm{w}_{1}, \ldots, \mathrm{w}_{\mathrm{m}}, \mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{r}}\right\}$ is a basis for V and hence is linearly independent. This means that $\gamma_{i}=0$ for each $i=1, \ldots, r$ and that $\delta_{j}=0$ for each $\mathrm{j}=1, \ldots, \mathrm{~m}$. Thus $\left\{\overline{\mathrm{v}}_{\mathrm{i}}\right\}$ is linearly independent and forms a basis for $\overline{\mathrm{V}}=$ $\mathrm{V} / \mathrm{W}$, and $\operatorname{dim} \mathrm{V} / \mathrm{W}=\mathrm{r}=\operatorname{dim} \mathrm{V}-\operatorname{dim} \mathrm{W}$.

There is a slightly different way of looking at this result that will be of use to us later.

Theorem 7.34 Let V be finite-dimensional over $\mathcal{F}$ and W be a subspace of V. Suppose that W has basis $w_{1}, \ldots, w_{m}$ and $\overline{\mathrm{V}}=\mathrm{V} / \mathrm{W}$ has basis $\overline{\mathrm{v}}_{1}, \ldots, \overline{\mathrm{v}}_{\mathrm{r}}$ where $\bar{v}_{i}=v_{i}+W$ for some $v_{i} \in V$. Then $\left\{w_{1}, \ldots, w_{m}, v_{1}, \ldots, v_{r}\right\}$ is a basis for V .

Proof Let $\mathrm{u} \in \mathrm{V}$ be arbitrary. Then $\overline{\mathrm{u}}=\mathrm{u}+\mathrm{W} \in \overline{\mathrm{V}}$, and hence there exists $\left\{\alpha_{i}\right\} \in \mathcal{F}$ such that

$$
\mathrm{u}+\mathrm{W}=\overline{\mathrm{u}}=\alpha_{1} \overline{\mathrm{v}}_{1}+\cdots+\alpha_{\mathrm{r}} \overline{\mathrm{v}}_{\mathrm{r}}=\alpha_{1} \mathrm{v}_{1}+\cdots+\alpha_{\mathrm{r}} \mathrm{v}_{\mathrm{r}}+\mathrm{W} .
$$

By Theorem 7.30 there exists $\mathrm{w}=\beta_{1} \mathrm{w}_{1}+\cdots+\beta_{\mathrm{m}} \mathrm{w}_{\mathrm{m}} \in \mathrm{W}$ such that

$$
u=\alpha_{1} v_{1}+\cdots+\alpha_{r} v_{r}+w=\alpha_{1} v_{1}+\cdots+\alpha_{r} v_{r}+\beta_{1} w_{1}+\cdots+\beta_{m} w_{m}
$$

This shows that $\left\{\mathrm{w}_{1}, \ldots, \mathrm{w}_{\mathrm{m}}, \mathrm{v}_{\mathrm{l}}, \ldots, \mathrm{v}_{\mathrm{r}}\right\}$ spans V .
To show that these vectors are linearly independent, we suppose that

$$
\gamma_{1} \mathrm{w}_{1}+\cdots+\gamma_{\mathrm{m}} \mathrm{w}_{\mathrm{m}}+\delta_{1} \mathrm{v}_{1}+\cdots+\delta_{\mathrm{r}} \mathrm{v}_{\mathrm{r}}=0 .
$$

Since the association between V and $\overline{\mathrm{V}}$ is linear (see the proof of Theorem 7.33) and $\overline{\mathrm{w}}_{\mathrm{i}}=\mathrm{w}_{\mathrm{i}}+\mathrm{W}=\mathrm{W}$, we see that

$$
\delta_{1} \overline{\mathrm{v}}_{1}+\cdots+\delta_{\mathrm{r}} \overline{\mathrm{v}}_{\mathrm{r}}=\overline{0}=\mathrm{W} .
$$

But the $\bar{v}_{i}$ are linearly independent (since they are a basis for $\overline{\mathrm{V}}$ ), and hence $\delta_{1}=\cdots=\delta_{\mathrm{r}}=0$. (This is just the definition of linear independence if we recall that $W$ is the zero vector in $\overline{\mathrm{V}}=\mathrm{V} / \mathrm{W}$.) This leaves us with $\gamma_{1} \mathrm{~W}_{1}+\cdots+\gamma_{\mathrm{m}} \mathrm{W}_{\mathrm{m}}$ $=0$. But again, the $w_{i}$ are linearly independent, and therefore $\gamma_{1}=\cdots=\gamma_{\mathrm{m}}=0$. This shows that $\left\{w_{1}, \ldots, w_{m}, v_{1}, \ldots, v_{r}\right\}$ is a basis for $V$, and hence $\operatorname{dim} V=$ $\operatorname{dim} \mathrm{W}+\operatorname{dim} \mathrm{V} / \mathrm{W}$.

## Exercises

1. Finish the proof of Theorem 7.32.
2. Let U and V be finite-dimensional, and suppose T is a linear transformation of U onto V . If $\mathrm{W}=$ Ker T , prove that V is isomorphic to $\mathrm{U} / \mathrm{W}$. [Hint: See Exercise 1.5.11.]
3. Let V be a vector space over $\mathcal{F}$, and let W be a subspace of V . Define a relation $R$ on the set $V$ by $x R y$ if $x-y \in W$.
(a) Show that R defines an equivalence relation on V .
(b) Let the equivalence class of $x \in V$ be denoted by [x], and define the quotient set $\mathrm{V} / \mathrm{R}$ to be the set of all such equivalence classes. For all $\mathrm{x}, \mathrm{y} \in$ V and $\mathrm{a} \in \mathcal{F}$ we define addition and scalar multiplication in $\mathrm{V} / \mathrm{R}$ by

$$
[\mathrm{x}]+[\mathrm{y}]=[\mathrm{x}+\mathrm{y}]
$$

and

$$
\mathrm{a}[\mathrm{x}]=[\mathrm{ax}]
$$

Show that these operation are well-defined, and that $\mathrm{V} / \mathrm{R}$ is a vector space over $\mathcal{F}$.
(c) Now assume that V is finite-dimensional, and define the mapping T : $\mathrm{V} \rightarrow \mathrm{V} / \mathrm{R}$ by $\mathrm{Tx}=[\mathrm{x}]$. Show that this defines a linear transformation.
(d) Using Theorem 5.6, prove that $\operatorname{dim} \mathrm{V} / \mathrm{R}+\operatorname{dim} \mathrm{W}=\operatorname{dim} \mathrm{V}$.

### 7.10 THE TRIANGULAR FORM THEOREM *

Now that we know something about quotient spaces, let us look at the effect of a linear transformation on such a space. Unless otherwise noted, we restrict our discussion to finite-dimensional vector spaces. In particular, suppose that $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and W is a T -invariant subspace of V . We first show that T induces a natural linear transformation on the space $\mathrm{V} / \mathrm{W}$. (The reader should be careful to note that $\overline{0}$ in Theorem 7.33 is the zero vector in $\overline{\mathrm{V}}=\mathrm{V} / \mathrm{W}$, while in the theorem below, $\overline{0}$ is the zero transformation on $\overline{\mathrm{V}}$.)

Theorem 7.35 Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and let W be a T-invariant subspace of V . Then T induces a linear transformation $\overline{\mathrm{T}} \in \mathrm{L}(\mathrm{V} / \mathrm{W})$ defined by

$$
\overline{\mathrm{T}}(\mathrm{v}+\mathrm{W})=\mathrm{T}(\mathrm{v})+\mathrm{W} .
$$

Furthermore, if T satisfies any polynomial $\mathrm{p}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$, then so does $\overline{\mathrm{T}}$. In particular, the minimal polynomial $\overline{\mathrm{m}}(\mathrm{x})$ for $\overline{\mathrm{T}}$ divides the minimal polynomial $m(x)$ for $T$.

Proof Our first task is to show that $\overline{\mathrm{T}}$ is well-defined and linear. Thus, suppose $\mathrm{v}+\mathrm{W}=\mathrm{v}^{\prime}+\mathrm{W}$. Then $\mathrm{v}-\mathrm{v}^{\prime} \in \mathrm{W}$ so that $\mathrm{T}\left(\mathrm{v}-\mathrm{v}^{\prime}\right)=\mathrm{T}(\mathrm{v})-\mathrm{T}\left(\mathrm{v}^{\prime}\right) \in \mathrm{W}$ since W is T-invariant. Therefore, using Theorem 7.30, we see that

$$
\overline{\mathrm{T}}(\mathrm{v}+\mathrm{W})=\mathrm{T}(\mathrm{v})+\mathrm{W}=\mathrm{T}\left(\mathrm{v}^{\prime}\right)+\mathrm{W}=\overline{\mathrm{T}}\left(\mathrm{v}^{\prime}+\mathrm{W}\right)
$$

and hence $\overline{\mathrm{T}}$ is well-defined. To show that $\overline{\mathrm{T}}$ is a linear transformation, we simply calculate

$$
\begin{aligned}
\bar{T}\left[\left(v_{1}+W\right)+\left(v_{2}+W\right)\right] & =\bar{T}\left(v_{1}+v_{2}+W\right)=T\left(v_{1}+v_{2}\right)+W \\
& =T\left(v_{1}\right)+T\left(v_{2}\right)+W=T\left(v_{1}\right)+W+T\left(v_{2}\right)+W \\
& =\bar{T}\left(v_{1}+W\right)+\bar{T}\left(v_{2}+W\right)
\end{aligned}
$$

and

$$
\begin{aligned}
\bar{T}[\alpha(v+W)] & =\bar{T}(\alpha v+W)=T(\alpha v)+W=\alpha T(v)+W \\
& =\alpha[T(v)+W]=\alpha \bar{T}(v+W)
\end{aligned}
$$

This proves that $\overline{\mathrm{T}}$ is indeed a linear transformation.
Next we observe that for any $\mathrm{T} \in \mathrm{L}(\mathrm{V}), \mathrm{T}^{2}$ is a linear transformation and W is also invariant under $\mathrm{T}^{2}$. This means that we can calculate the effect of $\overline{\mathrm{T}^{2}}$ on any $\mathrm{v}+\mathrm{W} \in \mathrm{V} / \mathrm{W}$ :

$$
\begin{aligned}
\overline{T^{2}}(v+W) & =T^{2}(v)+W=T[T(v)]+W=\bar{T}[T(v)+W] \\
& =\bar{T}[\bar{T}(v+W)]=\bar{T}^{2}(v+W) .
\end{aligned}
$$

This shows that $\overline{\mathrm{T}^{2}}=\overline{\mathrm{T}}^{2}$, and it is easy to see that in general $\overline{\mathrm{T}}^{\overline{\mathrm{m}}}=\overline{\mathrm{T}}^{\mathrm{m}}$ for any $\mathrm{m} \geq 0$. Then for any $\mathrm{p}(\mathrm{x})=\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{x}+\cdots+\mathrm{a}_{\mathrm{n}} \mathrm{x}^{\mathrm{n}} \in \mathcal{F}[\mathrm{x}]$, we have

$$
\begin{aligned}
\overline{p(T)}(v+W) & =p(T)(v)+W=\sum a_{m} T^{m}(v)+W=\sum a_{m}\left[T^{m}(v)+W\right] \\
& =\sum a_{m} \overline{T^{m}}(v+W)=\Sigma a_{m} \bar{T}^{m}(v+W)=p(\bar{T})(v+W)
\end{aligned}
$$

so that $\overline{\mathrm{p}(\mathrm{T})}=\mathrm{p}(\overline{\mathrm{T}})$.
Now note that for any $v+W \in V / W$ we have $\overline{0}(v+W)=0(v)+W=W$, and hence $\overline{0}$ is the zero transformation on V/W (since W is the zero vector in $\mathrm{V} / \mathrm{W})$. Therefore, if $\mathrm{p}(\mathrm{T})=0$ for some $\mathrm{p}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$, we see that $\overline{0}=\overline{\mathrm{p}(\mathrm{T})}=$ $p(\bar{T})$ and hence $\bar{T}$ satisfies $p(x)$ also.

Finally, let $\overline{\mathrm{m}}(\mathrm{x})$ be the minimal polynomial for $\overline{\mathrm{T}}$. If $\mathrm{p}(\mathrm{x})$ is such that $p(\bar{T})=\overline{0}$, then we know that $\bar{m} \mid p$ (Theorem 7.4). If $m(x)$ is the minimal polynomial for T , then $\mathrm{m}(\mathrm{T})=0$, and therefore $\mathrm{m}(\overline{\mathrm{T}})=\overline{0}$ by what we just proved. Hence $\overline{\mathrm{m}} \mid \mathrm{m}$.

We now come to the main result of this section. By way of motivation, we saw in Theorem 7.24 that a linear transformation $T \in L(V)$ is diagonalizable if and only if the minimal polynomial for T can be written as the product of distinct linear factors. What we wish to do now is take into account the more general case where all of the distinct factors of the minimal polynomial are linear, but possibly with a multiplicity greater than one. We shall see that this leads to a triangular form for the matrix representation of T.

For definiteness we consider upper-triangular matrices, so what we are looking for is a basis $\left\{\mathrm{v}_{\mathrm{i}}\right\}$ for V in which the action of T takes the form

$$
\begin{aligned}
T\left(v_{1}\right) & =v_{1} a_{11} \\
T\left(v_{2}\right) & =v_{1} a_{12}+v_{2} a_{22} \\
& \vdots \\
T\left(v_{n}\right) & =v_{1} a_{1 n}+\cdots+v_{n} a_{n n} .
\end{aligned}
$$

We present two versions of this theorem in the present chapter. The first is more intuitive, while the second requires the development of some additional material that is also of use in other applications. Because of this, we postpone the second version until after we have discussed nilpotent transformations in the next section. (Actually, Exercises 7.5.7 and 7.5.8 outlined another way to prove the first version that used the formalism of ideals and annihilators, and made no reference whatsoever to quotient spaces.) Furthermore, in Section 8.1 we give a completely independent proof for the special case of matrices over an algebraically closed field.

Recall from Theorem 7.9 that an element $\lambda \in \mathcal{F}$ is an eigenvalue of $T$ if and only if $\lambda$ is a root of the characteristic polynomial of T . Thus all the eigenvalues of T lie in $\mathcal{F}$ if and only if the characteristic polynomial factors into linear terms.

Theorem 7.36 (Triangular Form Theorem) Suppose $T \in L(V)$ has a characteristic polynomial that factors into (not necessarily distinct) linear terms. Then V has a basis in which the matrix of T is triangular.

Proof If $\operatorname{dim} \mathrm{V}=1$, then T is represented by a $1 \times 1$ matrix which is certainly triangular. Now suppose that $\operatorname{dim} \mathrm{V}=\mathrm{n}>1$. We assume the theorem is true for $\operatorname{dim} \mathrm{V}=\mathrm{n}-1$, and proceed by induction to prove the result for $\operatorname{dim} \mathrm{V}=\mathrm{n}$. Since the characteristic polynomial of T factors into linear polynomials, there exists at least one nonzero eigenvalue $\lambda_{1}$ and corresponding eigenvector $v_{1}$ such that $\mathrm{T}\left(\mathrm{v}_{1}\right)=\lambda_{1} \mathrm{v}_{1} \equiv \mathrm{a}_{11} \mathrm{v}_{1}$. Let W be the one-dimensional T -invariant subspace spanned by $\mathrm{v}_{1}$, and define $\overline{\mathrm{V}}=\mathrm{V} / \mathrm{W}$ so that (by Theorem 7.33) $\operatorname{dim} \overline{\mathrm{V}}=\operatorname{dim} \mathrm{V}-\operatorname{dim} \mathrm{W}=\mathrm{n}-1$.

According to Theorem 7.35, T induces a linear transformation $\overline{\mathrm{T}}$ on $\overline{\mathrm{V}}$ such that the minimal polynomial $\overline{\mathrm{m}}(\mathrm{x})$ for $\overline{\mathrm{T}}$ divides the minimal polynomial $\mathrm{m}(\mathrm{x})$ for T . But this means that any root of $\overline{\mathrm{m}}(\mathrm{x})$ is also a root of $\mathrm{m}(\mathrm{x})$. Since the characteristic polynomial of T factors into linear polynomials by hypothesis, so does $m(x)$ (see Theorem 7.12). Therefore $\bar{m}(x)$ must also factor into linear polynomials, and hence so does the characteristic polynomial of T. This shows that $\overline{\mathrm{T}}$ and $\overline{\mathrm{V}}$ satisfy the hypotheses of the theorem, and hence there exists a basis $\left\{\bar{v}_{2}, \ldots, \bar{v}_{n}\right\}$ for $\overline{\mathrm{V}}=\mathrm{V} / \mathrm{W}$ such that

$$
\begin{aligned}
& \bar{T}\left(\bar{v}_{2}\right)=\bar{v}_{2} a_{22} \\
& \bar{T}\left(\bar{v}_{3}\right)=\bar{v}_{2} a_{23}+\bar{v}_{3} a_{33} \\
& \quad \vdots \\
& \bar{T}\left(\bar{v}_{n}\right)=\bar{v}_{2} a_{2 n}+\cdots+\bar{v}_{n} a_{n n} .
\end{aligned}
$$

We now let $\mathrm{v}_{2}, \ldots, \mathrm{v}_{\mathrm{n}}$ be elements of V such that $\overline{\mathrm{v}}_{\mathrm{i}}=\mathrm{v}_{\mathrm{i}}+\mathrm{W}$ for each $\mathrm{i}=$ $2, \ldots, n$. Since W has basis $\left\{\mathrm{v}_{1}\right\}$, Theorem 7.35 tells us that $\left\{\mathrm{v}_{1}, \mathrm{v}_{2}, \ldots, \mathrm{v}_{\mathrm{n}}\right\}$ is a basis for V. According to our above result, we have $\bar{T}\left(\bar{v}_{2}\right)=\bar{v}_{2} a_{22}$ which is equivalent to $\overline{\mathrm{T}}\left(\overline{\mathrm{v}}_{2}\right)-\overline{\mathrm{v}}_{2} \mathrm{a}_{22}=\overline{0}$, and hence the definition of $\overline{\mathrm{T}}$ (see Theorem 7.35) along with Theorem 7.30 tells us that $T\left(v_{2}\right)-v_{2} a_{22} \in W$. Since $W$ is spanned by $v_{1}$, this says there exists $a_{12} \in \mathcal{F}$ such that $T\left(v_{2}\right)-v_{2} a_{22}=v_{1} a_{12}$, i.e., $T\left(v_{2}\right)=v_{1} a_{12}+v_{2} a_{22}$. Clearly, an identical argument holds for any of the $\overline{\mathrm{T}}\left(\overline{\mathrm{v}}_{\mathrm{i}}\right)$, and thus for each $\mathrm{i}=2, \ldots, \mathrm{n}$ there exists $\mathrm{a}_{1 \mathrm{i}} \in \mathcal{F}$ such that $\mathrm{T}\left(\mathrm{v}_{\mathrm{i}}\right)$ $\mathrm{v}_{2} \mathrm{a}_{2 \mathrm{i}}-\cdots-\mathrm{v}_{\mathrm{i}} \mathrm{a}_{\mathrm{ii}} \in \mathrm{W}$ implies

$$
T\left(v_{i}\right)=v_{1} a_{1 i}+v_{2} a_{2 i}+\cdots+v_{i} a_{i i}
$$

Written out, this is just

$$
\begin{aligned}
T\left(v_{1}\right) & =v_{1} a_{11} \\
T\left(v_{2}\right) & =v_{1} a_{12}+v_{2} a_{22} \\
& \vdots \\
T\left(v_{n}\right) & =v_{1} a_{1 n}+\cdots+v_{n} a_{n n} .
\end{aligned}
$$

In other words, the elements $\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{n}} \in \mathrm{V}$ are a basis for V in which every $T\left(v_{i}\right)$ is a linear combination of $v_{j}$ for $j \leq i$. This is precisely the definition of triangular form.

We now give a restatement of this theorem in terms of matrices. For ease of reference, this version is presented as a corollary.

Corollary Let $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ be a matrix whose characteristic polynomial factors into linear polynomials. Then A is similar to a triangular matrix.

Proof The matrix $\mathrm{A}=\left(\mathrm{a}_{\mathrm{ij}}\right)$ defines a linear transformation T on the space $\mathcal{F}^{\mathrm{n}}$ by $\mathrm{T}\left(\mathrm{v}_{\mathrm{i}}\right)=\sum_{\mathrm{j}=1}^{\mathrm{n}} \mathrm{v}_{\mathrm{j}} \mathrm{a}_{\mathrm{ji}}$ where $\left\{\mathrm{v}_{\mathrm{i}}\right\}$ is a basis for $\mathcal{F}^{\mathrm{n}}$. In particular, relative to the basis $\left\{\mathrm{v}_{\mathrm{i}}\right\}$, the matrix representation of T is precisely the matrix A (since T takes the i th basis vector into the i th column of the matrix representing T ). Since the characteristic polynomial of T is independent of the basis used in the matrix representation of T, and the characteristic polynomial of A factors into linear polynomials, we see that Theorem 7.36 applies to T . Thus there is a basis for $\mathcal{F}^{\mathrm{n}}$ in which the matrix of T (i.e., the matrix A ) is triangular. By Theorem 5.18 we then see that A must be similar to a triangular matrix.

If a linear transformation T can be represented by a triangular matrix, then we say that T can be brought into triangular form. Since $\lambda$ is an eigenvalue of T if and only if $\operatorname{det}(\lambda \mathrm{I}-\mathrm{T})=0$ (Theorem 7.9), Theorem 4.5 tells us that the eigenvalues of a triangular matrix are precisely the diagonal elements of the matrix (this was also discussed in the previous section).

### 7.11 NILPOTENT TRANSFORMATIONS *

An operator $T \in L(V)$ is said to be nilpotent if $T^{n}=0$ for some positive integer $n$. If $T^{k}=0$ but $T^{k-1} \neq 0$, then $k$ is called the index of nilpotency of $T$ (note that $\mathrm{T}^{\mathrm{k}-1} \neq 0$ implies that $\mathrm{T}^{\mathrm{j}} \neq 0$ for all $\mathrm{j} \leq \mathrm{k}-1$ ). This same terminology applies to any square matrix $A$ with the property that $A^{n}=0$. Some elementary facts about nilpotent transformations are contained in the following theorem. Note Theorem 7.1 implies that if A is the matrix representation of T and T is nilpotent with index k , then A is also nilpotent with index k .

Theorem 7.37 Suppose $T \in L(V)$, and assume that for some $v \in V$ we have $\mathrm{T}^{\mathrm{k}}(\mathrm{v})=0$ but $\mathrm{T}^{\mathrm{k}-1}(\mathrm{v}) \neq 0$. Define the set

$$
\mathrm{S}=\left\{\mathrm{v}, \mathrm{~T}(\mathrm{v}), \mathrm{T}^{2}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{k}-1}(\mathrm{v})\right\}
$$

Then $S$ has the following properties:
(a) The elements of S are linearly independent.
(b) The linear span W of S is a T -invariant subspace of V .
(c) The operator $\mathrm{T}_{\mathrm{W}}=\mathrm{T} \mid \mathrm{W}$ is nilpotent with index k .
(d) With respect to the ordered basis $\left\{\mathrm{T}^{\mathrm{k}-1}(\mathrm{v}), \ldots, \mathrm{T}(\mathrm{v}), \mathrm{v}\right\}$ for W , the matrix of $\mathrm{T}_{\mathrm{W}}$ has the form

$$
\left(\begin{array}{cccccc}
0 & 1 & 0 & \cdots & 0 & 0 \\
0 & 0 & 1 & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & \cdots & 0 & 1 \\
0 & 0 & 0 & \cdots & 0 & 0
\end{array}\right)
$$

Thus the matrix of $\mathrm{T}_{\mathrm{W}}$ has all zero entries except on the superdiagonal where they are all equal to one. This shows that the matrix representation of $\mathrm{T}_{\mathrm{W}}$ is a nilpotent matrix with index k .

Proof (a) Suppose that

$$
\alpha_{0} \mathrm{v}+\alpha_{1} \mathrm{~T}(\mathrm{v})+\cdots+\alpha_{\mathrm{k}-1} \mathrm{~T}^{\mathrm{k}-1}(\mathrm{v})=0
$$

for some set of scalars $\alpha_{i} \in \mathcal{F}$. Applying $\mathrm{T}^{\mathrm{k}-1}$ to this equation results in $\alpha_{0} \mathrm{~T}^{\mathrm{k}-1}(\mathrm{v})=0$. Since $\mathrm{T}^{\mathrm{k}-1}(\mathrm{v}) \neq 0$, this implies that $\alpha_{0}=0$. Using this result, apply $\mathrm{T}^{\mathrm{k}-2}$ to the above equation to obtain $\alpha_{1}=0$. Continuing this procedure, we eventually arrive at $\alpha_{i}=0$ for each $\mathrm{i}=0,1, \ldots, \mathrm{k}-1$ and hence the elements of $S$ are linearly independent.
(b) Since any $\mathrm{w} \in \mathrm{W}$ may be written in the form

$$
\mathrm{w}=\beta_{0} \mathrm{v}+\beta_{1} \mathrm{~T}(\mathrm{v})+\cdots+\beta_{\mathrm{k}-1} \mathrm{~T}^{\mathrm{k}-1}(\mathrm{v})
$$

we see that $T(w)=\beta_{0} T(v)+\cdots+\beta_{k-2} T^{k-1}(v) \in W$, and hence $T(W) \subset W$.
(c) Using $\mathrm{T}^{\mathrm{k}}(\mathrm{v})=0$, it follows that $\mathrm{T}_{\mathrm{W}}{ }^{\mathrm{k}}\left(\mathrm{T}^{\mathrm{i}}(\mathrm{v})\right)=\mathrm{T}^{\mathrm{k}+\mathrm{i}}(\mathrm{v})=0$ for each $\mathrm{i}=$ $0, \ldots, k-1$. This shows that $T_{W}{ }^{k}$ applied to each element of $S$ (i.e., each of the basis vectors for $W$ ) is zero, and thus $T_{W}{ }^{k}=0$. In addition, since $v \in W$ we see that $\mathrm{T}_{\mathrm{W}}{ }^{\mathrm{k}-1}(\mathrm{v})=\mathrm{T}^{\mathrm{k}-1}(\mathrm{v}) \neq 0$, and therefore $\mathrm{T}_{\mathrm{W}}$ is nilpotent with index k.
(d) Using $\mathrm{T}_{\mathrm{W}}\left(\mathrm{T}^{\mathrm{i}}(\mathrm{v})\right)=\mathrm{T}^{\mathrm{i}+1}(\mathrm{v})$ along with the fact that the i th column of $\left[\mathrm{T}_{\mathrm{W}}\right]$ is the image of the $\mathrm{i} t h$ basis vector for W , it is easy to see that $\left[\mathrm{T}_{\mathrm{W}}\right]$ has the desired form.

One must be careful to understand exactly what Theorem 7.37 says and what it does not say. In particular, if $T$ is nilpotent with index $k$, then $T^{k}(u)=$ 0 for all $\mathrm{u} \in \mathrm{V}$, while $\mathrm{T}^{\mathrm{k}-1}(\mathrm{v}) \neq 0$ for some $\mathrm{v} \in \mathrm{V}$. This is because if $\mathrm{w}=\mathrm{T}(\mathrm{v})$, then $T^{k-1}(w)=T^{k-1}(T(v))=T^{k}(v)=0$. Hence it is impossible for $T^{k-1}(v)$ to be nonzero for all $\mathrm{v} \in \mathrm{V}$.

It is also interesting to note that according to Theorem 7.37(b), the subspace W is T-invariant, and hence by Theorem 7.19, the matrix of T must be of the block form

$$
\left(\begin{array}{cc}
M_{k} & B \\
0 & C
\end{array}\right)
$$

where $\mathrm{M}_{\mathrm{k}}$ is the kx k matrix of $\mathrm{T}_{\mathrm{W}}=\mathrm{T} \mid \mathrm{W}$ defined in part (d) of Theorem 7.37. If we can find another T-invariant subspace U of V such that $\mathrm{V}=\mathrm{W} \oplus$ U , then the matrix representation of T will be in block diagonal form (Theorem 7.20). We now proceed to show that this can in fact be done. Let us first prove two more easy results.

Theorem 7.38 Let $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ be nilpotent, and let $\mathrm{S}=\alpha_{1} \mathrm{~T}+\cdots+\alpha_{\mathrm{m}} \mathrm{T}^{\mathrm{m}}$ where each $\alpha_{i} \in \mathcal{F}$. Then $\alpha_{0} 1+S$ is invertible for any nonzero $\alpha_{0} \in \mathcal{F}$.

Proof Suppose the index of T is k . Then $\mathrm{T}^{\mathrm{k}}=0$, and therefore $\mathrm{S}^{\mathrm{k}}=0$ also since the lowest power of T in the expansion of $S^{k}$ is $k$. If $\alpha_{0} \neq 0$, we leave it as a trivial exercise for the reader to show that

$$
\left(\alpha_{0} 1+S\right)\left(1 / \alpha_{0}-S / \alpha_{0}^{2}+S^{2} / \alpha_{0}^{3}+\cdots+(-1)^{\mathrm{k}-1} \mathrm{~S}^{\mathrm{k}-1} / \alpha_{0}{ }^{\mathrm{k}}\right)=1 .
$$

This shows that $\alpha_{0} 1+\mathrm{S}$ is invertible, and that its inverse is given by the above polynomial in S .

Theorem 7.39 Let $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ be nilpotent with index n , and let W be the T invariant subspace spanned by $\left\{\mathrm{T}^{\mathrm{n}-1}(\mathrm{v}), \ldots, \mathrm{T}(\mathrm{v}), \mathrm{v}\right\}$ where $\mathrm{v} \in \mathrm{V}$ is such that $\mathrm{T}^{\mathrm{n}-1}(\mathrm{v}) \neq 0$. If $\mathrm{w} \in \mathrm{W}$ is such that $\mathrm{T}^{\mathrm{n}-\mathrm{k}}(\mathrm{w})=0$ for some $0<\mathrm{k} \leq \mathrm{n}$, then there exists $\mathrm{w}_{0} \in \mathrm{~W}$ such that $\mathrm{T}^{\mathrm{k}}\left(\mathrm{w}_{0}\right)=\mathrm{w}$.

Proof Since $\mathrm{w} \in \mathrm{W}$, we have

$$
\mathrm{w}=\alpha_{\mathrm{n}} \mathrm{~T}^{\mathrm{n}-1}(\mathrm{v})+\cdots+\alpha_{\mathrm{k}+1} \mathrm{~T}^{\mathrm{k}}(\mathrm{v})+\alpha_{\mathrm{k}} \mathrm{~T}^{\mathrm{k}-1}(\mathrm{v})+\cdots+\alpha_{2} \mathrm{~T}(\mathrm{v})+\alpha_{1} \mathrm{v}
$$

and therefore $\left(\right.$ since $\left.\mathrm{T}^{\mathrm{n}}=0\right)$

$$
0=\mathrm{T}^{\mathrm{n}-\mathrm{k}}(\mathrm{w})=\alpha_{\mathrm{k}} \mathrm{~T}^{\mathrm{n}-\mathrm{l}}(\mathrm{v})+\cdots+\alpha_{1} \mathrm{~T}^{\mathrm{n}-\mathrm{k}}(\mathrm{v})
$$

But $\left\{\mathrm{T}^{\mathrm{n}-1}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{n}-\mathrm{k}}(\mathrm{v})\right\}$ is linearly independent (Theorem 7.37), and thus $\alpha_{k}=\cdots=\alpha_{1}=0$. This means that

$$
\mathrm{w}=\alpha_{\mathrm{n}} \mathrm{~T}^{\mathrm{n}-1}(\mathrm{v})+\cdots+\alpha_{\mathrm{k}+1} \mathrm{~T}^{\mathrm{k}}(\mathrm{v})=\mathrm{T}^{\mathrm{k}}\left(\mathrm{w}_{0}\right)
$$

where $\mathrm{w}_{0}=\alpha_{\mathrm{n}} \mathrm{T}^{\mathrm{n}-\mathrm{k}-1}(\mathrm{v})+\cdots+\alpha_{\mathrm{k}+1} \mathrm{v} \in \mathrm{W}$.
We are now in a position to prove our above assertion on the decomposition of V. This (by no means trivial) result will form the basis of the principle theorem dealing with nilpotent transformations (Theorem 7.41 below). It is worth pointing out that while the following theorem will be quite useful to us, its proof is not very constructive.

Theorem 7.40 Let T and W be as defined in the previous theorem. Then there exists a T-invariant subspace U of V such that $\mathrm{V}=\mathrm{W} \oplus \mathrm{U}$.

Proof Let $\mathrm{U} \subset \mathrm{V}$ be a T -invariant subspace of largest dimension with the property that $\mathrm{W} \cap \mathrm{U}=\{0\}$. (Such a space exists since even $\{0\}$ is T-invariant, and $\mathrm{W} \cap\{0\}=\{0\}$.) We first show that $\mathrm{V}=\mathrm{W}+\mathrm{U}$. If this is not the case, then there exists $z \in V$ such that $z \notin W+U$. Since $T^{0}(z)=z \notin W+U$ while $\mathrm{T}^{\mathrm{n}}(\mathrm{z})=0 \in \mathrm{~W}+\mathrm{U}$, it follows that there must exist an integer k with $0<\mathrm{k} \leq \mathrm{n}$ such that $\mathrm{T}^{\mathrm{k}}(\mathrm{z}) \in \mathrm{W}+\mathrm{U}$ and $\mathrm{T}^{\mathrm{j}}(\mathrm{z}) \notin \mathrm{W}+\mathrm{U}$ for $\mathrm{j}<\mathrm{k}$. We write $\mathrm{T}^{\mathrm{k}}(\mathrm{z})=\mathrm{w}+$ $u$ where $w \in W$ and $u \in U$, and therefore

$$
0=\mathrm{T}^{\mathrm{n}}(\mathrm{z})=\mathrm{T}^{\mathrm{n}-\mathrm{k}}\left(\mathrm{~T}^{\mathrm{k}}(\mathrm{z})\right)=\mathrm{T}^{\mathrm{n}-\mathrm{k}}(\mathrm{w})+\mathrm{T}^{\mathrm{n}-\mathrm{k}}(\mathrm{u}) .
$$

Since both W and U are T -invariant, we have $\mathrm{T}^{\mathrm{n}-\mathrm{k}}(\mathrm{w}) \in \mathrm{W}$ and $\mathrm{T}^{\mathrm{n}-\mathrm{k}}(\mathrm{u}) \in \mathrm{U}$. But $W \cap U=\{0\}$ so that

$$
\mathrm{T}^{\mathrm{n}-\mathrm{k}}(\mathrm{w})=-\mathrm{T}^{\mathrm{n}-\mathrm{k}}(\mathrm{u}) \in \mathrm{W} \cap \mathrm{U}=0
$$

(Remember that $W$ and $U$ are subspaces so $x \in U$ implies that $-x \in U$ also.) We now apply Theorem 7.39 to conclude that there exists $\mathrm{w}_{0} \in \mathrm{~W}$ such that $\mathrm{T}^{\mathrm{k}}\left(\mathrm{w}_{0}\right)=\mathrm{w}$, and hence $\mathrm{T}^{\mathrm{k}}(\mathrm{z})=\mathrm{w}+\mathrm{u}=\mathrm{T}^{\mathrm{k}}\left(\mathrm{w}_{0}\right)+\mathrm{u}$. Defining $\mathrm{x}=\mathrm{z}-\mathrm{w}_{0}$, we then have

$$
\mathrm{T}^{\mathrm{k}}(\mathrm{x})=\mathrm{T}^{\mathrm{k}}(\mathrm{z})-\mathrm{T}^{\mathrm{k}}\left(\mathrm{w}_{0}\right)=\mathrm{u} \in \mathrm{U} .
$$

But $U$ is $T$-invariant, and hence it follows that $T^{m}(x) \in U$ for any $m \geq k$. Considering lower powers of T , let us assume that $\mathrm{j}<\mathrm{k}$. Then the T invariance of W implies $\mathrm{T}^{\mathrm{j}}\left(\mathrm{w}_{0}\right) \in \mathrm{W}$, while we saw above that $\mathrm{T}^{\mathrm{j}}(\mathrm{z}) \notin \mathrm{W}+$ U . This means that

$$
\mathrm{T}^{\mathrm{j}}(\mathrm{x})=\mathrm{T}^{\mathrm{j}}(\mathrm{z})-\mathrm{T}^{\mathrm{j}}\left(\mathrm{w}_{0}\right) \notin \mathrm{U}
$$

(because if $\mathrm{T}^{\mathrm{j}}(\mathrm{x}) \in \mathrm{U}$, then $\mathrm{T}^{\mathrm{j}}(\mathrm{z})=\mathrm{T}^{\mathrm{j}}\left(\mathrm{w}_{0}\right)+\mathrm{T}^{\mathrm{j}}(\mathrm{x}) \in \mathrm{W}+\mathrm{U}$, a contradiction)
Now let $U_{x}$ be that subspace of $V$ spanned by $U$ together with the set $\left\{T^{k-1}(x), \ldots, T(x), x\right\}$. Since $U$ is $T$-invariant and $T^{j}(x) \notin U$, it must be true that $x \notin U$. Together with $U \subset U_{x}$, this means $\operatorname{dim} U_{x}>\operatorname{dim} U$. Applying $T$ to $\left\{\mathrm{T}^{\mathrm{k}-1}(\mathrm{x}), \ldots, \mathrm{T}(\mathrm{x}), \mathrm{x}\right\}$ we obtain the set $\left\{\mathrm{T}^{\mathrm{k}}(\mathrm{x}), \mathrm{T}^{\mathrm{k}-1}(\mathrm{x}), \ldots, \mathrm{T}^{2}(\mathrm{x}), \mathrm{T}(\mathrm{x})\right\}$. Since $T^{k}(x) \in U$ and the rest of the vectors in this set are included in the set that spans $U_{x}$, it follows that $U_{x}$ is also T-invariant.

By assumption, U is the subspace of largest dimension that is both T invariant and satisfies $\mathrm{W} \cap \mathrm{U}=\{0\}$. Since $\operatorname{dim} \mathrm{U}_{\mathrm{x}}>\operatorname{dim} \mathrm{U}$ and $\mathrm{U}_{\mathrm{x}}$ is $T$ invariant, we must have $\mathrm{W} \cap \mathrm{U}_{\mathrm{x}} \neq\{0\}$. Therefore there exists a nonzero element in $\mathrm{W} \cap \mathrm{U}_{\mathrm{x}}$ of the form $\mathrm{u}_{0}+\alpha_{\mathrm{k}} \mathrm{T}^{\mathrm{k}-1}(\mathrm{x})+\cdots+\alpha_{2} \mathrm{~T}(\mathrm{x})+\alpha_{1} \mathrm{x}$ where $\mathrm{u}_{0} \in \mathrm{U}$. We can not have $\alpha_{i}=0$ for every $\mathrm{i}=1, \ldots, \mathrm{k}$ because this would imply that $0 \neq u_{0} \in W \cap U=\{0\}$, a contradiction. If we let $\alpha_{r} \neq 0$ be the first nonzero $\alpha_{i}$, then we have

$$
\begin{equation*}
u_{0}+\left(\alpha_{k} T^{k-r}+\cdots+\alpha_{r+1} T+\alpha_{r}\right) T^{r-1}(x) \in W \tag{*}
\end{equation*}
$$

From Theorem 7.38 we see that $\alpha_{k} \mathrm{~T}^{\mathrm{k}-\mathrm{r}}+\cdots+\alpha_{\mathrm{r}+1} \mathrm{~T}+\alpha_{\mathrm{r}}$ is invertible, and its inverse is given by some polynomial $\mathrm{p}(\mathrm{T})$. Since W and U are T-invariant, they are also invariant under $\mathrm{p}(\mathrm{T})$.

Applying $\mathrm{p}(\mathrm{T})$ to $(*)$, we see that

$$
\mathrm{p}(\mathrm{~T})\left(\mathrm{u}_{0}\right)+\mathrm{T}^{\mathrm{r}-1}(\mathrm{x}) \in \mathrm{p}(\mathrm{~T})(\mathrm{W}) \subset \mathrm{W} .
$$

This means that $\mathrm{T}^{\mathrm{r}-1}(\mathrm{x}) \in \mathrm{W}+\mathrm{p}(\mathrm{T})(\mathrm{U}) \subset \mathrm{W}+\mathrm{U}$. But $\mathrm{r}-1<\mathrm{r}<\mathrm{k}$, and hence this result contradicts the earlier conclusion that $T^{j}(x) \notin U$ for $j<k$. Since this contradiction arose from the assumed existence of an element $\mathrm{z} \in \mathrm{V}$ with $\mathrm{z} \notin \mathrm{W}+\mathrm{U}$, we conclude that $\mathrm{V}=\mathrm{W}+\mathrm{U}$. Finally, since $\mathrm{W} \cap \mathrm{U}=\{0\}$ by hypothesis, we have $\mathrm{V}=\mathrm{W} \oplus \mathrm{U}$.

Combining several previous results, the next major theorem follows quite easily.

Theorem 7.41(a) Let $T \in L(V)$ be nilpotent with index of nilpotence $n_{1}$, and let $\mathrm{M}_{\mathrm{k}}$ be the $\mathrm{kx} k$ matrix containing all 0 's except for 1 's on the superdiagonal (see Theorem 7.37). Then there exists a basis for V in which the matrix of T has the block diagonal form

$$
\left(\begin{array}{cccc}
M_{n_{1}} & 0 & \cdots & 0 \\
0 & M_{n_{2}} & \cdots & 0 \\
\vdots & \vdots & & \vdots \\
0 & 0 & \cdots & M_{n_{r}}
\end{array}\right)
$$

where $n_{1} \geq \cdots \geq n_{r}$ and $n_{1}+\cdots+n_{r}=\operatorname{dim} V$.
Proof Since $\mathrm{T}^{\mathrm{n}_{1}}=0$ but $\mathrm{T}^{\mathrm{n}_{1}-1} \neq 0$, there exists $\mathrm{v} \in \mathrm{V}$ such that $\mathrm{T}^{\mathrm{n}_{1}-1}(\mathrm{v}) \neq 0$. Applying Theorem 7.37, we see that the vectors $\mathrm{v}_{1}=\mathrm{T}^{\mathrm{n}_{1}-1}(\mathrm{v}), \ldots, \mathrm{v}_{\mathrm{n}_{1}-1}=$ $\mathrm{T}(\mathrm{v}), \mathrm{v}_{\mathrm{n}_{1}}=\mathrm{v}$ are linearly independent and form the basis for a T -invariant subspace $W_{1}$ of $V$. Moreover, the matrix of $T_{1}=T \mid W_{1}$ in this basis is just $M_{n_{1}}$.

By Theorem 7.40, there exists a T-invariant subspace $\mathrm{U} \subset \mathrm{V}$ such that $\mathrm{V}=$ $\mathrm{W}_{1} \oplus \mathrm{U}$. Define a basis for V by taking the basis $\left\{\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{n}_{1}}\right\}$ for $\mathrm{W}_{1}$ together with any basis for U . Then, according to Theorem 7.20, the matrix of T with respect to this basis is of the form

$$
\left(\begin{array}{cc}
M_{n_{1}} & 0 \\
0 & A_{2}
\end{array}\right)
$$

where $A_{2}$ is the matrix of $T_{2}=T \mid U$. For any $u \in U$ and positive integer $m$ we have $T_{2}{ }^{m}(u)=T^{m}(u)$. Since $T^{n_{1}}=0$, we see that $T_{2}{ }^{m}=0$ for all $m \geq n_{1}$, and thus there exists an integer $\mathrm{n}_{2} \leq \mathrm{n}_{1}$ such that $\mathrm{T}_{2}{ }^{\mathrm{n}_{2}}=0$. This shows that $\mathrm{T}_{2}$ is nilpotent with index $\mathrm{n}_{2}$.

We now repeat the above argument using $\mathrm{T}_{2}$ and U instead of T and V . This time we will decompose $A_{2}$ into

$$
\left(\begin{array}{cc}
M_{n_{2}} & 0 \\
0 & A_{3}
\end{array}\right)
$$

and therefore the representation of T becomes

$$
\left(\begin{array}{ccc}
M_{n_{1}} & 0 & 0 \\
0 & M_{n_{2}} & 0 \\
0 & 0 & A_{3}
\end{array}\right)
$$

Continuing this process, it should be clear that we will eventually arrive at a basis for V in which the matrix of T has the desired form. It is also obvious that $\sum_{\mathrm{i}=1}^{\mathrm{r}} \mathrm{n}_{\mathrm{i}}=\operatorname{dim} \mathrm{V}=\mathrm{n}$ since the matrix of T must be of size n .

Our next result is a rephrasing of Theorem 7.41(a).
Theorem 7.41(b) Let $T \in L(V)$ be nilpotent with index $k$. Then there exists a basis for V in which the matrix representation of T is block diagonal, and where each of these diagonal entries (i.e., square matrices) is of the superdiagonal form M given in Theorem 7.37. Moreover,
(a) There is at least one M matrix of size $k$, and every other $M$ matrix is of size $\leq k$.
(b) The total number of M matrices in the representation of T (i.e., the total number of blocks in the representation of $T$ ) is just nul $T=\operatorname{dim}($ Ker $T)$.

## Proof See Exercise 7.11.1.

Example 7.13 Consider the matrix

$$
A=\left(\begin{array}{lllll}
0 & 1 & 1 & 0 & 1 \\
0 & 0 & 1 & 1 & 1 \\
0 & 0 & 0 & 0 & 0 \\
0 & 0 & 0 & 0 & 0 \\
0 & 0 & 0 & 0 & 0
\end{array}\right)
$$

We leave it to the reader to show that $A$ is nilpotent with index 3 , i.e., $A^{3}=0$ but $A^{2} \neq 0$. We seek the diagonal representation of A described in Theorem 7.41(b). It is obvious that $\mathrm{r}(\mathrm{A})=2$, and therefore (using Theorem 5.6) nul $\mathrm{A}=$ $5-2=3$. Thus there are three M matrices in the diagonal representation of A , and one of them must be of size 3 . This means that the only possibility for the remaining two matrices is that they both be of size 1 . Thus the block diagonal form for A must be

$$
A=\left(\begin{array}{ccc}
\begin{array}{|ccc|}
0 & 1 & 0 \\
0 & 0 & 1 \\
0 & 0 & 0 \\
\hline
\end{array} & & \\
& & \boxed{0} \\
& & \\
& & \\
0
\end{array}\right)
$$

It is easy to see that this matrix is also nilpotent with index 3. //

## Exercises

1. Prove Theorem 7.41 (b). [Hint: What is the rank of the matrix in Theorem 7.41(a)?]
2. Suppose $S, T \in L(V)$ are nilpotent with the property that $S T=T S$. Show that $\mathrm{S}+\mathrm{T}$ and ST are also nilpotent.
3. Suppose $A$ is a supertriangular matrix, i.e., all entries of $A$ on or below the main diagonal are zero. Show that A is nilpotent.
4. Let $\mathrm{V}_{\mathrm{n}}$ be the vector space of all polynomials of degree $\leq \mathrm{n}$, and let $\mathrm{D} \in$ $\mathrm{L}\left(\mathrm{V}_{\mathrm{n}}\right)$ be the usual differentiation operator. Show that D is nilpotent with index $n+1$.
5. Show that the following nilpotent matrices of size n are similar:

$$
\left(\begin{array}{ccccc}
0 & 1 & 0 & \cdots & 0 \\
0 & 0 & 1 & \cdots & 0 \\
\vdots & \vdots & \vdots & & \vdots \\
0 & 0 & 0 & \cdots & 1 \\
0 & 0 & 0 & \cdots & 0
\end{array}\right) \quad \text { and } \quad\left(\begin{array}{ccccc}
0 & 0 & \cdots & 0 & 0 \\
1 & 0 & \cdots & 0 & 0 \\
0 & 1 & \cdots & 0 & 0 \\
\vdots & \vdots & & \vdots & \vdots \\
0 & 0 & \cdots & 1 & 0
\end{array}\right)
$$

6. Show that two nilpotent $3 \times 3$ matrices are similar if and only if they have the same index of nilpotency. Give an example to show that this is not true for nilpotent $4 \times 4$ matrices.

### 7.12 THE TRIANGULAR FORM THEOREM AGAIN *

After all the discussion on nilpotent transformations in the previous section, let us return to our second version of the triangular form theorem which, as we shall see in the next chapter, is just the Jordan canonical form. While this theorem applies to a finite-dimensional vector space over an arbitrary field, the minimal polynomial $\mathrm{m}(\mathrm{x})$ for T must be factorable into linear polynomials. This means that all the roots of $\mathrm{m}(\mathrm{x})$ must lie in $\mathcal{F}$. Clearly, this will always be true if $\mathcal{F}$ is algebraically closed.

Theorem 7.42 (Jordan Form) Suppose $T \in L(V)$, and assume that the minimal polynomial $\mathrm{m}(\mathrm{x})$ for T can be written in the form

$$
\mathrm{m}(\mathrm{x})=\left(\mathrm{x}-\lambda_{\mathrm{t}}\right)^{\mathrm{n}_{1}} \cdots\left(\mathrm{x}-\lambda_{\mathrm{r}}\right)^{\mathrm{n}_{\mathrm{r}}}
$$

where each $n_{i}$ is a positive integer, and the $\lambda_{i}$ are distinct elements of $\mathcal{F}$. Then there exists a basis for V in which the matrix representation A of T has the block diagonal form $\mathrm{A}=\mathrm{A}_{1} \oplus \cdots \oplus \mathrm{~A}_{\mathrm{r}}$ where each $\mathrm{A}_{\mathrm{i}} \in \mathrm{M}_{\mathrm{k}_{\mathrm{i}}}(\mathcal{F})$ for some integer $k_{i} \geq n_{i}$, and each $A_{i}$ has the upper triangular form

$$
\left(\begin{array}{cccccc}
\lambda_{i} & * & 0 & \cdots & 0 & 0 \\
0 & \lambda_{i} & * & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & \cdots & \lambda_{i} & * \\
0 & 0 & 0 & \cdots & 0 & \lambda_{i}
\end{array}\right)
$$

where the ${ }^{*}$ 's may be either 0 or 1 .
Proof For each $\mathrm{i}=1, \ldots, r$ we define $\mathrm{W}_{\mathrm{i}}=\operatorname{Ker}\left(\mathrm{T}-\lambda_{\mathrm{i}} 1\right)^{\mathrm{n}_{\mathrm{i}}}$ and $\mathrm{k}_{\mathrm{i}}=\operatorname{dim} \mathrm{W}_{\mathrm{i}}$. By the primary decomposition theorem (Theorem 7.23), we see that $\mathrm{V}=\mathrm{W}_{1} \oplus$ $\cdots \oplus \mathrm{W}_{\mathrm{r}}$ (where each $\mathrm{W}_{\mathrm{i}}$ is T-invariant), and hence according to Theorem 2.15, V has a basis which is just the union of bases of the $\mathrm{W}_{\mathrm{i}}$. Letting the basis for V be the union of the bases for $\mathrm{W}_{1}, \ldots, \mathrm{~W}_{\mathrm{r}}$ taken in this order, it follows from Theorem 7.20 that $\mathrm{A}=\mathrm{A}_{1} \oplus \cdots \oplus \mathrm{~A}_{\mathrm{r}}$ where each $\mathrm{A}_{\mathrm{i}}$ is the matrix representation of $T_{i}=T \mid W_{i}$. We must show that each $A_{i}$ has the required form, and that $\mathrm{k}_{\mathrm{i}} \geq \mathrm{n}_{\mathrm{i}}$ for each $\mathrm{i}=1, \ldots, \mathrm{r}$.

If we define $\mathrm{N}_{\mathrm{i}}=\mathrm{T}-\lambda_{\mathrm{i}} 1$, then $\mathrm{N}_{\mathrm{i}} \in \mathrm{L}\left(\mathrm{W}_{\mathrm{i}}\right)$ since $\mathrm{W}_{\mathrm{i}}$ is T-invariant. In other words, $\mathrm{N}_{\mathrm{i}}$ is a linear operator defined on the space $\mathrm{W}_{\mathrm{i}}$, and hence so is $N_{i}{ }^{n_{i}}$. However, since $W_{i}=\operatorname{Ker} N_{i}{ }^{n_{i}}$, it follows from the definition of kernel that $\mathrm{N}_{\mathrm{i}}{ }^{\mathrm{n}_{\mathrm{i}}}=0$ so that $\mathrm{N}_{\mathrm{i}}$ is nilpotent. The result now follows by applying Theorem 7.41(a) to each $\mathrm{N}_{\mathrm{i}}$ and writing $\mathrm{T}=\mathrm{N}_{\mathrm{i}}+\lambda_{\mathrm{i}} 1$.

Note that each $A_{i}$ in this theorem is a direct sum of matrices of the form

$$
\left(\begin{array}{cccccc}
\lambda_{i} & 1 & 0 & \cdots & 0 & 0 \\
0 & \lambda_{i} & 1 & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & \cdots & \lambda_{i} & 1 \\
0 & 0 & 0 & \cdots & 0 & \lambda_{i}
\end{array}\right)
$$

which are referred to as basic Jordan blocks belonging to $\lambda_{i}$. This theorem will be discussed in much more detail (and from an entirely different point of view) in Chapter 8.

Many of our earlier results follow in a natural manner as corollaries to Theorem 7.42. In particular, suppose that V is finite-dimensional over an algebraically closed field $\mathcal{F}$, and $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ satisfies the hypotheses of Theorem 7.42. We wish to know the form of the characteristic polynomial $\Delta_{\mathrm{T}}(\mathrm{x})$. Relative to the basis for V given by Theorem 7.42, we see that the characteristic matrix $\mathrm{xI}-\mathrm{A}$ is given by

and hence (using Theorem 4.5)

$$
\Delta_{T}(x)=\operatorname{det}(x I-A)=\prod_{i=1}^{r}\left(x-\lambda_{i}\right)^{k_{i}}
$$

On the other hand, since $m(x)=\prod_{i=1}^{r}\left(x-\lambda_{i}\right)^{n_{i}}$ and $k_{i} \geq n_{i}$, properties (a) and (b) in the following corollary should be clear, and property (c) follows from the proof of Theorem 7.42. Note that property (c) is just the Cayley-Hamilton theorem again.

Corollary Suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ where V is finite-dimensional over an algebraically closed field $\mathcal{F}$, and let the minimal and characteristic polynomials of T be $\mathrm{m}(\mathrm{x})$ and $\Delta_{\mathrm{T}}(\mathrm{x})$ respectively. Then
(a) $\mathrm{m}(\mathrm{x}) \mid \Delta_{\mathrm{T}}(\mathrm{x})$.
(b) $\mathrm{m}(\mathrm{x})$ and $\Delta_{\mathrm{T}}(\mathrm{x})$ have the same roots (although they are not necessarily of the same algebraic multiplicity).
(c) $\Delta_{\mathrm{T}}(\mathrm{T})=0$.

Example 7.14 Let $\mathrm{V}=\mathbb{C}^{3}$ have basis $\left\{\mathrm{v}_{1}, \mathrm{v}_{2}, \mathrm{v}_{3}\right\}$ and define $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ by

$$
\begin{aligned}
& T\left(v_{1}\right)=-v_{1}+2 v_{3} \\
& T\left(v_{2}\right)=3 v_{1}+2 v_{2}+v_{3} \\
& T\left(v_{3}\right)=-v_{3} .
\end{aligned}
$$

Then the matrix representation of T in this basis is

$$
A=\left(\begin{array}{rrr}
-1 & 3 & 0 \\
0 & 2 & 0 \\
2 & 1 & -1
\end{array}\right)
$$

We first find the minimal polynomial for T. Note that while we have given many theorems dealing with the minimal polynomial, there has as yet been no general method presented for actually finding it. (We shall see that such a method can be based on Theorem 8.8.) Since the minimal polynomial has the same irreducible factors as does the characteristic polynomial (Theorem 7.12), we begin by finding $\Delta_{T}(x)=\operatorname{det}(x I-A)$. A simple calculation yields

$$
\Delta_{\mathrm{T}}(\mathrm{x})=(\mathrm{x}+1)^{2}(\mathrm{x}-2)
$$

and therefore the minimal polynomial must be either

$$
(x+1)^{2}(x-2)
$$

or

$$
(x+1)(x-2) .
$$

To decide between these two possibilities, we could simply substitute A and multiply them out. However, it is worthwhile to instead apply Theorem 7.23. In other words, we find the subspaces $W_{i}=\operatorname{Ker} f_{i}(x)^{n_{i}}$ and see which value of $n_{1}$ (i.e., either 1 or 2 ) results in $V=W_{1} \oplus W_{2}$. We must therefore find the kernel (i.e., the null space) of $(\mathrm{T}+1),(\mathrm{T}+1)^{2}$ and $(\mathrm{T}-2)$. Applying the operator $\mathrm{T}+1$ to each of the basis vectors yields

$$
\begin{aligned}
& (T+1)\left(v_{1}\right)=2 v_{3} \\
& (T+1)\left(v_{2}\right)=3 v_{1}+3 v_{2}+v_{3} \\
& (T+1)\left(v_{3}\right)=0 .
\end{aligned}
$$

Since $\operatorname{Im}(T+1)$ is spanned by these three vectors, only two of which are obviously independent, we see that $\mathrm{r}(\mathrm{T}+1)=2$. Therefore, applying Theorem 5.6, we find that $\operatorname{nul}(T+1)=\operatorname{dim} V-r(T+1)=1$. Similarly, we have

$$
\begin{aligned}
& (T+1)^{2}\left(v_{1}\right)=0 \\
& (T+1)^{2}\left(v_{2}\right)=9\left(v_{1}+v_{2}+v_{3}\right) \\
& (T+1)^{2}\left(v_{3}\right)=0
\end{aligned}
$$

and so $\mathrm{r}(\mathrm{T}+1)^{2}=1$ implies $\operatorname{nul}(\mathrm{T}+1)^{2}=2$. It should also be clear that the space $\operatorname{Ker}(T+1)^{2}$ is spanned by the set $\left\{\mathrm{v}_{1}, \mathrm{v}_{3}\right\}$. Finally,

$$
\begin{aligned}
& (T-2)\left(v_{1}\right)=-3 v_{1}+2 v_{3} \\
& (T-2)\left(v_{2}\right)=3 v_{1}+v_{3} \\
& (T-2)\left(v_{3}\right)=-3 v_{3}
\end{aligned}
$$

so that $r(T-2)=2$ and hence $\operatorname{nul}(T-2)=1$. We also note that since

$$
(T-2)\left(v_{1}+v_{2}+v_{3}\right)=0
$$

and $\operatorname{nul}(T-2)=1$, it follows that the space $\mathrm{W}_{2}=\operatorname{Ker}(\mathrm{T}-2)$ must be spanned by the vector $\left\{\mathrm{v}_{1}+\mathrm{v}_{2}+\mathrm{v}_{3}\right\}$. Alternatively, we could assume that $\mathrm{W}_{2}$ is spanned by some vector $u=\alpha_{1} v_{1}+\alpha_{2} v_{2}+\alpha_{3} v_{3}$ and proceed to find $\alpha_{1}, \alpha_{2}$ and $\alpha_{3}$ by requiring that $(T-2)(u)=0$. This results in

$$
(\mathrm{T}-2)(\mathrm{u})=\alpha_{1}\left(-3 \mathrm{v}_{1}+2 \mathrm{v}_{3}\right)+\alpha_{2}\left(3 \mathrm{v}_{1}+\mathrm{v}_{3}\right)+\alpha_{3}\left(-3 \mathrm{v}_{3}\right)=0
$$

so that we have the simultaneous set of equations

$$
\begin{array}{r}
-3 \alpha_{1}+3 \alpha_{2}=0 \\
2 \alpha_{1}+\alpha_{2}-3 \alpha_{3}=0 .
\end{array}
$$

This yields $\alpha_{1}=\alpha_{2}=\alpha_{3}$ so that $u=v_{1}+v_{2}+v_{3}$ will span $W_{2}$ as we found by inspection.

From the corollary to Theorem 2.15 we have $\operatorname{dim} \mathrm{V}=\operatorname{dim} \mathrm{W}_{1}+\operatorname{dim} \mathrm{W}_{2}$, and since $\operatorname{dim} \mathrm{W}_{2}=\operatorname{nul}(\mathrm{T}-2)=1$, it follows that we must have $\operatorname{dim} \mathrm{W}_{1}=2$, and hence $W_{1}=\operatorname{Ker}(T+1)^{2}$. Thus the minimal polynomial for $T$ must be given by

$$
\mathrm{m}(\mathrm{x})=(\mathrm{x}+1)^{2}(\mathrm{x}-2)
$$

Note that because of this form for $\mathrm{m}(\mathrm{x})$, Theorem 7.24 tells us that T is not diagonalizable.

According to Theorem 7.42, the matrix $\mathrm{A}_{1}$ corresponding to $\lambda=-1$ must be at least a $2 \times 2$ matrix, and the matrix $\mathrm{A}_{2}$ corresponding to $\lambda=2$ must be at least a $1 \times 1$ matrix. However, since $\operatorname{dim} V=3$, these are in fact the actual
sizes required. While $A_{2}$ is unambiguous, the matrix $A_{1}$ could be either a single $2 \times 2$ matrix, or it could be a direct sum of two $1 \times 1$ matrices. To resolve this we use Theorem 7.41(b) which tells us that the number of blocks in the representation of the nilpotent operator $\mathrm{T}+1$ is $\operatorname{dim}(\operatorname{Ker}(\mathrm{T}+1))=1$. This means that the Jordan form of T must be

$$
\left(\begin{array}{rrr}
-1 & 1 & 0 \\
0 & -1 & 0 \\
0 & 0 & 2
\end{array}\right)
$$

## Exercises

1. Let $V=\mathbb{C}^{4}$ and suppose $T \in L(V)$. If $T$ has only one eigenvalue $\lambda$ of multiplicity 4, describe all possible Jordan forms of T.
2. Let $\mathrm{V}=\mathbb{C}^{4}$ and suppose $\mathrm{m}(\mathrm{x})=\left(\mathrm{x}-\lambda_{1}\right)^{\mathrm{n}_{1}} \cdots\left(\mathrm{x}-\lambda_{\mathrm{r}}\right)^{\mathrm{n}_{\mathrm{r}}}$ is the minimal polynomial for $\mathrm{T} \in \mathrm{L}(\mathrm{V})$. Let $\mathrm{A}=\mathrm{A}_{1} \oplus \cdots \oplus \mathrm{~A}_{\mathrm{r}}$ be the Jordan form of T . Prove directly from the structure of the $\mathrm{A}_{\mathrm{i}}$ that the largest Jordan block belonging to $\lambda_{i}$ has size $n_{i} \times n_{i}$.
3. Let $V=\mathbb{C}^{\mathrm{n}}$. If the Jordan form of $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ consists of just one Jordan block (counting $1 \times 1$ blocks), what is the Jordan form of $\mathrm{T}^{2}$ ? Explain.
4. Let $V=\mathbb{C}^{n}$, suppose $T \in L(V)$, and let $\lambda$ be an eigenvalue of $T$. What is the relationship between the number of Jordan blocks belonging to $\lambda$ and the rank of $\mathrm{T}-\lambda \mathrm{I}$ ? Explain.
5. Let $\mathrm{V}=\mathbb{C}^{\mathrm{n}}$, and suppose that each matrix below represents $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ relative to the standard basis. Determine the Jordan form of T. [Hint: Use the previous exercise.]
(a) $\left(\begin{array}{rrrr}1 & -1 & -1 & -1 \\ 0 & 1 & -1 & -1 \\ 0 & 0 & 2 & 2 \\ 0 & 0 & 0 & 2\end{array}\right)$
(b) $\left(\begin{array}{rrrr}1 & -1 & -1 & -1 \\ 0 & 1 & -1 & -1 \\ 0 & 0 & 2 & 0 \\ 0 & 0 & 0 & 2\end{array}\right)$
(c) $\left(\begin{array}{rrrr}1 & -1 & -1 & -1 \\ -1 & 1 & -1 & -1 \\ 0 & 0 & 2 & 2 \\ 0 & 0 & 0 & 2\end{array}\right)$
(d) $\left(\begin{array}{rrrr}1 & -1 & -1 & -1 \\ -1 & 1 & -1 & -1 \\ 0 & 0 & 2 & 0 \\ 0 & 0 & 0 & 2\end{array}\right)$

## CHAPTER 8

## Canonical Forms

Recall that at the beginning of Section 7.5 we stated that a canonical form for $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is simply a representation in which the matrix takes on an especially simple form. For example, if there exists a basis of eigenvectors of T, then the matrix representation will be diagonal. In this case, it is then quite trivial to assess the various properties of T such as its rank, determinant and eigenvalues. Unfortunately, while this is generally the most desirable form for a matrix representation, it is also generally impossible to achieve.

We now wish to determine certain properties of T that will allow us to learn as much as we can about the possible forms its matrix representation can take. There are three major canonical forms that we will consider in this chapter : triangular, rational and Jordan. (This does not count the Smith form, which is really a tool, used to find the rational and Jordan forms.) As we have done before, our approach will be to study each of these forms in more than one way. By so doing, we shall gain much insight into their meaning, as well as learning additional techniques that are of great use in various branches of mathematics.

### 8.1 ELEMENTARY CANONICAL FORMS

In order to ease into the subject, this section presents a simple and direct method of treating two important results: the triangular form for complex matrices and the diagonalization of normal matrices. To begin with, suppose
that we have a matrix $A \in M_{n}(\mathbb{C})$. We define the adjoint (or Hermitian adjoint) of $A$ to be the matrix $\mathrm{A}^{\dagger}=\mathrm{A}^{* \mathrm{~T}}$. In other words, the adjoint of A is its complex conjugate transpose. From Theorem 3.18(d), it is easy to see that

$$
(\mathrm{AB})^{\dagger}=\mathrm{B}^{\dagger} \mathrm{A}^{\dagger} .
$$

If it so happens that $\mathrm{A}^{\dagger}=\mathrm{A}$, then A is said to be a Hermitian matrix.
If a matrix $U \in M_{n}(\mathbb{C})$ has the property that $U^{\dagger}=U^{-1}$, then we say that $U$ is unitary. Thus a matrix $U$ is unitary if $U^{\dagger}=U^{\dagger} U=I$. (Note that by Theorem 3.21, it is only necessary to require either $\mathrm{UU}^{\dagger}=\mathrm{I}$ or $\mathrm{U}^{\dagger} \mathrm{U}=\mathrm{I}$.) We also see that the product of two unitary matrices U and V is unitary since $(\mathrm{UV})^{\dagger} \mathrm{UV}=\mathrm{V}^{\dagger} \mathrm{U}^{\dagger} \mathrm{UV}=\mathrm{V}^{\dagger} \mathrm{IV}=\mathrm{V}^{\dagger} \mathrm{V}=\mathrm{I}$. If a matrix $\mathrm{N} \in \mathrm{M}_{\mathrm{n}}(\mathbb{C})$ has the property that it commutes with its adjoint, i.e., $\mathrm{NN}^{\dagger}=\mathrm{N}^{\dagger} \mathrm{N}$, then N is said to be a normal matrix. Note that Hermitian and unitary matrices are automatically normal.

Example 8.1 Consider the matrix $A \in M_{2}(\mathbb{C})$ given by

$$
A=\frac{1}{\sqrt{2}}\left(\begin{array}{rr}
1 & -1 \\
i & i
\end{array}\right) .
$$

Then the adjoint of A is given by

$$
A^{\dagger}=\frac{1}{\sqrt{2}}\left(\begin{array}{rr}
1 & -i \\
-1 & -i
\end{array}\right)
$$

and we leave it to the reader to verify that $\mathrm{AA}^{\dagger}=\mathrm{A}^{\dagger} \mathrm{A}=1$, and hence show that A is unitary. //

We will devote considerable time in Chapter 10 to the study of these matrices. However, for our present purposes, we wish to point out one important property of unitary matrices. Note that since $U \in M_{n}(\mathbb{C})$, the rows $U_{i}$ and columns $U^{\mathrm{i}}$ of U are just vectors in $\mathbb{C}^{\mathrm{n}}$. This means that we can take their inner product relative to the standard inner product on $\mathbb{C}^{\mathrm{n}}$ (see Example 2.9). Writing out the relation $\mathrm{UU}^{\dagger}=\mathrm{I}$ in terms of components, we have

$$
\left(U U^{\dagger}\right)_{i j}=\sum_{k=1}^{n} u_{i k} u_{k j}^{\dagger}=\sum_{k=1}^{n} u_{i k} u_{j k}^{*}=\sum_{k=1}^{n} u_{j k}^{*} u_{i k}=\left\langle U_{j}, U_{i}\right\rangle=\delta_{i j}
$$

and from $\mathrm{U}^{\dagger} \mathrm{U}=\mathrm{I}$ we see that

$$
\left(U^{\dagger} U\right)_{i j}=\sum_{k=1}^{n} u^{\dagger}{ }_{i k} u_{k j}=\sum_{k=1}^{n} u_{k i}^{*} u_{k j}=\left\langle U^{i}, U^{j}\right\rangle=\delta_{i j} .
$$

In other words, a matrix is unitary if and only if its rows (or columns) each form an orthonormal set. Note we have shown that if the rows (columns) of $\mathrm{U} \in \mathrm{M}_{\mathrm{n}}(\mathbb{C})$ form an orthonormal set, then so do the columns (rows), and either of these is a sufficient condition for $U$ to be unitary. For example, the reader can easily verify that the matrix A in Example 8.1 satisfies these conditions.

It is also worth pointing out that Hermitian and unitary matrices have important analogues over the real number system. If $A \in M_{n}(\mathbb{R})$ is Hermitian, then $A=A^{\dagger}=A^{T}$, and we say that $A$ is symmetric. If $U \in M_{n}(\mathbb{R})$ is unitary, then $\mathrm{U}^{-1}=\mathrm{U}^{\dagger}=\mathrm{U}^{\mathrm{T}}$, and we say that U is orthogonal. Repeating the above calculations over $\mathbb{R}$, it is easy to see that a real matrix is orthogonal if and only if its rows (or columns) form an orthonormal set.

It will also be useful to recall from Section 3.6 that if A and B are two matrices for which the product $A B$ is defined, then the ith row of $A B$ is given by $(A B)_{i}=A_{i} B$ and the $i$ th column of $A B$ is given by $(A B)^{i}=A B^{i}$. We now prove yet another version of the triangular form theorem.

Theorem 8.1 (Schur Canonical Form) If $A \in M_{n}(\mathbb{C})$, then there exists a unitary matrix $U \in M_{n}(\mathbb{C})$ such that $U^{\dagger} A U$ is upper-triangular. Furthermore, the diagonal entries of $\mathrm{U}^{\dagger} \mathrm{AU}$ are just the eigenvalues of A .

Proof If $\mathrm{n}=1$ there is nothing to prove, so we assume that the theorem holds for any square matrix of size $n-1 \geq 1$, and suppose A is of size $n$. Since we are dealing with the algebraically closed field $\mathbb{C}$, we know that A has n (not necessarily distinct) eigenvalues (see Section 7.3). Let $\lambda$ be one of these eigenvalues, and denote the corresponding eigenvector by $\widetilde{\mathrm{U}}^{1}$. By Theorem 2.10 we extend $\widetilde{\mathrm{U}}^{1}$ to a basis for $\mathbb{C}^{\mathrm{n}}$, and by the Gram-Schmidt process (Theorem 2.21) we assume that this basis is orthonormal. From our discussion above, we see that this basis may be used as the columns of a unitary matrix $\widetilde{U}$ with $\widetilde{\mathrm{U}}^{1}$ as its first column. We then see that

$$
\begin{aligned}
\left(\tilde{U}^{\dagger} A \tilde{U}\right)^{1} & =\tilde{U}^{\dagger}(A \tilde{U})^{1}=\tilde{U}^{\dagger}\left(A \tilde{U}^{1}\right)=\tilde{U}^{\dagger}\left(\lambda \tilde{U}^{1}\right)=\lambda\left(\tilde{U}^{\dagger} \tilde{U}^{1}\right) \\
& =\lambda\left(\tilde{U}^{\dagger} \tilde{U}\right)^{1}=\lambda I^{1}
\end{aligned}
$$

and hence $\widetilde{\mathrm{U}}^{\dagger} \mathrm{A} \widetilde{\mathrm{U}}$ has the form

$$
\tilde{U}^{\dagger} A \tilde{U}=\left(\begin{array}{ccc}
\lambda & * & \ldots
\end{array}\right)
$$

where $\mathrm{B} \in \mathrm{M}_{\mathrm{n}-1}(\mathbb{C})$ and the *'s are (in general) nonzero scalars. By our induction hypothesis, we may choose a unitary matrix $\mathrm{W} \in \mathrm{M}_{\mathrm{n}-1}(\mathbb{C})$ such that $\mathrm{W}^{\dagger} \mathrm{BW}$ is upper-triangular. Let $\mathrm{V} \in \mathrm{M}_{\mathrm{n}}(\mathbb{C})$ be a unitary matrix of the form

$$
V=\left(\begin{array}{cccc}
1 & 0 & \cdots & 0 \\
0 & & & \\
\vdots & & W & \\
0 & & &
\end{array}\right)
$$

and define the unitary matrix $U=\widetilde{U} V \in M_{n}(\mathbb{C})$. Then

$$
\mathrm{U}^{\dagger} \mathrm{AU}=(\widetilde{\mathrm{U}} \mathrm{~V})^{\dagger} \mathrm{A}(\widetilde{\mathrm{U}} \mathrm{~V})=\mathrm{V}^{\dagger}\left(\widetilde{\mathrm{U}}^{\dagger} \mathrm{A} \widetilde{\mathrm{U}}\right) \mathrm{V}
$$

is upper-triangular since (in an obvious shorthand notation)

$$
\begin{aligned}
V^{\dagger}\left(\tilde{U}^{\dagger} A \tilde{U}\right) V & =\left(\begin{array}{cc}
1 & 0 \\
0 & W^{\dagger}
\end{array}\right)\left(\begin{array}{cc}
\lambda & * \\
0 & B
\end{array}\right)\left(\begin{array}{cc}
1 & 0 \\
0 & W
\end{array}\right)=\left(\begin{array}{cc}
1 & 0 \\
0 & W^{\dagger}
\end{array}\right)\left(\begin{array}{cc}
\lambda & * \\
0 & B W
\end{array}\right) \\
& =\left(\begin{array}{cc}
\lambda & * \\
0 & W^{\dagger} B W
\end{array}\right)
\end{aligned}
$$

and $\mathrm{W}^{\dagger} \mathrm{BW}$ is upper-triangular by the induction hypothesis.
Noting that $\lambda \mathrm{I}-\mathrm{U}^{\dagger} \mathrm{AU}$ is upper-triangular, it is easy to see (using Theorem 4.5) that the roots of $\operatorname{det}\left(\lambda I-U^{\dagger} A U\right)$ are just the diagonal entries of $\mathrm{U}^{\dagger} \mathrm{AU}$. But

$$
\operatorname{det}\left(\lambda \mathrm{I}-\mathrm{U}^{\dagger} \mathrm{AU}\right)=\operatorname{det}\left[\mathrm{U}^{\dagger}(\lambda \mathrm{I}-\mathrm{A}) \mathrm{U}\right]=\operatorname{det}(\lambda \mathrm{I}-\mathrm{A})
$$

so that A and $\mathrm{U}^{\dagger} \mathrm{AU}$ have the same eigenvalues.
Corollary If $A \in M_{n}(\mathbb{R})$ has all its eigenvalues in $\mathbb{R}$, then the matrix $U$ defined in Theorem 8.1 may be chosen to have all real entries.

Proof If $\lambda \in \mathbb{R}$ is an eigenvalue of $A$, then $A-\lambda I$ is a real matrix with determinant $\operatorname{det}(\mathrm{A}-\lambda \mathrm{I})=0$, and therefore the homogeneous system of equations $(A-\lambda I) X=0$ has a real solution. Defining $\widetilde{U}^{1}=X$, we may now proceed as in Theorem 8.1. The details are left to the reader (see Exercise 8.8.1).

We say that two matrices $A, B \in M_{n}(\mathbb{C})$ are unitarily similar (written A ~ B) if there exists a unitary matrix $U$ such that $B=U^{\dagger} A U=U^{-1} A U$. Since this
defines an equivalence relation on the set of all matrices in $M_{n}(\mathbb{C})$, many authors say that A and B are unitarily equivalent. However, we will be using the term "equivalent" in a somewhat more general context later in this chapter, and the word "similar" is in accord with our earlier terminology.

We leave it to the reader to show that if A and B are unitarily similar and A is normal, then B is also normal (see Exercise 8.8.2). In particular, suppose that U is unitary and N is such that $\mathrm{U}^{\dagger} \mathrm{NU}=\mathrm{D}$ is diagonal. Since any diagonal matrix is automatically normal, it follows that N must be normal also. We now show that the converse is also true, i.e., that any normal matrix is unitarily similar to a diagonal matrix.

To see this, suppose N is normal, and let $\mathrm{U}^{\dagger} \mathrm{NU}=\mathrm{D}$ be the Schur canonical form of N . Then D is both upper-triangular and normal (since it is unitarily similar to a normal matrix). We claim that the only such matrices are diagonal. For, consider the $(1,1)$ elements of $\mathrm{DD}^{\dagger}$ and $\mathrm{D}^{\dagger} \mathrm{D}$. From what we showed above, we have

$$
\left(D D^{\dagger}\right)_{11}=\left\langle D_{1}, D_{1}\right\rangle=\left|d_{11}\right|^{2}+\left|d_{12}\right|^{2}+\cdots+\left|d_{1 n}\right|^{2}
$$

and

$$
\left(D^{\dagger} D\right)_{11}=\left\langle D^{1}, D^{1}\right\rangle=\left|d_{11}\right|^{2}+\left|d_{21}\right|^{2}+\cdots+\left|d_{n 1}\right|^{2} .
$$

But $D$ is upper-triangular so that $d_{21}=\cdots=d_{n 1}=0$. By normality we must have $\left(\mathrm{DD}^{\dagger}\right)_{11}=\left(\mathrm{D}^{\dagger} \mathrm{D}\right)_{11}$, and therefore $\mathrm{d}_{12}=\cdots=d_{1 n}=0$ also. In other words, with the possible exception of the $(1,1)$ entry, all entries in the first row and column of D must be zero. In the same manner, we see that

$$
\left(\mathrm{DD}^{\dagger}\right)_{22}=\left\langle\mathrm{D}_{2}, \mathrm{D}_{2}\right\rangle=\left|\mathrm{d}_{21}\right|^{2}+\left|\mathrm{d}_{22}\right|^{2}+\cdots+\left|\mathrm{d}_{2 \mathrm{n}}\right|^{2}
$$

and

$$
\left(D^{\dagger} D\right)_{22}=\left\langle D^{2}, D^{2}\right\rangle=\left|d_{12}\right|^{2}+\left|d_{22}\right|^{2}+\cdots+\left|d_{n 2}\right|
$$

Since the fact that $D$ is upper-triangular means $d_{32}=\cdots=d_{n 2}=0$ and we just showed that $d_{21}=d_{12}=0$, it again follows by normality that $d_{23}=\cdots=d_{2 n}=$ 0 . Thus all entries in the second row and column with the possible exception of the $(2,2)$ entry must be zero.

Continuing this procedure, it is clear that D must be diagonal as claimed. In other words, an upper-triangular normal matrix is necessarily diagonal. This discussion proves the following very important theorem.

Theorem 8.2 A matrix $N \in M_{n}(\mathbb{C})$ is normal if and only if there exists a unitary matrix U such that $\mathrm{U}^{\dagger} \mathrm{NU}$ is diagonal.

Corollary If $A=\left(a_{i j}\right) \in M_{n}(\mathbb{R})$ is symmetric, then there exists an orthogonal matrix $S$ such that $S^{T} A S$ is diagonal.

Proof If we can show that a real symmetric matrix has all real eigenvalues, then this corollary will follow from the corollary to Theorem 8.1 and the real analogue of the proof of Theorem 8.2. Now suppose $A=A^{T}$ so that $a_{i j}=a_{j i}$. If $\lambda$ is an eigenvalue of $A$, then there exists a (nonzero and not necessarily real) vector $\mathrm{x} \in \mathbb{C}^{\mathrm{n}}$ such that $\mathrm{Ax}=\lambda \mathrm{x}$ or

$$
\begin{equation*}
\sum_{j=1}^{n} a_{i j} x_{j}=\lambda x_{i} \tag{1}
\end{equation*}
$$

Multiplying (1) by $\mathrm{x}_{\mathrm{i}}{ }^{*}$, summing over i and using the standard inner product on $\mathbb{C}^{\mathrm{n}}$ we obtain

$$
\begin{equation*}
\sum_{i, j=1}^{n} x_{i}^{*} a_{i j} x_{j}=\lambda\|x\|^{2} \tag{2}
\end{equation*}
$$

On the other hand, we may take the complex conjugate of (1), then multiply by $\mathrm{x}_{\mathrm{i}}$ and sum over i to obtain (since each $\mathrm{a}_{\mathrm{ij}}$ is real)

$$
\begin{equation*}
\sum_{i, j=1}^{n} x_{i} a_{i j} x_{j}^{*}=\lambda^{*}\|x\|^{2} \tag{3}
\end{equation*}
$$

But $\mathrm{a}_{\mathrm{ij}}=\mathrm{a}_{\mathrm{ji}}$ and therefore the left hand side of (3) becomes

$$
\sum_{i, j=1}^{n} x_{i} a_{i j} x_{j}^{*}=\sum_{i, j=1}^{n} x_{j}^{*} a_{j i} x_{i}=\sum_{i, j=1}^{n} x_{i}^{*} a_{i j} x_{j}
$$

where in the last step we relabelled the index i by j and the index j by i . Since this shows that the left hand sides of (2) and (3) are equal, it follows that $\lambda=$ $\lambda^{*}$ as claimed.

We will return to this theorem in Chapter 10 where it will be proved in an entirely different manner.

## Exercises

1. Finish the proof of the corollary to Theorem 8.1.
2. Show that if $A, B \in M_{n}(\mathbb{C})$ are unitarily similar and $A$ is normal, then $B$ is also normal.
3. Suppose $A, B \in M_{n}(\mathbb{C})$ commute (i.e., $\left.A B=B A\right)$.
(a) Prove there exists a unitary matrix U such that $\mathrm{U}^{\dagger} \mathrm{AU}$ and $\mathrm{U}^{\dagger} \mathrm{BU}$ are both upper-triangular. [Hint: Let $\mathrm{V}_{\lambda} \subset \mathbb{C}^{\mathrm{n}}$ be the eigenspace of B corresponding to the eigenvalue $\lambda$. Show that $\mathrm{V}_{\lambda}$ is invariant under A , and
hence show that A and B have a common eigenvector $\widetilde{\mathrm{U}}^{1}$. Now proceed as in the proof of Theorem 8.1.]
(b) Show that if A and B are also normal, then there exists a unitary matrix U such that $\mathrm{U}^{\dagger} \mathrm{AU}$ and $\mathrm{U}^{\dagger} \mathrm{BU}$ are diagonal.
4. Can every matrix $A \in M_{n}(\mathbb{C})$ be written as a product of two unitary matrices? Explain.
5. (a) Prove that if H is Hermitian, then $\operatorname{det} \mathrm{H}$ is real.
(b) Is it the case that every square matrix A can be written as the product of finitely many Hermitian matrices? Explain.
6. A matrix M is skew-Hermitian if $\mathrm{M}^{\dagger}=-\mathrm{M}$.
(a) Show that skew-Hermitian matrices are normal.
(b) Show that any square matrix A can be written as a sum of a skewHermitian matrix and a Hermitian matrix.
7. Describe all diagonal unitary matrices. Prove that any $\mathrm{n} x \mathrm{n}$ diagonal matrix can be written as a finite sum of unitary diagonal matrices. [Hint: Do the cases $\mathrm{n}=1$ and $\mathrm{n}=2$ to get the idea.]
8. Using the previous exercise, show that any $\mathrm{n} x \mathrm{n}$ normal matrix can be written as the sum of finitely many unitary matrices.
9. If A is unitary, does this imply that $\operatorname{det} \mathrm{A}^{\mathrm{k}}=1$ for some integer k ? What if A is a real unitary matrix (i.e., orthogonal)?
10. (a) Is an $\mathrm{n} x \mathrm{n}$ matrix A that is similar (but not necessarily unitarily similar) to a Hermitian matrix necessarily Hermitian?
(b) If A is similar to a normal matrix, is A necessarily normal?
11. If N is normal and $\mathrm{N}_{\mathrm{x}}=\lambda \mathrm{x}$, prove that $\mathrm{N}^{\dagger} \mathrm{x}=\lambda^{*} \mathrm{x}$. [Hint: First treat the case where N is diagonal.]
12. Does the fact that A is similar to a diagonal matrix imply that A is normal?
13. Discuss the following conjecture: If $\mathrm{N}_{1}$ and $\mathrm{N}_{2}$ are normal, then $\mathrm{N}_{1}+\mathrm{N}_{2}$ is normal if and only if $\mathrm{N}_{1} \mathrm{~N}_{2}{ }^{\dagger}=\mathrm{N}_{2}{ }^{\dagger} \mathrm{N}_{1}$.
14. (a) If $A \in M_{n}(\mathbb{R})$ is nonzero and skew-symmetric, show that $A$ can not have any real eigenvalues.
(b) What can you say about the eigenvalues of such a matrix?
(c) What can you say about the rank of A?
15. Let $\sigma \in \mathrm{S}_{\mathrm{n}}$ be a permutation, and let $\mathrm{f}:\{1, \ldots, \mathrm{n}\} \rightarrow\{+1,-1\}$. Define the signed permutation matrix $P_{\sigma}^{f}$ by

$$
P_{\sigma}^{f}(i, j)=\left\{\begin{array}{cl}
f(j) & \text { if } \sigma(j)=i \\
0 & \text { otherwise }
\end{array}\right.
$$

Show that signed permutation matrices are orthogonal.
16. (a) Prove that a real $n \times n$ matrix $A$ that commutes with all $n$-square real orthogonal matrices is a multiple of $\mathrm{I}_{\mathrm{n}}$. [Hint: Show that the matrices $\mathrm{E}_{\mathrm{ij}}$ of Section 3.6 can be represented as sums of signed permutation matrices.]
(b) What is true for a complex matrix that commutes with all unitary matrices?

### 8.2 MATRICES OVER THE RING OF POLYNOMIALS

For the remainder of this chapter we will be discussing matrices with polynomial entries. Unfortunately, this requires some care since the ring of polynomials $\mathcal{F}[x]$ does not form a field (see Theorem 6.2, Corollary 3). However, the reader should recall that it is possible to embed $\mathcal{F}[\mathrm{x}]$ (or any integral domain for that matter) in a field of quotients as we saw in Section 6.5 (see Theorem 6.16). This simply means that quotients (i.e., rational functions) such as $f(x) / g(x)$ are defined (if $g \neq 0)$, along with their inverses $g(x) / f(x)($ if $f \neq 0)$.

First of all, we will generally restrict ourselves to only the real and complex number fields. In other words, $\mathcal{F}$ will be taken to mean either $\mathbb{R}$ or $\mathbb{C}$ unless otherwise stated. Next, we introduce some additional simplifying notation. We denote $\mathcal{F}[\mathrm{x}]$ (the ring of polynomials) by $\mathcal{P}$, and the associated field of quotients by $\mathcal{R}$ (think of $\mathcal{P}$ as meaning polynomial and $\mathcal{R}$ as meaning ratio). Thus, an $\mathrm{m} \times \mathrm{n}$ matrix with polynomial entries is an element of $\mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$, and an $m \times n$ matrix over the field of quotients is an element of $M_{m x n}(\mathcal{R})$. Note that $\mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$ is actually a subset of $\mathrm{M}_{\mathrm{mxn}}(\mathcal{R})$ since any polynomial $\mathrm{p}(\mathrm{x})$ may be written as $\mathrm{p}(\mathrm{x}) / 1$.

It is important to realize that since $\mathcal{R}$ is a field, all of our previous results apply to $\mathrm{M}_{\mathrm{m} \times \mathrm{n}}(\mathcal{R})$ just as they do to $\mathrm{M}_{\mathrm{m} \times \mathrm{n}}(\mathcal{F})$. However, we need to reformulate some of our definitions in order to handle $\mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$. In other words, as
long as we allow all operations in $\mathcal{R}$ there is no problem. Where we must be careful is when we restrict ourselves to multiplication by polynomials only (rather than by rational functions). To begin with, we must modify the definition of elementary row and column operations that we gave in Section 3.2. In particular, we now define the $\mathcal{P}$-elementary row (column) operations as follows. The type $\alpha$ operation remains the same, the type $\beta$ operation is multiplication by $\mathrm{c} \in \mathcal{F}$, and the type $\gamma$ operation is now taken to be the addition of a polynomial multiple of one row (column) to another. In other words, if $\mathrm{A}_{\mathrm{i}}$ is the ith row of $\mathrm{A} \in \mathrm{M}_{\mathrm{m} \times \mathrm{n}}(\mathcal{P})$, then the $\mathcal{P}$-elementary operations are:
( $\alpha$ ) Interchange $\mathrm{A}_{\mathrm{i}}$ and $\mathrm{A}_{\mathrm{j}}$.
( $\beta$ ) $\mathrm{A}_{\mathrm{i}} \rightarrow \mathrm{cA}_{\mathrm{i}}$ where $\mathrm{c} \in \mathcal{F}$.
$(\gamma) \mathrm{A}_{\mathrm{i}} \rightarrow \mathrm{A}_{\mathrm{i}}+\mathrm{pA}_{\mathrm{j}}$ where $\mathrm{p} \in \mathcal{P}$.
With these modifications, it is easy to see that all of our discussion on the techniques of reduction to row-echelon form remains valid, although now the distinguished elements of the matrix (i.e., the first nonzero entry in each row) will in general be polynomials (which we will assume to be monic). In other words, the row-echelon form of a matrix $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ will in general be an upper-triangular matrix in $\mathrm{M}_{\mathrm{n}}(\mathcal{P})$ (which may, however, have zeros on the main diagonal). However, if $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ is nonsingular, then $\mathrm{r}(\mathrm{A})=\mathrm{n}$, and the row-echelon form of A will be upper-triangular with nonzero monic polynomials down the main diagonal. (This is true since $\mathrm{M}_{\mathrm{n}}(\mathcal{P}) \subset \mathrm{M}_{\mathrm{n}}(\mathcal{R})$, and hence all of our results dealing with the rank remain valid for elements of $\mathrm{M}_{\mathrm{n}}(\mathcal{P})$ ). In other words, the row-echelon form of $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ will be

$$
\left(\begin{array}{ccccc}
p_{11} & p_{12} & p_{13} & \cdots & p_{1 n} \\
0 & p_{22} & p_{23} & \cdots & p_{2 n} \\
0 & 0 & p_{33} & \cdots & p_{3 n} \\
\vdots & \vdots & \vdots & & \vdots \\
0 & 0 & 0 & \cdots & p_{n n}
\end{array}\right)
$$

where each $\mathrm{p}_{\mathrm{ij}} \in \mathcal{P}$.
Example 8.2 Let us illustrate the basic approach in applying $\mathcal{P}$-elementary operations. For notational simplicity we will consider only the first column of a matrix $\mathrm{A} \in \mathrm{M}_{3}(\mathcal{P})$. Thus, suppose we have

$$
\left(\begin{array}{c}
x^{2}-2 x+1 \\
x-1 \\
x^{2}+2
\end{array}\right)
$$

Multiplying the second row by -x and adding to the third yields

$$
\left(\begin{array}{c}
x^{2}-2 x+1 \\
x-1 \\
x+2
\end{array}\right)
$$

Adding -1 times the second row to the third and then multiplying the third by $1 / 3$ now yields

$$
\left(\begin{array}{c}
x^{2}-2 x+1 \\
x-1 \\
1
\end{array}\right)
$$

Adding $-(x-1)$ times the third row to the second, and $-\left(x^{2}-2 x+1\right)$ times the third to the first gives us

$$
\left(\begin{array}{l}
0 \\
0 \\
1
\end{array}\right) .
$$

Finally, interchanging rows 1 and 3 will put this into row-echelon form. Note that while we came up with a field element in this last form, we could have ended up with some other nonconstant polynomial.

We now repeat this procedure on column 2, but only on rows 2 and 3 since only these rows have zeros in the first column. This results in a matrix that will in general have nonzero elements in row 1 of column 1 , in rows 1 and 2 of column 2, and in all three rows of column 3. It should now be clear that when applied to any $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$, this procedure will result in an uppertriangular matrix. //

A moments thought should convince the reader that it will not be possible in general to transform a matrix in $\mathrm{M}_{\mathrm{n}}(\mathcal{P})$ to reduced row-echelon form if we allow only $\mathcal{P}$-elementary operations. For example, if the row-echelon form of $\mathrm{A} \in \mathrm{M}_{2}(\mathcal{P})$ is

$$
\left(\begin{array}{cc}
x^{2}+1 & 2 x-3 \\
0 & x^{2}
\end{array}\right)
$$

then it is impossible to add any polynomial multiple of the second row to the first to eliminate the $2 \mathrm{x}-3$ term. This is exactly the type of difference that occurs between operations in the ring $\mathcal{P}$ and those in the field $\mathcal{R}$.

It should also be clear that we can define $\mathcal{P}$-elementary matrices in the obvious way, and that each $\mathcal{P}$-elementary matrix is also in $\mathrm{M}_{\mathrm{n}}(\mathcal{P})$. Moreover, each $\mathcal{P}$-elementary matrix has an inverse which is also in $\mathrm{M}_{\mathrm{n}}(\mathcal{P})$, as is its transpose (see Theorem 3.23). In addition, Theorem 4.5 remains valid for matrices over $\mathcal{P}$, as does Theorem 4.4 since replacing row $\mathrm{A}_{i}$ by $\mathrm{A}_{\mathrm{i}}+\mathrm{pA}_{j}$ where $p$ is a polynomial also has no effect on det $A$. This shows that if we reduce a matrix $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ to its row-echelon form $\widetilde{\mathrm{A}}$, then the fact that $\widetilde{\mathrm{A}}$ is upper-triangular means that

$$
\operatorname{det} \widetilde{\mathrm{A}}=\mathrm{k} \operatorname{det} \mathrm{~A}
$$

where k is a unit in $\mathcal{P}$ (recall from Example 6.4 that the units of the ring $\mathcal{P}=$ $\mathcal{F}[\mathrm{x}]$ are just the elements of $\mathcal{F}$, i.e., the nonzero constant polynomials). We will refer to units of $P$ as (nonzero) scalars.

We say that a matrix $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(P)$ is a unit matrix if $\mathrm{A}^{-1}$ exists and is also an element of $\mathrm{M}_{\mathrm{n}}(\mathcal{P})$. (Do not confuse a unit matrix with the identity matrix.) Note this is more restrictive than to say that $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ is merely invertible, because we now also require that $\mathrm{A}^{-1}$ have entries only in $\mathcal{P}$, whereas in general it could have entries in $\mathcal{R}$. From our discussion above, we see that $\mathcal{P}_{\text {- }}$ elementary matrices are also unit matrices. The main properties of unit matrices that we shall need are summarized in the following theorem.

Theorem 8.3 If $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ and $\widetilde{\mathrm{A}} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ is the row-echelon form of A , then
(a) A is a unit matrix if and only if A can be row-reduced to $\widetilde{\mathrm{A}}=\mathrm{I}$.
(b) A is a unit matrix if and only if $\operatorname{det} \mathrm{A}$ is a nonzero scalar.
(c) A is a unit matrix if and only if A is a product of $\mathcal{P}$-elementary matrices.

Proof (a) If A is a unit matrix, then $\mathrm{A}^{-1}$ exists so that $\mathrm{r}(\mathrm{A})=\mathrm{n}$ (Theorem 3.21). This means that the row-echelon form of A is an upper-triangular matrix $\widetilde{\mathrm{A}}=\left(\mathrm{p}_{\mathrm{ij}}\right) \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ with n nonzero diagonal entries. Since $\mathrm{AA}^{-1}=\mathrm{I}$, it follows that $(\operatorname{det} \mathrm{A})\left(\operatorname{det} \mathrm{A}^{-1}\right)=1$ (Theorem 4.8 is also still valid) and hence $\operatorname{det} \mathrm{A} \neq 0$. Furthermore, since both det A and $\operatorname{det} \mathrm{A}^{-1}$ are in $\mathcal{P}$, Theorem 6.2(b) shows us that $\operatorname{deg}(\operatorname{det} A)=\operatorname{deg}\left(\operatorname{det} A^{-1}\right)=0$ and thus $\operatorname{det} A$ is a scalar. Our discussion above showed that

$$
k \operatorname{det} A=\operatorname{det} \tilde{A}=\prod_{i=1}^{n} p_{i i}
$$

where k is a scalar, and therefore each polynomial $\mathrm{p}_{\mathrm{ii}}$ must also be of degree zero (i.e., a scalar). In this case we can apply $\mathcal{P}$-elementary row operations to further reduce $\widetilde{A}$ to the identity matrix I.

Conversely, if A is row-equivalent to $\widetilde{\mathrm{A}}=\mathrm{I}$, then we may write

$$
\mathrm{E}_{1} \cdots \mathrm{E}_{\mathrm{r}} \mathrm{~A}=\mathrm{I}
$$

where each $\mathrm{E}_{\mathrm{i}} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ is an elementary matrix. It follows that $\mathrm{A}^{-1}$ exists and is given by $\mathrm{A}^{-1}=\mathrm{E}_{1} \cdots \mathrm{E}_{\mathrm{r}} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$. Thus A is a unit matrix.
(b) If A is a unit matrix, then the proof of part (a) showed that $\operatorname{det} \mathrm{A}$ is a nonzero scalar. On the other hand, if $\operatorname{det} \mathrm{A}$ is a nonzero scalar, then the proof of part (a) showed that $\widetilde{\mathrm{A}}=\mathrm{E}_{1} \cdots \mathrm{E}_{\mathrm{r}} \mathrm{A}=\mathrm{I}$, and hence $\mathrm{A}^{-1}=\mathrm{E}_{1} \cdots \mathrm{E}_{\mathrm{r}} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ so that A is a unit matrix.
(c) If A is a unit matrix, then the proof of part (a) showed that A may be written as a product of $\mathcal{P}$-elementary matrices. Conversely, if A is the product of $\mathcal{P}$-elementary matrices, then we may write $\mathrm{A}=\mathrm{E}_{\mathrm{r}}^{-1} \cdots \mathrm{E}_{1}^{-1} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$. Therefore $\mathrm{A}^{-1}=\mathrm{E}_{1} \cdots \mathrm{E}_{\mathrm{r}} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ also and hence A is a unit matrix.

Recall from Section 5.4 that two matrices $\mathrm{A}, \mathrm{B} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ are said to be similar if there exists a nonsingular matrix $\mathrm{S} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ such that $\mathrm{A}=\mathrm{S}^{-1} \mathrm{BS}$. In order to generalize this, we say that two matrices $\mathrm{A}, \mathrm{B} \in \mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$ are equivalent over $\mathcal{P}$ if there exist unit matrices $\mathrm{P} \in \mathrm{M}_{\mathrm{m}}(\mathcal{P})$ and $\mathrm{Q} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ such that $\mathrm{A}=\mathrm{PBQ}$. The reader should have no trouble showing that this defines an equivalence relation on the set of all $\mathrm{m} \times \mathrm{n}$ matrices over $\mathcal{P}$.

Note that since P and Q are unit matrices, they may be written as a product of $\mathcal{P}$-elementary matrices (Theorem 8.3). Now recall from our discussion at the end of Section 3.8 that multiplying B from the right by an elementary matrix E has the same effect on the columns of B as multiplying from the left by $\mathrm{E}^{\mathrm{T}}$ does on the rows. We thus conclude that if A and B are equivalent over $\mathcal{P}$, then A is obtainable from B by a sequence of $\mathcal{P}$-elementary row and column operations. Conversely, if A is obtainable from B by a sequence of $\mathcal{P}_{-}$ elementary row and column operations, the fact that each $\mathrm{E}_{\mathrm{i}} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ is a unit matrix means that A and B are $\mathcal{P}$-equivalent.

Theorem 8.4 (a) Two matrices $\mathrm{A}, \mathrm{B} \in \mathrm{M}_{\mathrm{mxn}}(P)$ are equivalent over $P$ if and only if A can be obtained from B by a sequence of $\mathcal{P}$-elementary row and column operations.
(b) $\mathcal{P}$-equivalent matrices have the same rank.

Proof (a) This was proved in the preceding discussion.
(b) Suppose $\mathrm{A}, \mathrm{B} \in \mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$ and $\mathrm{A}=\mathrm{PBQ}$ where $\mathrm{P} \in \mathrm{M}_{\mathrm{m}}(\mathcal{P})$ and $\mathrm{Q} \in$ $\mathrm{M}_{\mathrm{n}}(\mathcal{P})$ are unit matrices and hence nonsingular. Then, applying the corollary to Theorem 3.20, we have

$$
\begin{aligned}
\mathrm{r}(A) & =\mathrm{r}(P B Q) \leq \min \{\mathrm{r}(P), \mathrm{r}(B Q)\}=\min \{m, \mathrm{r}(B Q)\}=\mathrm{r}(B Q) \\
& \leq \min \{\mathrm{r}(B), \mathrm{r}(Q)\}=\mathrm{r}(B)
\end{aligned}
$$

Similarly, we see that $r(B)=r\left(P^{-1} A Q^{-1}\right) \leq r(A)$ and hence $r(A)=r(B)$.
Another point that should be clarified is the following computational technicality that we will need to apply several times in the remainder of this chapter. Referring to Section 6.1, we know that the product of two polynomials $p(x)=\sum_{1=0}^{n} a_{i} x^{i}$ and $q(x)=\sum_{j}^{n}=0 b_{j} x^{j}$ is given by

$$
p(x) q(x)=\sum_{k=0}^{m+n} \sum_{t=0}^{k} a_{t} x^{t} b_{k-t} x^{k-t}
$$

where we have been careful to write everything in its original order. In the special case that $\mathrm{x}, \mathrm{a}_{\mathrm{i}}, \mathrm{b}_{\mathrm{j}} \in \mathcal{F}$, this may be written in the more common and simpler form

$$
p(x) q(x)=\sum_{k=0}^{m+n} c_{k} x^{k}
$$

where $c_{k}=\sum_{t=0}^{k} a_{t} b_{k-t}$. However, we will need to evaluate the product of two polynomials when the coefficients as well as the indeterminate x are matrices. In this case, none of the terms in the general form for pq can be assumed to commute with each other, and we shall have to be very careful in evaluating such products. We do though, have the following useful special case.

Theorem 8.5 Let $p(x)=\sum_{i=0}^{m} a_{i} x^{i}$ and $q(x)=\sum_{j=0}^{n} b_{j} x^{j}$ be polynomials with (matrix) coefficients $\mathrm{a}_{\mathrm{i}}, \mathrm{b}_{\mathrm{j}} \in \mathrm{M}_{\mathrm{s}}(\mathcal{F})$, and let $r(x)=\sum_{k=0}^{m+n} c_{k} x^{k}$ where $\mathrm{c}_{\mathrm{k}}=$ $\sum_{t=0}^{k} a_{t} b_{k-t}$. Then if $\mathrm{A} \in \mathrm{M}_{\mathrm{s}}(\mathcal{F})$ commutes with all of the $\mathrm{b}_{\mathrm{j}} \in \mathrm{M}_{\mathrm{s}}(\mathcal{F})$, we have $\mathrm{p}(\mathrm{A}) \mathrm{q}(\mathrm{A})=\mathrm{r}(\mathrm{A})$.

Proof We simply compute using $\mathrm{Ab}_{\mathrm{j}}=\mathrm{b}_{\mathrm{j}} \mathrm{A}$ :

$$
\begin{aligned}
p(A) q(A) & =\sum_{k=0}^{m+n} \sum_{t=0}^{k} a_{t} A^{t} b_{k-t} A^{k-t}=\sum_{k=0}^{m+n} \sum_{t=0}^{k} a_{t} b_{k-t} A^{k} \\
& =\sum_{k=0}^{m+n} c_{k} A^{k}=r(A) .
\end{aligned}
$$

What this theorem has shown us is that if A commutes with all of the $b_{j}$, then we may use the simpler form for the product of two polynomials. As an interesting application of this result, we now give yet another (very simple) proof of the Cayley-Hamilton theorem. Suppose $A \in M_{n}(\mathcal{F})$, and consider its characteristic matrix xI - A along with the characteristic polynomial $\Delta_{\mathrm{A}}(\mathrm{x})=\operatorname{det}(\mathrm{xI}-\mathrm{A})$. Writing equation (1b) of Section 4.3 in matrix notation we obtain

$$
[\operatorname{adj}(x I-A)](x I-A)=\Delta_{A}(x) I
$$

Now notice that any matrix with polynomial entries may be written as a polynomial with (constant) matrix coefficients (see the proof of Theorem 7.10). Then $\operatorname{adj}(\mathrm{xI}-\mathrm{A})$ is just a polynomial in x of degree $\mathrm{n}-1$ with (constant) matrix coefficients, and $\mathrm{xI}-\mathrm{A}$ is similarly a polynomial in x of degree 1 . Since A obviously commutes with I and A, we can apply Theorem 8.5 with $p(x)=\operatorname{adj}(x I-A)$ and $q(x)=x I-A$ to obtain $p(A) q(A)=\Delta_{A}(A)$. But $q(A)=$ 0 , and hence we find that $\Delta_{\mathrm{A}}(\mathrm{A})=0$.

The last technical point that we wish to address is the possibility of dividing two polynomials with matrix coefficients. The reason that this is a problem is that all of our work in Chapter 6 was based on the assumption that we were dealing with polynomials over a field, and the set of all square matrices of any fixed size certainly does not in general form a field. Referring back to the proof of the division algorithm (Theorem 6.3), we see that the process of dividing $f(x)=a_{m} x^{m}+\cdots+a_{1} x+a_{0}$ by $g(x)=b_{n} x^{n}+\cdots+b_{1} x+$ $b_{0}$ depends on the existence of $b_{n}{ }^{-1}$. This then allows us to show that $x-c$ is a factor of $f(x)$ if and only if $c$ is a root of $f(x)$ (Corollary to Theorem 6.4).

We would like to apply Theorem 6.4 to a special case of polynomials with matrix coefficients. Thus, consider the polynomials $f(x)=B_{n} x+\cdots+B_{1} x+$ $B_{0}$ and $g(x)=x I-A$ where $A, B_{i} \in M_{n}(\mathcal{F})$. In this case, $I$ is obviously invertible and we may divide $g(x)$ into $f(x)$ in the usual manner. The first two terms of the quotient $\mathrm{q}(\mathrm{x})$ are then given by

$$
\begin{aligned}
& \left.x I - A \longdiv { B _ { n } x ^ { n } + ( B _ { n - 1 } + B _ { n } A ) x ^ { n - 2 } } B _ { n - 1 } x ^ { n - 1 } + \cdots + B _ { 1 } x + B _ { 0 }\right) \\
& \frac{B_{n} x^{n}-\frac{B_{n} A x^{n-1}}{\left(B_{n-1}+B_{n} A\right) x^{n-1}}}{}
\end{aligned}
$$

It should now be clear (using Theorem 8.5) that Theorem 6.4 applies in this special case, and if $f(A)=0$, then we may write $f(x)=q(x)(x I-A)$. In other words, if A is a root of $f(x)$, then $x I-A$ is a factor of $f(x)$. Note that in order
to divide $\mathrm{f}(\mathrm{x})=\mathrm{B}_{\mathrm{n}} \mathrm{x}^{\mathrm{n}}+\cdots+\mathrm{B}_{0}$ by $\mathrm{g}(\mathrm{x})=\mathrm{A}_{\mathrm{m}} \mathrm{x}^{\mathrm{m}}+\cdots+\mathrm{A}_{0}$, only the leading coefficient $A_{m}$ of $g(x)$ need be invertible.

Let us also point out that because matrix multiplication is not generally commutative, the order in which we multiply the divisor and quotient is important when dealing with matrix coefficients. We will adhere to the convention used in the above example.

Another point that we should take note of is the following. Two polyno-
 defined to be equal if $A_{k}=B_{k}$ for every $k=1, \ldots, m$. For example, recalling that x is just an indeterminate, we consider the polynomial $\mathrm{p}(\mathrm{x})=\mathrm{A}_{0}+\mathrm{A}_{1} \mathrm{x}=$ $A_{0}+x A_{1}$. If $C \in M_{n}(\mathcal{F})$ does not commute with $A_{1}$ (i.e., $C A_{1} \neq A_{1} C$ ), then $A_{0}+A_{1} C \neq A_{0}+C A_{1}$. This means that going from an equality such as $p(x)=$ $q(x)$ to $p(C)=q(C)$ must be done with care in that the same convention for placing the indeterminate be applied to both $p(x)$ and $q(x)$.

## Exercise

Determine whether or not each of the following matrices is a unit matrix by verifying each of the properties listed in Theorem 8.3:
(a) $\left(\begin{array}{ccc}x+2 & 1 & -3 x^{3}-6 x^{2} \\ 2 x+6 & 2 & -6 x^{3}-18 x^{2} \\ x^{2}+2 x & x^{2}+x+1 & -3 x^{4}-6 x^{3}-3\end{array}\right)$
(b) $\left(\begin{array}{ccc}x+1 & x^{2} & -2 \\ x^{2}-1 & x^{3}-x^{2} & x+7 \\ 3 x^{2}+3 x & 3 & 0\end{array}\right)$
(c) $\left(\begin{array}{cccc}x^{2}+3 x+2 & 0 & x & x^{3}-3 x^{2} \\ 2 x^{2}+4 x & x^{2} & 0 & x-3 \\ x+2 & -x^{2} & 1 & x^{2}-3 x \\ 3 x+6 & -6 x^{2} & 3 & 3 x^{2}-9 x\end{array}\right)$

### 8.3 THE SMITH CANONICAL FORM

If the reader has not studied (or does not remember) the Cauchy-Binet theorem (Section 4.6), now is the time to go back and read it. We will need this result several times in what follows, as well as the notation defined in that section.

We know that the norm of any integer is just its absolute value, and the greatest common divisor of a set of nonzero integers is just the largest positive integer that divides them all. Similarly, we define the norm of any polynomial to be its degree, and the greatest common divisor (frequently denoted by gcd ) of a set of nonzero polynomials is the polynomial of highest degree that divides all of them. By convention, we will assume that the gcd is monic (i.e., the leading coefficient is 1).

Suppose $\mathrm{A} \in \mathrm{M}_{\mathrm{m} \times \mathrm{n}}(\mathcal{P})$, and assume that $1 \leq \mathrm{k} \leq \min \{\mathrm{m}, \mathrm{n}\}$. If A has at least one nonzero k-square subdeterminant, then we define $f_{k}$ to be the greatest common divisor of all $\mathrm{k} t h$ order subdeterminants of A . In other words,

$$
\mathrm{f}_{\mathrm{k}}=\operatorname{gcd}\{\operatorname{det} \mathrm{A}[\alpha \mid \beta]: \alpha \in \mathrm{INC}(\mathrm{k}, \mathrm{~m}), \beta \in \mathrm{INC}(\mathrm{k}, \mathrm{n})\} .
$$

If there is no nonzero $k$ th order subdeterminant, then we define $f_{k}=0$. Furthermore, for notational convenience we define $f_{0}=1$. The numbers $f_{k}$ are called the determinantal divisors of $A$. We will sometimes write $f_{k}(A)$ if there is more than one matrix under consideration.

## Example 8.3 Suppose

$$
A=\left(\begin{array}{ccc}
x & 2 x^{2} & 1 \\
x^{3} & x+2 & x^{2} \\
x+2 & x-1 & 0
\end{array}\right) .
$$

Then the sets of nonzero 1-, 2- and 3-square subdeterminants are, respectively,

$$
\begin{aligned}
& \left\{x, 2 x^{2}, 1, x^{3}, x+2, x^{2}, x+2, x-1\right\} \\
& \begin{aligned}
\left\{-x\left(2 x^{4}-x-2\right),\right. & 2 x^{4}-x-2, x^{4}-x^{3}-x^{2}-4 x-4,-x^{2}(x+2), \\
& \left.-x^{2}(x-1),-x\left(2 x^{2}+3 x+1\right),-(x+2),-(x-1)\right\}
\end{aligned} \\
& \left\{2 x^{5}+4 x^{4}-x^{2}-4 x-4\right\}
\end{aligned}
$$

and hence $\mathrm{f}_{1}=1, \mathrm{f}_{2}=1$ and $\mathrm{f}_{3}=\mathrm{x}^{5}+2 \mathrm{x}^{4}-(1 / 2) \mathrm{x}^{2}-2 \mathrm{x}-2 . / /$

Our next result contains two very simple but important properties of determinantal divisors. Recall that the notation plq means p divides q .

Theorem 8.6 (a) If $f_{k}=0$, then $f_{k+1}=0$.
(b) If $\mathrm{f}_{\mathrm{k}} \neq 0$, then $\mathrm{f}_{\mathrm{k}} \mid \mathrm{f}_{\mathrm{k}+1}$.

Proof Using Theorem 6.6, it is easy to see that these are both immediate consequences of Theorem 4.10 since a $(\mathrm{k}+1)$ th order subdeterminant may be written as a linear combination of $\mathrm{k} t h$ order subdeterminants.

If $\mathrm{A} \in \mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$ has rank r , then Theorem 4.12 tells us that $\mathrm{f}_{\mathrm{r}} \neq 0$ while $f_{r+1}=0$. Hence, according to Theorem 8.6(b), we may define the quotients $q_{k}$ by

$$
\mathrm{f}_{\mathrm{k}}=\mathrm{q}_{\mathrm{k}} \mathrm{f}_{\mathrm{k}-1}
$$

for each $\mathrm{k}=1, \ldots, \mathrm{r}$. The polynomials $\mathrm{q}_{\mathrm{k}}$ are called the invariant factors of A. Note that $f_{0}=1$ implies $f_{1}=q_{1}$, and hence

$$
\mathrm{f}_{\mathrm{k}}=\mathrm{q}_{\mathrm{k}} \mathrm{f}_{\mathrm{k}-1}=\mathrm{q}_{\mathrm{k}} \mathrm{q}_{\mathrm{k}-1} \mathrm{f}_{\mathrm{k}-2}=\cdots=\mathrm{q}_{\mathrm{k}} \mathrm{q}_{\mathrm{k}-1} \cdots \mathrm{q}_{1} .
$$

Because each $f_{k}$ is defined to be monic, it follows that each $q_{k}$ is also monic. Moreover, the unique factorization theorem (Theorem 6.6) shows that each $q_{k}$ ( $\mathrm{k}=1, \ldots, \mathrm{r}$ ) can be factored uniquely (except for order) into products of powers of prime (i.e., irreducible) polynomials as

$$
\mathrm{q}_{\mathrm{k}}=\mathrm{p}_{1}{ }^{\mathrm{e}_{1}} \mathrm{p}_{2}{ }^{\mathrm{e}_{2}} \cdots \mathrm{p}_{\mathrm{s}}{ }^{\mathrm{e}_{\mathrm{s}}}
$$

where $\mathrm{p}_{1}, \ldots, \mathrm{p}_{\mathrm{s}}$ are all the distinct prime factors of the invariant factors, and each $e_{i}$ is a nonnegative integer. Of course, since every $q_{k}$ will not necessarily contain all of the $\mathrm{p}_{\mathrm{i}}$ 's as factors, some of the $\mathrm{e}_{\mathrm{i}}$ 's may be zero.

Each of the factors $p_{i}{ }^{e_{i}}$ for which $e_{i}>0$ is called an elementary divisor of A. We count an elementary divisor once for each time that it appears as a factor of an invariant factor. This is because a given elementary divisor can appear as a factor in more than one invariant factor. Note also that the elementary divisors clearly depend on the field under consideration (see Example 6.7). However, the elementary divisors of a matrix over $\mathbb{C}[x]$ are always powers of linear polynomials (Theorem 6.13). As we shall see following Theorem 8.8 below, the list of elementary divisors determines the list of invariant factors, and hence the determinantal divisors.

Example 8.4 Let $\mathrm{A}_{1}$ be the $3 \times 3$ matrix

$$
A_{1}=\left(\begin{array}{rrc}
x & -1 & 0 \\
0 & x & -1 \\
-1 & 1 & x-1
\end{array}\right)
$$

and note that det $A_{1}=(x-1)\left(x^{2}+1\right)$. Now consider the block diagonal matrix

$$
A=\left(\begin{array}{cc}
A_{1} & 0 \\
0 & A_{1}
\end{array}\right)
$$

Using Theorem 4.14, we immediately find

$$
\mathrm{f}_{6}=\operatorname{det} \mathrm{A}=(\mathrm{x}-1)^{2}\left(\mathrm{x}^{2}+1\right)^{2} .
$$

We now observe that every $5 \times 5$ submatrix of A is either block triangular with a 3 x 3 matrix on its diagonal that contains one zero row (so the determinant is zero), or else is block diagonal with $\mathrm{A}_{1}$ as one of the blocks (you should try to write out some of these and see this for yourself). Therefore

$$
\mathrm{f}_{5}=(\mathrm{x}-1)\left(\mathrm{x}^{2}+1\right)
$$

and hence

$$
\mathrm{q}_{6}=\mathrm{f}_{6} / \mathrm{f}_{5}=(\mathrm{x}-1)\left(\mathrm{x}^{2}+1\right)
$$

As to $f_{4}$, we see that some of the $4 \times 4$ subdeterminants contain det $A_{1}$ while others (such as the one obtained by deleting both rows and columns 3 and 4) do not contain any factors in common with this. Thus $f_{4}=1$ and we must have $q_{5}=f_{5}$. Since $f_{6}=q_{1} q_{2} \cdots q_{6}$, it follows that $q_{4}=q_{3}=q_{2}=q_{1}=1$.

If we regard $A$ as a matrix over $\mathbb{R}[x]$, then the elementary divisors of $A$ are $x-1, x^{2}+1, x-1, x^{2}+1$. However, if we regard $A$ as a matrix over $\mathbb{C}[x]$, then its elementary divisors are $\mathrm{x}-1, \mathrm{x}+i, \mathrm{x}-i, \mathrm{x}-1, \mathrm{x}+i, \mathrm{x}-i$. //

Theorem 8.7 Equivalent matrices have the same determinantal divisors.

Proof Suppose that A = PBQ. Applying the Cauchy-Binet theorem (the corollary to Theorem 4.15), we see that any kth order subdeterminant of A is just a sum of multiples of $\mathrm{k} t h$ order subdeterminants of B . But then the gcd of all k th order subdeterminants of B must divide all the $\mathrm{k} t h$ order subdeterminants of $A$. In other words, $f_{k}(B) \mid f_{k}(A)$. Conversely, writing $B=P^{-1} A Q^{-1}$ we see that $f_{k}(A) \mid f_{k}(B)$, and therefore $f_{k}(A)=f_{k}(B)$.

From Example 8.4 (which was based on a relatively simple block diagonal matrix), it should be obvious that a brute force approach to finding invariant factors leaves much to be desired. The proof of our next theorem is actually nothing more than an algorithm for finding the invariant factors of any matrix A. The matrix B defined in the theorem is called the Smith canonical (or normal) form of A. After the proof, we give an example that should clarify the various steps outlined in the algorithm.

Theorem 8.8 (Smith Canonical Form) Suppose $A \in \mathrm{M}_{\mathrm{m} \times \mathrm{n}}(\mathcal{P})$ has rank r . Then A has precisely $r+1$ nonzero determinantal divisors $f_{0}, f_{1}, \ldots, f_{r}$, and A is equivalent over $\mathcal{P}$ to a unique diagonal matrix $\mathrm{B}=\left(\mathrm{b}_{\mathrm{ij}}\right) \in \mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$ with $b_{i i}=q_{i}=f_{i} / f_{i-1}$ for $i=1, \ldots, r$ and $b_{i j}=0$ otherwise. Moreover $q_{i} \mid q_{i+1}$ for each $\mathrm{i}=1, \ldots, r-1$.

Proof While we have already seen that A has precisely $\mathrm{r}+1$ nonzero determinantal divisors, this will also fall out of the proof below. Furthermore, the uniqueness of B follows from the fact that equivalence classes are disjoint, along with Theorem 8.7 (because determinantal divisors are defined to be monic). As to existence, we assume that $\mathrm{A} \neq 0$ or it is already in Smith form. Note in the following that all we will do is perform a sequence of $\mathcal{P}$ elementary row and column operations on A. Recall that if E is an elementary matrix, then EA represents the same elementary row operation applied to A, and $\mathrm{AE}^{\mathrm{T}}$ is the same operation applied to the columns of A . Therefore, what we will finally arrive at is a matrix of the form $B=P A Q$ where $P=E_{i_{1}} \cdots E_{i_{r}}$ and $\mathrm{Q}=\mathrm{E}_{\mathrm{j}_{1}}{ }^{\mathrm{T}} \cdots \mathrm{E}_{\mathrm{j}_{\mathrm{s}}}{ }^{\mathrm{T}}$. Recall also that the norm of a polynomial is defined to be its degree.

Step 1. Search A for a nonzero entry of least norm and bring it to the $(1,1)$ position by row and column interchanges. By subtracting the appropriate multiples of row 1 from rows $2, \ldots, m$, we obtain a matrix in which every element of column 1 below the $(1,1)$ entry is either 0 or of smaller norm than the $(1,1)$ entry. Now perform the appropriate column operations to make every element of row 1 to the right of the $(1,1)$ entry either 0 or of smaller norm than the $(1,1)$ entry. Denote this new matrix by $\widetilde{\mathrm{A}}$.

Step 2. Search the first row and column of $\widetilde{A}$ for a nonzero entry of least norm and bring it to the $(1,1)$ position. Now repeat the procedure of Step 1 to decrease the norm of every element of the first row and column outside the $(1,1)$ position by at least 1 . Repeating this step a finite number of times, we must eventually arrive at a matrix $\mathrm{A}_{1}$ equivalent to A which is 0 everywhere in the first row and column outside the $(1,1)$ position. Let us denote the $(1,1)$ entry of $\mathrm{A}_{1}$ by a.

Step 3. Suppose $b$ is the $(i, j)$ element of $A_{1}(w h e r e ~ i, j>1)$ and $a \not b$. If no such b exists, then go on to Step 4 . Put b in the $(1, \mathrm{j})$ position by adding row i
to row 1 . Since $\mathrm{a} \mid \mathrm{b}$, we may write $\mathrm{b}=\mathrm{aq}+\mathrm{r}$ where $\mathrm{r} \neq 0$ and $\operatorname{deg} \mathrm{r}<\operatorname{deg} \mathrm{a}$ (Theorem 6.3). We place r in the $(1, \mathrm{j})$ position by subtracting q times column 1 from column j . This results in a matrix with an entry of smaller norm than that of a. Now repeat Steps 1 and 2 with this matrix to obtain a new matrix $A_{2}$ equivalent to A which is 0 everywhere in the first row and column outside the $(1,1)$ position.

This process is repeated with $\mathrm{A}_{2}$ to obtain $\mathrm{A}_{3}$ and so forth. We thus obtain a sequence $A_{1}, A_{2}, \ldots, A_{s}$ of matrices in which the norms of the $(1,1)$ entries are strictly decreasing, and in which all elements of row 1 and column 1 are 0 outside the $(1,1)$ position. Furthermore, we go on from $A_{p}$ to obtain $A_{p+1}$ only as long as there is an element of $A_{p}(1 \mid 1)$ that is not divisible by the $(1,1)$ element of $A_{p}$. Since the norms of the $(1,1)$ entries are strictly decreasing, this process must terminate with a matrix $\mathrm{C}=\left(\mathrm{c}_{\mathrm{ij}}\right) \in \mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$ equivalent to A and having the following properties:
(i) $\mathrm{c}_{11} \mid \mathrm{c}_{\mathrm{ij}}$ for every $\mathrm{i}, \mathrm{j}>1$;
(ii) $\mathrm{c}_{1 \mathrm{j}}=0$ for every $\mathrm{j}=2, \ldots, \mathrm{n}$;
(iii) $\mathrm{c}_{\mathrm{i} 1}=0$ for every $\mathrm{i}=2, \ldots, \mathrm{~m}$.

Step 4. Now repeat the entire procedure on the matrix C, except that this time apply the $\mathcal{P}$-elementary row and column operations to rows $2, \ldots, \mathrm{~m}$ and columns $2, \ldots, \mathrm{n}$. This will result in a matrix $\mathrm{D}=\left(\mathrm{d}_{\mathrm{ij}}\right)$ that has all 0 entries in the first two rows and columns except for the $(1,1)$ and $(2,2)$ entries. Since $\mathrm{c}_{11} \mid \mathrm{c}_{\mathrm{ij}}$ (for $\mathrm{i}, \mathrm{j}>1$ ), it follows that $\mathrm{c}_{11} \mid \mathrm{d}_{\mathrm{ij}}$ for all $\mathrm{i}, \mathrm{j}$. (This true because every element of D is just a linear combination of elements of C.) Thus the form of D is

$$
D=\left(\begin{array}{cccc}
c_{11} & 0 & \cdots & 0 \\
0 & d & \cdots & 0 \\
\vdots & \vdots & G & \\
0 & 0 & &
\end{array}\right)
$$

where $\mathrm{G}=\left(\mathrm{g}_{\mathrm{ij}}\right) \in \mathrm{M}_{(\mathrm{m}-2) \mathrm{x}(\mathrm{n}-2)}(\mathcal{P}), \mathrm{c}_{11} \mid \mathrm{d}$ and $\mathrm{c}_{11} \mid \mathrm{g}_{\mathrm{ij}}$ for $\mathrm{i}=1, \ldots, \mathrm{~m}-2$ and $\mathrm{j}=$ $1, \ldots, n-2$. It is clear that we can continue this process until we eventually obtain a diagonal matrix $\mathrm{H}=\left(\mathrm{h}_{\mathrm{ij}}\right) \in \mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$ with the property that $\mathrm{h}_{\mathrm{ij}} \mid \mathrm{h}_{\mathrm{i}+1} \mathrm{i+1}$ and $h_{i+1} \mathrm{i}+1 \neq 0$ for $\mathrm{i}=1, \ldots, \mathrm{p}-1$ (where $\mathrm{p}=$ rank H ). But H is equivalent to A so that H and A have the same determinantal divisors (Theorem 8.7) and p $=r(H)=r(A)=r($ Theorem 8.4(b)).

For each $k$ with $1 \leq k \leq r$, we observe that the only nonzero $k$-square subdeterminants of H are of the form $\prod_{s=1}^{k} h_{i_{s} i_{s}}$, and the gcd of all such products is

$$
f_{k}=\prod_{i=1}^{k} h_{i i}
$$

(since $h_{i \mathrm{ij}} \mid \mathrm{h}_{\mathrm{i}+1} \mathrm{i}+1$ for $\mathrm{i}=1, \ldots, \mathrm{r}-1$ ). But then applying the definition of invariant factors, we see that

$$
\prod_{i=1}^{k} h_{i i}=f_{k}=\prod_{i=1}^{k} q_{i} .
$$

In particular, this shows us that

$$
\begin{aligned}
h_{11} & =q_{1} \\
h_{11} h_{22} & =q_{1} q_{2} \\
& \vdots \\
h_{11} \cdots h_{r r} & =q_{1} \cdots q_{r}
\end{aligned}
$$

and hence $h_{22}=q_{2}, \ldots, h_{r r}=q_{r}$ also. In other words, $H$ is precisely the desired matrix B. Finally, note that $\mathrm{h}_{\mathrm{ij}} \mid \mathrm{h}_{\mathrm{i}+1} \mathrm{i}+1$ is just the statement that $q_{i} \mid q_{i+1}$.

Suppose $\mathrm{A} \in \mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$ has rank r , and suppose that we are given a list of all the elementary divisors of A. From Theorem 8.8, we know that $q_{i} \mid q_{i+1}$ for $\mathrm{i}=1, \ldots, \mathrm{r}-1$. Therefore, to compute the invariant factors of A , we first multiply together the highest powers of all the distinct primes that appear in the list of elementary divisors. This gives us $\mathrm{q}_{\mathrm{r}}$. Next, we multiply together the highest powers of the remaining distinct primes to obtain $\mathrm{q}_{\mathrm{r}-1}$. Continuing this process until the list of elementary divisors is exhausted, suppose that $\mathrm{q}_{\mathrm{k}}$ is the last invariant factor so obtained. If $\mathrm{k}>1$, we then set $\mathrm{q}_{1}=\cdots=\mathrm{q}_{\mathrm{k}-1}=1$. The reader should try this on the list of elementary divisors given at the end of Example 8.4.

Corollary If $\mathrm{A}, \mathrm{B} \in \mathrm{M}_{\mathrm{mxn}}(\mathcal{P})$, then A is $\mathcal{P}$-equivalent to B if and only if A and B have the same invariant factors (or determinantal divisors or elementary divisors).

Proof Let $\mathrm{A}_{\mathrm{S}}$ and $\mathrm{B}_{\mathrm{S}}$ be the Smith forms for A and B . If A and B have the same invariant factors then they have the same Smith form. If we denote $\mathcal{P}_{-}$ equivalence by $\approx$, then $\mathrm{A} \approx \mathrm{A}_{\mathrm{S}}=\mathrm{B}_{\mathrm{S}} \approx \mathrm{B}$ so that $\mathrm{A} \approx \mathrm{B}$. Conversely, if $\mathrm{A} \approx \mathrm{B}$ then $\mathrm{A} \approx \mathrm{B} \approx \mathrm{B}_{\mathrm{S}}$ implies that $\mathrm{A} \approx \mathrm{B}_{\mathrm{S}}$, and hence the uniqueness of $\mathrm{A}_{\mathrm{S}}$ implies that $\mathrm{A}_{\mathrm{S}}=\mathrm{B}_{\mathrm{S}}$, and thus A and B have the same invariant factors.

If we recall Theorem 6.6, then the statement for elementary divisors follows immediately. Now note that $f_{0}=1$ so that $f_{1}=q_{1}$, and in general we then have $f_{k}=q_{k} f_{k-1}$. This takes care of the statement for determinantal divisors.

Example 8.5 Consider the matrix A given in Example 7.3. We shall compute the invariant factors of the associated characteristic matrix xI - A. The reason for using the characteristic matrix will become clear in a later section. According to Step 1, we obtain the following sequence of equivalent matrices. Start with

$$
x I-A=\left(\begin{array}{cccc}
x-2 & -1 & 0 & 0 \\
0 & x-2 & 0 & 0 \\
0 & 0 & x-2 & 0 \\
0 & 0 & 0 & x-5
\end{array}\right) .
$$

Put -1 in the $(1,1)$ position:

$$
\left(\begin{array}{cccc}
-1 & x-2 & 0 & 0 \\
x-2 & 0 & 0 & 0 \\
0 & 0 & x-2 & 0 \\
0 & 0 & 0 & x-5
\end{array}\right)
$$

Add $\mathrm{x}-2$ times row 1 to row 2 , and $\mathrm{x}-2$ times column 1 to column 2 :

$$
\left(\begin{array}{cccc}
-1 & 0 & 0 & 0 \\
0 & (x-2)^{2} & 0 & 0 \\
0 & 0 & x-2 & 0 \\
0 & 0 & 0 & x-5
\end{array}\right)
$$

Since all entries in row 1 and column 1 are 0 except for the $(1,1)$ entry, this last matrix is $\mathrm{A}_{1}$ and we have also finished Step 2. Furthermore, there is no element $b \in A_{1}$ that is not divisible by -1 , so we go on to Step 4 applied to the $3 \times 3$ matrix in the lower right hand corner. In this case, we first apply Step 1 and then follow Step 3. We thus obtain the following sequence of matrices. Put $\mathrm{x}-2$ in the $(2,2)$ position:

$$
\left(\begin{array}{cccc}
-1 & 0 & 0 & 0 \\
0 & x-2 & 0 & 0 \\
0 & 0 & (x-2)^{2} & 0 \\
0 & 0 & 0 & x-5
\end{array}\right)
$$

$(x-2) \xi(x-5)$ so add row 4 to row 2 :

$$
\left(\begin{array}{cccc}
-1 & 0 & 0 & 0 \\
0 & x-2 & 0 & x-5 \\
0 & 0 & (x-2)^{2} & 0 \\
0 & 0 & 0 & x-5
\end{array}\right)
$$

Note $x-5=1(x-2)+(-3)$, so subtract 1 times column 2 from column 4:

$$
\left(\begin{array}{cccc}
-1 & 0 & 0 & 0 \\
0 & x-2 & 0 & -3 \\
0 & 0 & (x-2)^{2} & 0 \\
0 & 0 & 0 & x-5
\end{array}\right)
$$

Now put -3 in the $(2,2)$ position:

$$
\left(\begin{array}{cccc}
-1 & 0 & 0 & 0 \\
0 & -3 & 0 & x-2 \\
0 & 0 & (x-2)^{2} & 0 \\
0 & x-5 & 0 & 0
\end{array}\right)
$$

Add $(x-5) / 3$ times row 2 to row 4 , and then add $(x-2) / 3$ times column 2 to column 4 to obtain

$$
\left(\begin{array}{rccc}
-1 & 0 & 0 & 0 \\
0 & -3 & 0 & 0 \\
0 & 0 & (x-2)^{2} & 0 \\
0 & 0 & 0 & (x-2)(x-5) / 3
\end{array}\right)
$$

Elementary long division (see Example 6.2) shows that $(x-2)(x-5) / 3$ divided by $(x-2)^{2}$ equals $1 / 3$ with a remainder of $-x+2$. Following Step 3, we add row 4 to row 3 and then subtract $1 / 3$ times column 3 from column 4 to obtain

$$
\left(\begin{array}{cccc}
-1 & 0 & 0 & 0 \\
0 & -3 & 0 & 0 \\
0 & 0 & (x-2)^{2} & -x+2 \\
0 & 0 & 0 & (x-2)(x-5) / 3
\end{array}\right)
$$

Going back to Step 1 , we first put $-x+2=-(x-2)$ in the $(3,3)$ position. We then add $(x-5) / 3$ times row 3 to row 4 and $(x-2)$ times column 3 to column 4 resulting in

$$
\left(\begin{array}{rrcc}
-1 & 0 & 0 & 0 \\
0 & -3 & 0 & 0 \\
0 & 0 & -(x-2) & 0 \\
0 & 0 & 0 & (x-2)^{2}(x-5) / 3
\end{array}\right)
$$

Lastly, multiplying each row by a suitable nonzero scalar we obtain the final (unique) Smith form

$$
\left(\begin{array}{cccc}
1 & 0 & 0 & 0 \\
0 & 1 & 0 & 0 \\
0 & 0 & x-2 & 0 \\
0 & 0 & 0 & (x-2)^{2}(x-5)
\end{array}\right)
$$

## Exercises

1. Find the invariant factors of the matrix A given in Example 8.4 by using the list of elementary divisors also given in that example.
2. For each of the following matrices A , find the invariant factors of the characteristic matrix xI-A:
(a) $\left(\begin{array}{rrr}-3 & 3 & -2 \\ -7 & 6 & -3 \\ 1 & -1 & 2\end{array}\right)$
(b) $\left(\begin{array}{rrr}0 & 1 & -1 \\ -4 & 4 & -2 \\ -2 & 1 & 1\end{array}\right)$
(c) $\left(\begin{array}{rrrr}2 & -4 & 2 & 2 \\ -2 & 0 & 1 & 3 \\ -2 & -2 & 3 & 3 \\ -2 & -6 & 3 & 7\end{array}\right)$
(d) $\left(\begin{array}{rrrr}0 & -3 & 1 & 2 \\ -2 & 1 & -1 & 2 \\ -2 & 1 & -1 & 2 \\ -2 & -3 & 1 & 4\end{array}\right)$

### 8.4 SIMILARITY INVARIANTS

Recall that $\mathrm{A}, \mathrm{B} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ are similar over $\mathcal{F}$ if there exists a nonsingular matrix $\mathrm{S} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ such that $\mathrm{A}=\mathrm{S}^{-1} \mathrm{BS}$. Note that similar matrices are therefore also equivalent, although the converse is certainly not true (since in general $\mathrm{P} \neq \mathrm{Q}^{-1}$ in our definition of equivalent matrices). For our present purposes however, the following theorem is quite useful.

Theorem 8.9 Two matrices $\mathrm{A}, \mathrm{B} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ are similar over $\mathcal{F}$ if and only if their characteristic matrices $\mathrm{xI}-\mathrm{A}$ and $\mathrm{xI}-\mathrm{B}$ are equivalent over $\mathcal{P}=\mathcal{F}[\mathrm{x}]$. In particular, if $x I-A=P(x I-B) Q$ where $Q^{-1}=R_{m} x^{m}+\cdots+R_{1} x+R_{0}$, then $A=S^{-1} B S$ where $S^{-1}=R_{m} B^{m}+\cdots+R_{1} B+R_{0}$.

Proof If A and B are similar, then there exists a nonsingular matrix $\mathrm{S} \in$ $\mathrm{M}_{\mathrm{n}}(\mathcal{F})$ for which $\mathrm{A}=\mathrm{S}^{-1} \mathrm{BS}$, and hence

$$
x I-A=x I-S^{-1} B S=S^{-1}(x I-B) S
$$

But S is a unit matrix in $\mathrm{M}_{\mathrm{n}}(\mathcal{P})$, and therefore $\mathrm{xI}-\mathrm{A}$ and $\mathrm{xI}-\mathrm{B}$ are $\mathcal{P}$ equivalent.

On the other hand, if $\mathrm{xI}-\mathrm{A}$ and $\mathrm{xI}-\mathrm{B}$ are $\mathcal{P}$-equivalent, then there exist unit matrices $\mathrm{P}, \mathrm{Q} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ such that

$$
x I-A=P(x I-B) Q
$$

We wish to find a matrix $\mathrm{S} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ for which $\mathrm{A}=\mathrm{S}^{-1} \mathrm{BS}$. Since $\mathrm{Q} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ is a unit matrix, we may apply Theorem 4.11 to find its inverse $\mathrm{R} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ which is also a unit matrix and hence will also have polynomial entries. In fact, we may write (as in the proof of Theorem 7.10)

$$
\begin{equation*}
R=R_{m} x^{m}+R_{m-1} x^{m-1}+\cdots+R_{1} x+R_{0} \tag{1}
\end{equation*}
$$

where $m$ is the highest degree of any polynomial entry of $R$ and each $R_{i} \in$ $\mathrm{M}_{\mathrm{n}}(\mathcal{F})$.

From $x I-A=P(x I-B) Q$ and the fact that $P$ and $Q$ are unit matrices we have

$$
\begin{equation*}
P^{-1}(x I-A)=(x I-B) Q=Q x-B Q . \tag{2}
\end{equation*}
$$

Now recall Theorem 8.5 and the discussion following its proof. If we write both $\mathrm{P}^{-1}$ and $\mathrm{Q} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ in the same form as we did in (1) for R , then we may replace x by A in the resulting polynomial expression for Q to obtain a matrix
$\mathrm{W} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$. Since A commutes with I and A , and $\mathrm{B} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$, we may apply Theorem 8.5 and replace x by A on both sides of (2), resulting in

$$
0=\mathrm{WA}-\mathrm{BW} .
$$

Since $R$ is the inverse of $Q$ and $Q x^{i}=x^{i} Q$, we have $R Q=I$ or (from (1))

$$
\mathrm{R}_{\mathrm{m}} \mathrm{Qx} \mathrm{x}^{\mathrm{m}}+\mathrm{R}_{\mathrm{m}-1} \mathrm{Qx}{ }^{\mathrm{m}-1}+\cdots+\mathrm{R}_{1} \mathrm{Qx}+\mathrm{R}_{0} \mathrm{Q}=\mathrm{I}
$$

Replacing x by A in this expression yields

$$
\begin{equation*}
\sum_{i=0}^{m} R_{i} W A^{i}=I \tag{3}
\end{equation*}
$$

But WA $=\mathrm{BW}$ so that $\mathrm{WA}^{2}=\mathrm{BWA}=\mathrm{B}^{2} \mathrm{~W}$ and, by induction, it follows that $W^{i}=B^{i} W$. Using this in (3) we have

$$
\left(\sum_{i=0}^{m} R_{i} B^{i}\right) W=I
$$

so defining

$$
\begin{equation*}
S^{-1}=\sum_{i=0}^{m} R_{i} B^{i} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F}) \tag{4}
\end{equation*}
$$

we see that $S^{-1}=W^{-1}$ and hence $W=S$. Finally, noting that $\mathrm{WA}=\mathrm{BW}$ implies $\mathrm{A}=\mathrm{W}^{-1} \mathrm{BW}$, we arrive at $\mathrm{A}=\mathrm{S}^{-1} \mathrm{BS}$ as desired.

Corollary 1 Two matrices $A, B \in M_{n}(\mathcal{F})$ are similar if and only if their characteristic matrices have the same invariant factors (or elementary divisors).

Proof This follows directly from Theorem 8.9 and the corollary to Theorem 8.8.

Corollary 2 If $A$ and $B$ are in $M_{n}(\mathbb{R})$, then $A$ and $B$ are similar over $\mathbb{C}$ if and only if they are similar over $\mathbb{R}$.

Proof Clearly, if A and B are similar over $\mathbb{R}$ then they are also similar over $\mathbb{C}$. On the other hand, suppose that A and B are similar over $\mathbb{C}$. We claim that the algorithm in the proof of Theorem 8.9 yields a real S if A and B are real. From the definition of $S$ in the proof of Theorem 8.9 (equation (4)), we see that $S$ will be real if all of the $R_{i}$ are real (since each $B^{i}$ is real by hypothesis), and this in turn requires that Q be real (since $\mathrm{R}=\mathrm{Q}^{-1}$ ). That P and Q can indeed be chosen to be real is left as an exercise for the reader (see Exercise 8.4.1).

The invariant factors of the characteristic matrix of A are called the similarity invariants of A . We will soon show that the similarity invariant of highest degree is just the minimal polynomial for A .

Example 8.6 Let us show that the matrices

$$
A=\left(\begin{array}{ll}
1 & 1 \\
1 & 1
\end{array}\right) \quad \text { and } \quad B=\left(\begin{array}{ll}
0 & 0 \\
0 & 2
\end{array}\right)
$$

are similar over $\mathbb{R}$. We have the characteristic matrices

$$
x I-A=\left(\begin{array}{cc}
x-1 & -1 \\
-1 & x-1
\end{array}\right) \quad \text { and } \quad x I-B=\left(\begin{array}{cc}
x & 0 \\
0 & x-2
\end{array}\right)
$$

and hence the determinantal divisors are easily seen to be $f_{1}(A)=1, f_{2}(A)=$ $x(x-2), f_{1}(B)=1, f_{2}(B)=x(x-2)$. Thus $f_{1}(A)=f_{1}(B)$ and $f_{2}(A)=f_{2}(B)$ so that A and B must be similar by the corollary to Theorem 8.9.

For the sake of illustration, we will show how to compute the matrix $\mathrm{S}^{-1}$ following the method used in the proof of Theorem 8.9 (see equation (4)). While there is no general method for finding the matrices P and Q , the reader can easily verify that if we choose

$$
P=\left(\begin{array}{rr}
1 & x^{2}-x+1 \\
-1 & -x^{2}+x+1
\end{array}\right) \quad Q=\frac{1}{2}\left(\begin{array}{cc}
-x^{2}+3 x-1 & -x^{2}+3 x-1 \\
1 & 1
\end{array}\right)
$$

then $x I-A=P(x I-B) Q$. It is then easy to show that

$$
\begin{aligned}
R=Q^{-1} & =\left(\begin{array}{rr}
1 & x^{2}-3 x+3 \\
-1 & -x^{2}+3 x-1
\end{array}\right) \\
& =\left(\begin{array}{rr}
0 & 1 \\
0 & -1
\end{array}\right) x^{2}+\left(\begin{array}{ll}
0 & -3 \\
0 & -3
\end{array}\right) x+\left(\begin{array}{rr}
1 & 3 \\
-1 & -1
\end{array}\right)
\end{aligned}
$$

and hence (from (4)) we have

$$
S^{-1}=\left(\begin{array}{rr}
0 & 1 \\
0 & -1
\end{array}\right)\left(\begin{array}{ll}
0 & 0 \\
0 & 2
\end{array}\right)^{2}+\left(\begin{array}{rr}
0 & -3 \\
0 & 3
\end{array}\right)\left(\begin{array}{ll}
0 & 0 \\
0 & 2
\end{array}\right)+\left(\begin{array}{rr}
1 & 3 \\
-1 & -1
\end{array}\right)=\left(\begin{array}{rr}
1 & 1 \\
-1 & 1
\end{array}\right)
$$

Now recall the definition of minimal polynomial given in Section 7.3 (see the discussion following the proof of Theorem 7.10). We also recall that the minimal polynomial $\mathrm{m}(\mathrm{x})$ for $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ divides the characteristic polynomial $\Delta_{\mathrm{A}}(\mathrm{x})$. In the particular case that $\mathrm{m}(\mathrm{x})=\Delta_{\mathrm{A}}(\mathrm{x})$, the matrix A is called nonderogatory, and if $m(x) \neq \Delta_{A}(x)$, then (as you might have guessed) $A$ is said to be derogatory. Our next theorem is of fundamental importance.

Theorem 8.10 The minimal polynomial $\mathrm{m}(\mathrm{x})$ for $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ is equal to its similarity invariant of highest degree.

Proof Since $\Delta_{\mathrm{A}}(\mathrm{x})=\operatorname{det}(\mathrm{xI}-\mathrm{A})$ is just a (monic) polynomial of degree n in x , it is clearly nonzero, and hence $\mathrm{q}_{\mathrm{n}}(\mathrm{x})$, the similarity invariant of highest degree, is also nonzero. Now define the matrix $\mathrm{Q}(\mathrm{x})=\operatorname{adj}(\mathrm{xI}-\mathrm{A})$, and note that the entries of $\mathrm{Q}(\mathrm{x})$ are precisely all the $(\mathrm{n}-1)$-square subdeterminants of $x I-A$. This means $f_{n-1}(x)$ (i.e., the $(n-1)$ th determinantal divisor of $\left.x I-A\right)$ is just the monic gcd of all the entries of $\mathrm{Q}(\mathrm{x})$, and therefore we may write

$$
\mathrm{Q}(\mathrm{x})=\mathrm{f}_{\mathrm{n}-1}(\mathrm{x}) \mathrm{D}(\mathrm{x})
$$

where the matrix $D(x)$ has entries that are relatively prime. Noting that by definition we have $\Delta_{\mathrm{A}}(\mathrm{x})=\mathrm{f}_{\mathrm{n}}(\mathrm{x})=\mathrm{q}_{\mathrm{n}}(\mathrm{x}) \mathrm{f}_{\mathrm{n}-1}(\mathrm{x})$, it follows that

$$
\begin{equation*}
f_{n-1}(x) D(x)(x I-A)=Q(x)(x I-A)=\Delta_{A}(x) I=q_{n}(x) f_{n-1}(x) I \tag{1}
\end{equation*}
$$

where we used equation (1b) of Section 4.3. Since $f_{n-1}(x) \neq 0$ (by Theorem 8.6(a) and the fact that $\left.\mathrm{f}_{\mathrm{n}}(\mathrm{x}) \neq 0\right)$, we must have

$$
\begin{equation*}
D(x)(x I-A)=q_{n}(x) I \tag{2}
\end{equation*}
$$

(this follows by equating the polynomial entries of the matrices on each side of (1) and then using Corollary 2(b) of Theorem 6.2).

By writing both sides of (2) as polynomials with matrix coefficients and then applying Theorem 8.5, it follows that $\mathrm{q}_{\mathrm{n}}(\mathrm{A})=0$ and hence $\mathrm{m}(\mathrm{x}) \mid \mathrm{q}_{\mathrm{n}}(\mathrm{x})$ (Theorem 7.4). We may now define the polynomial $p(x)$ by writing

$$
\begin{equation*}
\left.\left.q_{n}(x)=m\right) x\right) p(x) . \tag{3}
\end{equation*}
$$

By definition, A is a root of $\mathrm{m}(\mathrm{x})$, and therefore our discussion at the end of Section 8.2 tells us that we may apply Theorem 6.4 to write

$$
m(x) I=C(x)(x I-A)
$$

where $C(x)$ is a polynomial with matrix coefficients. Using this in (2) we have

$$
\begin{equation*}
D(x)(x I-A)=q_{n}(x) I=p(x) m(x) I=p(x) C(x)(x I-A) \tag{4}
\end{equation*}
$$

where we used the fact that $\mathrm{m}(\mathrm{x})$ and $\mathrm{p}(\mathrm{x})$ are just polynomials with scalar coefficients so that $m(x) p(x)=p(x) m(x)$.

Since $\operatorname{det}(\mathrm{xI}-\mathrm{A}) \neq 0$, we know that $(\mathrm{xI}-\mathrm{A})^{-1}$ exists over $\mathrm{M}_{\mathrm{n}}(\mathcal{R})$, and thus (4) implies that

$$
\mathrm{D}(\mathrm{x})=\mathrm{p}(\mathrm{x}) \mathrm{C}(\mathrm{x}) .
$$

Now regarding both $\mathrm{D}(\mathrm{x})$ and $\mathrm{C}(\mathrm{x})$ as matrices with polynomial entries, this equation shows that $p(x)$ divides each of the entries of $D(x)$. But the entries of $D(x)$ are relatively prime, and hence $p(x)$ must be a unit (i.e., a nonzero scalar). Since both $m(x)$ and $q_{n}(x)$ are monic by convention, (3) implies that $\mathrm{p}(\mathrm{x})=1$, and therefore $\mathrm{q}_{\mathrm{n}}(\mathrm{x})=\mathrm{m}(\mathrm{x})$.

Corollary A matrix $A \in M_{n}(\mathcal{F})$ is nonderogatory if and only if its first $n-1$ similarity invariants are equal to 1 .

Proof Let A have characteristic polynomial $\Delta_{\mathrm{A}}(\mathrm{x})$ and minimal polynomial $\mathrm{m}(\mathrm{x})$. Using the definition of invariant factors and Theorem 8.10 we have

$$
\begin{aligned}
\Delta_{A}(x) & =\operatorname{det}(x I-A)=f_{n}(x)=q_{n}(x) q_{n-1}(x) \cdots q_{1}(x) \\
& =m(x) q_{n-1}(x) \cdots q_{1}(x) .
\end{aligned}
$$

Clearly, if $q_{n-1}(x)=\cdots=q_{1}(x)=1$ then $\Delta_{A}(x)=m(x)$. On the other hand, if $\Delta_{\mathrm{A}}(\mathrm{x})=\mathrm{m}(\mathrm{x})$, then $\mathrm{q}_{\mathrm{n}-1}(\mathrm{x}) \cdots \mathrm{q}_{1}(\mathrm{x})=1$ (Theorem 6.2, Corollary 2(b)) and hence each $\mathrm{q}_{\mathrm{i}}(\mathrm{x})(\mathrm{i}=1, \ldots, \mathrm{n}-1$ ) is a nonzero scalar (Theorem 6.2, Corollary 3). Since each $\mathrm{q}_{\mathrm{k}}(\mathrm{x})$ is defined to be monic, it follows that $\mathrm{q}_{\mathrm{n}-1}(\mathrm{x})=$ $\cdots=q_{1}(x)=1$.

Example 8.7 Comparison of Examples 7.3 and 8.8 shows that the minimal polynomial of the matrix A is indeed the same as its similarity invariant of highest degree. //

## Exercises

1. Finish the proof of Corollary 2 to Theorem 8.9.
2. Show that the minimal polynomial for $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ is the least common multiple of the elementary divisors of $\mathrm{xI}-\mathrm{A}$.
3. If $\left(x^{2}-4\right)^{4}$ is the minimal polynomial of an $n$-square matrix $A$, can $A^{6}-$ $A^{4}+A^{2}-I_{n}$ ever be zero? If $\left(x^{2}-4\right)^{3}$ is the minimal polynomial, can $A^{8}-A^{4}+A^{2}-I_{n}=0$ ? Explain.
4. Is the matrix

$$
\left(\begin{array}{llll}
0 & 0 & 1 & 0 \\
0 & 0 & 0 & 1 \\
1 & 0 & 0 & 0 \\
0 & 1 & 0 & 0
\end{array}\right)
$$

derogatory or nonderogatory? Explain.
5. Suppose $A$ is an $n$-square matrix and $p$ is a polynomial with complex coefficients. If $p(A)=0$, show that $p\left(S A S^{-1}\right)=0$ for any nonsingular $n$-square S . Is this true if p is a polynomial with n -square matrices as coefficients?
6. Prove or disprove:
(a) The elementary divisors of A are all linear if and only if the characteristic polynomial of A is a product of distinct linear factors.
(b) The elementary divisors of A are all linear if and only if the minimal polynomial of A is a product of distinct linear factors.
7. Prove or disprove:
(a) There exists a real nonsingular matrix S such that $\mathrm{SAS}^{-1}=\mathrm{B}$ where

$$
A=\left(\begin{array}{rr}
3 & 0 \\
-1 & 2
\end{array}\right) \quad \text { and } \quad B=\left(\begin{array}{rr}
4 & 2 \\
-1 & 1
\end{array}\right) .
$$

(b) There exists a complex nonsingular matrix S such that $\mathrm{SAS}^{-1}=\mathrm{B}$ where

$$
A=\left(\begin{array}{rr}
3 & 0 \\
-1 & 2
\end{array}\right) \quad \text { and } \quad B=\left(\begin{array}{cc}
4 & 2 i \\
i & 1
\end{array}\right)
$$

### 8.5 THE RATIONAL CANONICAL FORM

Given any monic polynomial $\mathrm{p}(\mathrm{x})=\mathrm{x}^{\mathrm{n}}-\mathrm{a}_{\mathrm{n}-1} \mathrm{x}^{\mathrm{n}-1}-\cdots-\mathrm{a}_{0} \in \mathcal{F}[\mathrm{x}]$, the matrix $\mathrm{C}(\mathrm{p}(\mathrm{x})) \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ defined by

$$
C(p(x))=\left(\begin{array}{cccccc}
0 & 0 & 0 & \cdots & 0 & a_{0} \\
1 & 0 & 0 & \cdots & 0 & a_{1} \\
0 & 1 & 0 & \cdots & 0 & a_{2} \\
\vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & \cdots & 1 & a_{n-1}
\end{array}\right)
$$

is called the companion matrix of the polynomial $p(x)$. If there is no possible ambiguity, we will denote the companion matrix simply by C . The companion matrix has several interesting properties that we will soon discover. We will also make use of the associated characteristic matrix $x I-C \in \mathrm{M}_{\mathrm{n}}(\mathcal{R})$ given by

$$
x I-C=\left(\begin{array}{rrlrcc}
x & 0 & \cdots & 0 & 0 & -a_{0} \\
-1 & x & \cdots & 0 & 0 & -a_{1} \\
\vdots & \vdots & & \vdots & \vdots & \vdots \\
0 & 0 & \cdots & -1 & x & -a_{n-2} \\
0 & 0 & \cdots & 0 & -1 & x-a_{n-1}
\end{array}\right)
$$

Our next theorem is quite useful.
Theorem 8.11 Let $p(x)=x^{n}-a_{n-1} x^{n-1}-\cdots-a_{0} \in \mathcal{F}[x]$ have companion matrix $C$. Then $\operatorname{det}(x I-C)=p(x)$.

Proof We proceed by induction on the degree of $\mathrm{p}(\mathrm{x})$. If $\mathrm{n}=1$, then $\mathrm{p}(\mathrm{x})=$ $\mathrm{x}-\mathrm{a}_{0}, \mathrm{C}=\left(\mathrm{a}_{0}\right)$ and $\mathrm{xI}-\mathrm{C}=\left(\mathrm{x}-\mathrm{a}_{0}\right)$ so that

$$
\operatorname{det}(x I-C)=x-a_{0}=p(x)
$$

Now assume that the theorem is true for all polynomials of degree less than $n$, and suppose $\operatorname{deg} \mathrm{p}(\mathrm{x})=\mathrm{n}>1$. If we expand $\operatorname{det}(\mathrm{xI}-\mathrm{C})$ by minors of the first row, we obtain (see Theorem 4.10)

$$
\operatorname{det}(\mathrm{xI}-\mathrm{C})=\mathrm{x} \operatorname{det} \mathrm{C}_{11}+\left(-\mathrm{a}_{0}\right)(-1)^{\mathrm{n}+1} \operatorname{det} \mathrm{C}_{1 \mathrm{n}}
$$

where the minor matrices $\mathrm{C}_{11}$ and $\mathrm{C}_{1 \mathrm{n}}$ are given by

$$
\begin{aligned}
& C_{11}=\left(\begin{array}{rrrrrc}
x & 0 & \cdots & 0 & 0 & -a_{1} \\
-1 & x & \cdots & 0 & 0 & -a_{2} \\
\vdots & \vdots & & \vdots & \vdots & \vdots \\
0 & 0 & \cdots & -1 & x & -a_{n-2} \\
0 & 0 & \cdots & 0 & -1 & x-a_{n-1}
\end{array}\right) \\
& C_{1 n}=\left(\begin{array}{rrrrrr}
-1 & x & 0 & \cdots & 0 & 0 \\
0 & -1 & x & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & \cdots & -1 & x \\
0 & 0 & 0 & \cdots & 0 & -1
\end{array}\right)
\end{aligned}
$$

Defining the polynomial $\mathrm{p}^{\prime}(\mathrm{x})=\mathrm{x}^{\mathrm{n}-1}-\mathrm{a}_{\mathrm{n}-1} \mathrm{x}^{\mathrm{n}-2}-\cdots-\mathrm{a}_{2} \mathrm{x}-\mathrm{a}_{1}$ along with its companion matrix $\mathrm{C}^{\prime}$, we see that $\mathrm{C}_{11}=\mathrm{xI}-\mathrm{C}^{\prime}$. By our induction hypothesis, it then follows that

$$
\operatorname{det} \mathrm{C}_{11}=\operatorname{det}\left(\mathrm{xI}-\mathrm{C}^{\prime}\right)=\mathrm{p}^{\prime}(\mathrm{x})
$$

Next we note that $\mathrm{C}_{1 \mathrm{n}}$ is an upper-triangular matrix, and hence (by Theorem 4.5) $\operatorname{det} \mathrm{C}_{1 \mathrm{n}}=(-1)^{\mathrm{n}-1}$. Putting all of this together we find that

$$
\operatorname{det}(x I-C)=x^{\prime}(x)-a_{0}=p(x)
$$

Recall that two matrix representations are similar if and only if they represent the same underlying operator in two different bases (see Theorem 5.18).

Theorem 8.12 (a) The companion matrix $\mathrm{C}=\mathrm{C}(\mathrm{p}(\mathrm{x}))$ of any monic polynomial $p(x) \in \mathcal{F}[x]$ has $p(x)$ as its minimal polynomial $m(x)$.
(b) If $\operatorname{dim} \mathrm{V}=\mathrm{n}$ and $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ has minimal polynomial $\mathrm{m}(\mathrm{x})$ of degree n , then $\mathrm{C}(\mathrm{m}(\mathrm{x}))$ represents T relative to some basis for V .

Proof (a) From the preceding proof, we see that deleting the first row and $\mathrm{n} t h$ column of $\mathrm{xI}-\mathrm{C}$ and taking the determinant yields $\operatorname{det} \mathrm{C}_{1 \mathrm{n}}=(-1)^{\mathrm{n}-1}$. Therefore $\mathrm{f}_{\mathrm{n}-1}(\mathrm{x})=1$ so that $\mathrm{q}_{1}(\mathrm{x})=\mathrm{q}_{2}(\mathrm{x})=\cdots=\mathrm{q}_{\mathrm{n}-1}(\mathrm{x})=1$. Hence C is nonderogatory (corollary to Theorem 8.10), so that by Theorem 8.11 we have $\mathrm{m}(\mathrm{x})=\mathrm{q}_{\mathrm{n}}(\mathrm{x})=\operatorname{det}(\mathrm{xI}-\mathrm{C})=\mathrm{p}(\mathrm{x})$.
(b) Note $\operatorname{dim} V=\operatorname{deg} \Delta_{T}(x)=n=\operatorname{deg} m(x)$ so that any [T] has similarity invariants $\mathrm{q}_{1}(\mathrm{x})=\cdots=\mathrm{q}_{\mathrm{n}-1}(\mathrm{x})=1$ and $\mathrm{q}_{\mathrm{n}}(\mathrm{x})=\mathrm{m}(\mathrm{x})$ (see Theorem 8.10 and its corollary). Since the proof of part (a) showed that $C=C(m(x))$ has the
same similarity invariants as [T], it follows from Corollary 1 of Theorem 8.9 that C and $[\mathrm{T}]$ are similar.

Note that Theorems 8.11 and 8.12(a) together show that the companion matrix is nonderogatory.

Given any $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$, we can interpret A as the matrix representation of a linear transformation T on an n -dimensional vector space V . If A has minimal polynomial $\mathrm{m}(\mathrm{x})$ with $\operatorname{deg} \mathrm{m}(\mathrm{x})=\mathrm{n}$, then so does T (by Theorem 7.1). Hence the companion matrix C of $\mathrm{m}(\mathrm{x})$ represents T relative to some basis for V (Theorem 8.12(b)). This means that A is similar to C (Theorem 5.18), and therefore $\mathrm{C}=\mathrm{P}^{-1} \mathrm{AP}$ for some nonsingular transition matrix $\mathrm{P} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$. But then

$$
\mathrm{xI}-\mathrm{C}=\mathrm{xI}-\mathrm{P}^{-1} \mathrm{AP}=\mathrm{P}^{-1}(\mathrm{xI}-\mathrm{A}) \mathrm{P}
$$

and hence $\operatorname{det}(x I-C)=\operatorname{det}(x I-A)$ by Theorem 4.8 and its corollary. Using Theorem 8.11, we then have the following result.

Theorem 8.13 Let $A \in M_{n}(\mathcal{F})$ have minimal polynomial $m(x)$ of degree $n$. Then $m(x)=\operatorname{det}(x I-A)$.

Our next theorem is a useful restatement of what we have done so far in this section.

Theorem 8.14 Let $p(x)=x^{n}-a_{n-1} x^{n-1}-\cdots-a_{0} \in \mathcal{F}[x]$. Then the companion matrix $\mathrm{C}(\mathrm{p}(\mathrm{x}))$ is nonderogatory, and its characteristic polynomial $\Delta_{\mathrm{C}}(\mathrm{x})$ and minimal polynomial $\mathrm{m}(\mathrm{x})$ both equal $\mathrm{p}(\mathrm{x})$. Moreover, $\mathrm{xI}-\mathrm{C}$ is equivalent over $\mathcal{P}$ to the n x n matrix (the Smith canonical form of $\mathrm{xI}-\mathrm{C}$ )

$$
\left(\begin{array}{ccccc}
1 & 0 & \cdots & 0 & 0 \\
\vdots & \vdots & & \vdots & \vdots \\
0 & 0 & \cdots & 1 & 0 \\
0 & 0 & \cdots & 0 & p(x)
\end{array}\right)
$$

For notational convenience, we sometimes write a diagonal matrix by listing its diagonal entries. For example, the matrix shown in the above theorem would be written as $\operatorname{diag}(1, \ldots, 1, p(x))$.

Theorem 8.15 If $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$, then A is similar over $\mathcal{F}$ to the direct sum of the companion matrices of its nonunit similarity invariants.

Proof The proof is an application of Theorem 8.9. Assume that A $\neq \mathrm{cI}$ (where $c \in \mathcal{F}$ ) or there is nothing to prove. Hence $\mathrm{f}_{1}(\mathrm{x})$, the first determinantal divisor of $x I-A$, must be 1 . But $f_{0}(x)=1$ by definition, and hence we have $\mathrm{f}_{1}(\mathrm{x})=\mathrm{q}_{1}(\mathrm{x}) \mathrm{f}_{0}(\mathrm{x})=\mathrm{q}_{1}(\mathrm{x})=1$. Since at least $\mathrm{q}_{1}(\mathrm{x})=1$, we now assume that in fact the first $k$ similarity invariants of $A$ are equal to 1 . In other words, we assume that $\mathrm{q}_{1}(\mathrm{x})=\cdots=\mathrm{q}_{\mathrm{k}}(\mathrm{x})=1$, and then $\operatorname{deg} \mathrm{q}_{\mathrm{i}}(\mathrm{x})=\mathrm{d}_{\mathrm{i}} \geq 1$ for $\mathrm{i}=\mathrm{k}+1$, ..., n.

Since $\mathrm{f}_{\mathrm{n}}(\mathrm{x})=\mathrm{q}_{1}(\mathrm{x}) \cdots \mathrm{q}_{\mathrm{n}}(\mathrm{x})$, Theorem 6.2(b) tells us that $\operatorname{deg} \mathrm{f}_{\mathrm{n}}(\mathrm{x})=$ $\sum_{j=1}^{n} \operatorname{deg} q_{j}(x)$ and hence (using $\operatorname{deg} \mathrm{q}_{\mathrm{j}}=0$ for $\mathrm{j}=1, \ldots, \mathrm{k}$ )

$$
n=\operatorname{deg} \Delta_{A}(x)=\operatorname{deg} f_{n}(x)=\sum_{j=k+1}^{n} \operatorname{deg} q_{j}(x)=\sum_{j=k+1}^{n} d_{j} .
$$

Let $\mathrm{Q}_{\mathrm{i}}=\mathrm{C}\left(\mathrm{q}_{\mathrm{i}}(\mathrm{x})\right) \in \mathrm{M}_{\mathrm{d}_{\mathrm{i}}}(\mathcal{P})$ for $\mathrm{i}=\mathrm{k}+1, \ldots, \mathrm{n}$. We want to show that $\mathrm{xI}-\mathrm{A}$ is equivalent over $\mathcal{P}$ to

$$
\mathrm{xI}-\left(\mathrm{Q}_{\mathrm{k}+1} \oplus \cdots \oplus \mathrm{Q}_{\mathrm{n}}\right)=\left(\mathrm{xI}-\mathrm{Q}_{\mathrm{k}+1}\right) \oplus \cdots \oplus\left(\mathrm{xI}-\mathrm{Q}_{\mathrm{n}}\right)
$$

(Note that each of the identity matrices in this equation may be of a different size.)

It should be clear that the Smith form of $\mathrm{xI}-\mathrm{A}$ is the diagonal nxn matrix

$$
(x I-A)_{S}=\operatorname{diag}\left(q_{1}(x), \ldots, q_{n}(x)\right)=\operatorname{diag}\left(1, \ldots, 1, q_{k+1}(x), \ldots, q_{n}(x)\right)
$$

From Theorem 8.14, we know that $\left(x I-Q_{i}\right)_{S}=\operatorname{diag}\left(1, \ldots, 1, q_{i}(x)\right) \in$ $\mathrm{M}_{\mathrm{d}_{\mathrm{i}}}(\mathcal{P})$. Since $\sum_{\mathrm{i}=\mathrm{k}+1}^{\mathrm{d}} \mathrm{d}_{\mathrm{i}}=\mathrm{n}$, we now see that by suitable row and column interchanges we have

$$
\begin{equation*}
x I-A \approx(x I-A)_{S} \approx\left(x I-Q_{k+1}\right)_{S} \oplus \cdots \oplus\left(x I-Q_{n}\right)_{S} \tag{*}
\end{equation*}
$$

where $\approx$ denotes equivalence over $\mathcal{P}$.
If we write $\left(x I-Q_{i}\right)_{S}=E_{i}\left(x I-Q_{i}\right) F_{i}$ where $E_{i}$ and $F_{i}$ are unit matrices, then (by multiplying out the block diagonal matrices) it is easy to see that

$$
\begin{aligned}
& \mathrm{E}_{\mathrm{k}+1}\left(\mathrm{xI}-\mathrm{Q}_{\mathrm{k}+1}\right) \mathrm{F}_{\mathrm{k}+1} \oplus \cdots \oplus \mathrm{E}_{\mathrm{n}}\left(\mathrm{xI}-\mathrm{Q}_{\mathrm{n}}\right) \mathrm{F}_{\mathrm{n}} \\
& \quad=\left[\mathrm{E}_{\mathrm{k}+1} \oplus \cdots \oplus \mathrm{E}_{\mathrm{n}}\right]\left[\left(\mathrm{xI}-\mathrm{Q}_{\mathrm{k}+1}\right) \oplus \cdots \oplus\left(\mathrm{xI}-\mathrm{Q}_{\mathrm{n}}\right)\right]\left[\mathrm{F}_{\mathrm{k}+1} \oplus \cdots \oplus \mathrm{~F}_{\mathrm{n}}\right]
\end{aligned}
$$

Since the direct sum of unit matrices is clearly a unit matrix (so that both $\mathrm{E}_{\mathrm{k}+1} \oplus \cdots \oplus \mathrm{E}_{\mathrm{n}}$ and $\mathrm{F}_{\mathrm{k}+1} \oplus \cdots \oplus \mathrm{~F}_{\mathrm{n}}$ are unit matrices), this shows that the right hand side of $(*)$ is equivalent to $\left(\mathrm{xI}-\mathrm{Q}_{\mathrm{k}+1}\right) \oplus \cdots \oplus\left(\mathrm{xI}-\mathrm{Q}_{\mathrm{n}}\right)$. (Note we have shown that if $\left\{\mathrm{S}_{\mathrm{i}}\right\}$ and $\left\{\mathrm{T}_{\mathrm{i}}\right\}$ are finite collections of matrices such that $\mathrm{S}_{\mathrm{i}} \approx \mathrm{T}_{\mathrm{i}}$, then it follows that $\mathrm{S}_{1} \oplus \cdots \oplus \mathrm{~S}_{\mathrm{n}} \approx \mathrm{T}_{1} \oplus \cdots \oplus \mathrm{~T}_{\mathrm{n}}$.) Therefore xI-A
is equivalent to $\mathrm{xI}-\left(\mathrm{Q}_{\mathrm{k}+1} \oplus \cdots \oplus \mathrm{Q}_{\mathrm{n}}\right)$ which is what we wanted to show. The theorem now follows directly from Theorem 8.9.

We are now in a position to prove the rational canonical form theorem. Note that the name is derived from the fact that the rational form of a matrix is obtained by the application of a finite number of rational operations (which essentially constitute the Smith algorithm).

Theorem 8.16 (Rational Canonical Form) A matrix $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ is similar over $\mathcal{F}$ to the direct sum of the companion matrices of the elementary divisors of $\mathrm{xI}-\mathrm{A}$.

Proof As in the proof of Theorem 8.15, we assume that the first k similarity invariants of $A$ are $q_{1}(x)=\cdots=q_{k}(x)=1$ and that $\operatorname{deg} q_{i}(x)=d_{i} \geq 1$ for $i=$ $\mathrm{k}+1, \ldots, \mathrm{n}$. Changing notation slightly from our first definition, we write each nonunit invariant factor as a product of powers of prime polynomials (i.e., as a product of elementary divisors): $q_{i}(x)=e_{i 1}(x) \cdots e_{i m_{i}}(x)$ for each $i=$ $\mathrm{k}+1, \ldots, \mathrm{n}$. From Theorem 8.14, we know that $\mathrm{xI}-\mathrm{Q}_{\mathrm{i}}=\mathrm{xI}-\mathrm{C}\left(\mathrm{q}_{\mathrm{i}}(\mathrm{x})\right)$ is $\mathcal{P}$ equivalent to the $\mathrm{d}_{\mathrm{i}} \times \mathrm{d}_{\mathrm{i}}$ matrix

$$
\mathrm{B}_{\mathrm{i}}=\operatorname{diag}\left(1, \ldots, 1, \mathrm{q}_{\mathrm{i}}(\mathrm{x})\right) .
$$

Similarly, if $\mathrm{c}_{\mathrm{ij}}=\operatorname{deg} \mathrm{e}_{\mathrm{ij}}(\mathrm{x})$, each $\mathrm{xI}-\mathrm{C}\left(\mathrm{e}_{\mathrm{ij}}(\mathrm{x})\right)\left(\mathrm{j}=1, \ldots, \mathrm{~m}_{\mathrm{i}}\right)$ is $\mathcal{P}$-equivalent to a $\mathrm{c}_{\mathrm{ij}} \mathrm{X} \mathrm{c}_{\mathrm{ij}}$ matrix

$$
\mathrm{D}_{\mathrm{ij}}=\operatorname{diag}\left(1, \ldots, 1, \mathrm{e}_{\mathrm{ij}}(\mathrm{x})\right)
$$

Since $\operatorname{deg} \mathrm{q}_{\mathrm{i}}(\mathrm{x})=\sum_{\mathrm{j}} \operatorname{deg} \mathrm{e}_{\mathrm{ij}}(\mathrm{x})$, it follows that the block diagonal matrix

$$
\begin{aligned}
\mathrm{D}_{\mathrm{i}} & =\mathrm{D}_{\mathrm{i} 1} \oplus \cdots \oplus \mathrm{D}_{\mathrm{im}}^{\mathrm{i}} \\
& =\operatorname{diag}\left(1, \ldots, 1, \mathrm{e}_{\mathrm{i} 1}(\mathrm{x})\right) \oplus \operatorname{diag}\left(1, \ldots, 1, \mathrm{e}_{\mathrm{i} 2}(\mathrm{x})\right) \\
& \oplus \cdots \oplus \operatorname{diag}\left(1, \ldots, 1, \mathrm{e}_{\mathrm{im}}(\mathrm{x})\right)
\end{aligned}
$$

is also a $\mathrm{d}_{\mathrm{i}} \times \mathrm{d}_{\mathrm{i}}$ matrix. We first show that $\mathrm{B}_{\mathrm{i}}$ (and hence also $\mathrm{xI}-\mathrm{Q}_{\mathrm{i}}$ ) is $\mathcal{P}$ equivalent to $D_{i}$.

Consider the collection of all $\left(d_{i}-1\right) x\left(d_{i}-1\right)$ subdeterminants of $D_{i}$. For each $r=1, \ldots, m_{i}$, this collection will contain that subdeterminant obtained by deleting the row and column containing $\mathrm{e}_{\mathrm{ir}}$. In particular, this subdeterminant will be $\prod_{j \neq r} e_{i j}$. But the gcd of all such subdeterminants taken over r (for a fixed i of course) is just 1 . (To see this, consider the product abcd. If we look at the collection of products obtained by deleting one of $a, b, c$ or $d$ we obtain \{bcd, acd, abd, abc\}. Since there is no factor in common with all four of these
products, it follows that the gcd of this collection is 1.) Therefore the $\left(d_{i}-\right.$ 1) th determinantal divisor $f_{d_{i}-1}(x)$ of $D_{i}$ is 1 , and hence the fact that $f_{k-1}(x)$ divides $\mathrm{f}_{\mathrm{k}}(\mathrm{x})$ means $\mathrm{f}_{1}(\mathrm{x})=\cdots=\mathrm{f}_{\mathrm{d}_{\mathrm{i}}-1}(\mathrm{x})=1$ and $\mathrm{f}_{\mathrm{d}_{\mathrm{i}}}(\mathrm{x})=\Pi_{\mathrm{j}} \mathrm{e}_{\mathrm{ij}}(\mathrm{x})=\mathrm{q}_{\mathrm{i}}(\mathrm{x})$. From the definition of determinantal divisor (or the definition of invariant factor along with the fact that $B_{i}$ is in its Smith canonical form), it is clear that $B_{i}$ has precisely these same determinantal divisors, and hence (by the corollary to Theorem 8.8) $\mathrm{B}_{\mathrm{i}}$ must be $\mathcal{P}$-equivalent to $\mathrm{D}_{\mathrm{i}}$.

All that remains is to put this all together and apply Theorem 8.9 again. We now take the direct sum of each side of the equivalence relation xI $-\mathrm{Q}_{\mathrm{i}} \approx$ $\mathrm{B}_{\mathrm{i}} \approx \mathrm{D}_{\mathrm{i} 1} \oplus \cdots \oplus \mathrm{D}_{\mathrm{im}}=\mathrm{D}_{\mathrm{i}}$ using the fact that (as we saw in the proof of Theorem 8.15) $\left(\mathrm{xI}-\mathrm{Q}_{\mathrm{k}+1}\right) \oplus \cdots \oplus\left(\mathrm{xI}-\mathrm{Q}_{\mathrm{n}}\right) \approx \mathrm{D}_{\mathrm{k}+1} \oplus \cdots \oplus \mathrm{D}_{\mathrm{n}}$. It will be convenient to denote direct sums by $\Sigma \oplus$. For example, we have already seen it is true in general that

$$
\sum_{i=k+1}^{n} \oplus\left(x I-Q_{i}\right)=x I-\sum_{i=k+1}^{n} \oplus Q_{i}
$$

(where we again remark that the identity matrices in this equation may be of different dimensions). Therefore, we have shown that

$$
\begin{aligned}
x I-\sum_{i=k+1}^{n} \oplus Q_{i} & =\sum_{i=k+1}^{n} \oplus\left(x I-Q_{i}\right) \approx \sum_{i=k+1}^{n} \oplus\left(D_{i 1} \oplus \cdots \oplus D_{i m_{i}}\right) \\
& =\sum_{i=k+1}^{n} \oplus\left(\sum_{j=1}^{m_{i}} \oplus D_{i j}\right) \approx \sum_{i=k+1}^{n} \oplus\left(\sum_{j=1}^{m_{i}} \oplus\left[x I-C\left(e_{i j}(x)\right)\right]\right) \\
& =x I-\sum_{i=k+1}^{n} \oplus\left(\sum_{j=1}^{m_{i}} \oplus C\left(e_{i j}(x)\right)\right)
\end{aligned}
$$

and hence $\sum_{i=\mathrm{k}+1}^{\mathrm{n}} \oplus \mathrm{Q}_{\mathrm{i}}$ is similar over $\mathcal{F}$ to $\sum_{i=k+1}^{n} \oplus\left[\sum_{j=1}^{m_{i}} \oplus C\left(e_{i j}(x)\right)\right]$. But Theorem 8.15 tells us that A is similar over $\mathcal{F}$ to $\sum_{\mathrm{i}=\mathrm{k}+1} \oplus \mathrm{Q}_{\mathrm{i}}$, and therefore we have shown that A is similar over $\mathcal{F}$ to

$$
\sum_{i=k+1}^{n} \oplus\left(\sum_{j=1}^{m_{i}} \oplus C\left(e_{i j}(x)\right)\right)
$$

Example 8.8 Consider the polynomial

$$
p(x)=(x-1)^{2}\left(x^{2}+1\right)^{2}=x^{6}-2 x^{5}+3 x^{4}-4 x^{3}+3 x^{2}-2 x+1 .
$$

Its companion matrix is

$$
C=\left(\begin{array}{rrrrrr}
0 & 0 & 0 & 0 & 0 & -1 \\
1 & 0 & 0 & 0 & 0 & 2 \\
0 & 1 & 0 & 0 & 0 & -3 \\
0 & 0 & 1 & 0 & 0 & 4 \\
0 & 0 & 0 & 1 & 0 & -3 \\
0 & 0 & 0 & 0 & 1 & 2
\end{array}\right) \in \mathrm{M}_{6}(\mathbb{R})
$$

According to Theorem 8.14, C is nonderogatory and its minimal polynomial is $p(x)$. Then by Theorem 8.10 and its corollary, the only nonunit similarity invariant of C is also $\mathrm{p}(\mathrm{x})$. This means that C is already in the form given by Theorem 8.15.

The elementary divisors (in $\mathbb{R}[x]$ ) of $\mathrm{xI}-\mathrm{C}$ are

$$
\mathrm{e}_{1}(\mathrm{x})=(\mathrm{x}-1)^{2}=\mathrm{x}^{2}-2 \mathrm{x}+1
$$

and

$$
\mathrm{e}_{2}(\mathrm{x})=\left(\mathrm{x}^{2}+1\right)^{2}=\mathrm{x}^{4}+2 \mathrm{x}^{2}+1
$$

These have the companion matrices

$$
\begin{gathered}
C\left(e_{1}(x)\right)=\left(\begin{array}{rr}
0 & -1 \\
0 & 2
\end{array}\right) \\
C\left(e_{2}(x)\right)=\left(\begin{array}{rrrr}
0 & 0 & 0 & -1 \\
1 & 0 & 0 & 0 \\
0 & 1 & 0 & -2 \\
0 & 0 & 1 & 0
\end{array}\right)
\end{gathered}
$$

and hence Theorem 8.16 tells us that C is similar over $\mathbb{R}$ to the direct sum $C\left(e_{1}(x)\right) \oplus C\left(e_{2}(x)\right)$. We leave it to the reader to find the rational canonical form of C if we regard it as a matrix over $\mathbb{C}[\mathrm{x}]$. //

## Exercises

1. Prove Corollary 1 of Theorem 7.24 using the rational canonical form.
2. (a) Let V be a real 6 -dimensional vector space, and suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ has minimal polynomial $m(x)=\left(x^{2}-x+3\right)(x-2)^{2}$. Write down all possible rational canonical forms for T (except for the order of the blocks).
(b) Let V be a real 7 -dimensional vector space, and suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ has minimal polynomial $m(x)=\left(x^{2}+2\right)(x+3)^{3}$. Write down all possible rational canonical forms for T (except for the order of the blocks).
3. Let A be a 4 x 4 matrix with minimal polynomial $\mathrm{m}(\mathrm{x})=\left(\mathrm{x}^{2}+1\right)\left(\mathrm{x}^{2}-3\right)$. Find the rational canonical form if A is a matrix over:
(a) The rational field $\mathbb{Q}$.
(b) The real field $\mathbb{R}$.
(c) The complex field $\mathbb{C}$.
4. Find the rational canonical form for the Jordan block

$$
\left(\begin{array}{llll}
a & 1 & 0 & 0 \\
0 & a & 1 & 0 \\
0 & 0 & a & 1 \\
0 & 0 & 0 & a
\end{array}\right)
$$

5. Find a $3 \times 3$ matrix $A$ with integral entries such that $A^{3}+3 A^{2}+2 A+2=$ 0 . Prove that your matrix satisfies this identity.
6. Discuss the validity of each of the following assertions:
(a) Two square matrices are similar if and only if they have the same eigenvalues (including multiplicities).
(b) Two square matrices are similar if and only if they have the same minimal polynomial.
(c) Two square matrices are similar if and only if they have the same elementary divisors.
(d) Two square matrices are similar if and only if they have the same determinantal divisors.
7. Suppose $\mathrm{A}=\mathrm{B} \oplus \mathrm{C}$ where B and C are square matrices. Is the list of elementary divisors of A equal to the list of elementary divisors of B concatenated with (i.e., "added on to") the list of elementary divisors of C? What if "elementary divisors" is replaced by "invariant factors" or "determinantal divisors" in this statement?

### 8.6 THE JORDAN CANONICAL FORM

We have defined a canonical form as that matrix representation A of a linear transformation $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ that is of a particularly simple form in some basis for V . If all the eigenvalues of T lie in the base field $\mathcal{F}$, then the minimal polynomial $\mathrm{m}(\mathrm{x})$ for T will factor into a product of linear terms. In addition, if the eigenvalues are all distinct, then T will be diagonalizable (Theorem 7.24). But in the general case of repeated roots, we must (so far) fall back to the triangular form described in Chapter 7 and in Section 8.1. However, in this more general case there is another very important form that follows easily from what we have already done. If $A \in \mathrm{M}_{\mathrm{n}}(\mathbb{C})$, then (by Theorem 6.13) all the elementary divisors of $x I-A$ will be of the simple form $(x-a)^{k}$. We shall now investigate the "simplest" form that such an A can take.

To begin with, given a polynomial $\mathrm{p}(\mathrm{x})=\left(\mathrm{x}-\mathrm{a}_{0}\right)^{\mathrm{n}} \in \mathcal{F}[\mathrm{x}]$, we define the hypercompanion matrix $\mathrm{H}(\mathrm{p}(\mathrm{x})) \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ to be the upper-triangular matrix

$$
\left(\begin{array}{cccccc}
a_{0} & 1 & 0 & \cdots & 0 & 0 \\
0 & a_{0} & 0 & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & \cdots & a_{0} & 1 \\
0 & 0 & 0 & \cdots & 0 & a_{0}
\end{array}\right)
$$

A matrix of this form is also referred to as a basic Jordan block belonging to $\mathrm{a}_{0}$. Now consider the characteristic matrix xI-H(p(x)). Note that if we delete the n th row and first column of this characteristic matrix, we obtain a lowertriangular matrix with all diagonal entries equal to -1 , and hence its determinant is equal to $(-1)^{\mathrm{n}-1}$. Thus the corresponding determinantal divisor $\mathrm{f}_{\mathrm{n}-1}(\mathrm{x})$ is equal to 1 , and therefore $f_{1}(x)=\cdots=f_{n-1}(x)=1$ (because $\left.f_{k-1}(x) \mid f_{k}(x)\right)$. Using $\mathrm{f}_{\mathrm{k}}(\mathrm{x})=\mathrm{q}_{\mathrm{k}}(\mathrm{x}) \mathrm{f}_{\mathrm{k}-1}(\mathrm{x})$, it follows that $\mathrm{q}_{1}(\mathrm{x})=\cdots=\mathrm{q}_{\mathrm{n}-1}(\mathrm{x})=1$, and thus H is nonderogatory (corollary to Theorem 8.10). Since it is obvious that $\Delta_{H}(x)=$ $\left(\mathrm{x}-\mathrm{a}_{0}\right)^{\mathrm{n}}=\mathrm{p}(\mathrm{x})$, we conclude that $\Delta_{\mathrm{H}}(\mathrm{x})=\mathrm{m}(\mathrm{x})=\mathrm{p}(\mathrm{x})$. (Alternatively, by Theorem 8.10, we see that the minimal polynomial for $H$ is $q_{n}(x)=f_{n}(x)=$ $(x-a)^{n}$ which is also just the characteristic polynomial of H.) Along with the definition of the Smith canonical form, this proves the following result analogous to Theorem 8.14.

Theorem 8.17 The hypercompanion matrix $\mathrm{H}(\mathrm{p}(\mathrm{x}))$ of the polynomial $\mathrm{p}(\mathrm{x})=$ $(\mathrm{x}-\mathrm{a})^{\mathrm{n}} \in \mathcal{F}[\mathrm{x}]$ is nonderogatory, and its characteristic and minimal polynomials both equal $p(x)$. Furthermore, the Smith form of $x I-H(p(x))$ is

$$
\left(\begin{array}{ccccc}
1 & 0 & \cdots & 0 & 0 \\
0 & 1 & \cdots & 0 & 0 \\
\vdots & \vdots & & \vdots & \vdots \\
0 & 0 & \cdots & 1 & 0 \\
0 & 0 & \cdots & 0 & p(x)
\end{array}\right) .
$$

Theorems 8.14 and 8.17 show that given the polynomial $p(x)=(x-a)^{n} \in$ $\mathbb{C}[\mathrm{x}]$, both $\mathrm{C}=\mathrm{C}(\mathrm{p}(\mathrm{x}))$ and $\mathrm{H}=\mathrm{H}(\mathrm{p}(\mathrm{x}))$ have precisely the same similarity invariants. Using Theorem 8.16, we then see that C and H are similar over $\mathbb{C}$. Now, if $A \in M_{n}(\mathbb{C})$, we know that the elementary divisors of xI - A will be of the form $(x-a)^{k}$. Furthermore, Theorem 8.16 shows us that A is similar over $\mathbb{C}$ to the direct sum of the companion matrices of these elementary divisors. But each companion matrix is similar over $\mathbb{C}$ to the corresponding hypercompanion matrix, and hence A is similar over $\mathbb{C}$ to the direct sum of the hypercompanion matrices of the elementary divisors of xI - A.

It may be worth briefly showing that the notions of similarity and direct sums may be treated in the manner just claimed. In other words, denoting similarity over $\mathbb{C}$ by $\sim$, we suppose that $A \sim \mathrm{C}_{1} \oplus \mathrm{C}_{2}=\mathrm{S}^{-1} \mathrm{AS}$ for some nonsingular matrix $S \in M_{n}(\mathbb{C})$. We now also assume that $C_{i} \sim H_{i}=T_{i}{ }^{-1} C_{i} T_{i}$ for each $\mathrm{i}=1,2$. Then we see that

$$
\begin{aligned}
\left(\begin{array}{cc}
H_{1} & 0 \\
0 & H_{2}
\end{array}\right) & =\left(\begin{array}{cc}
T_{1}^{-1} C_{1} T_{1} & 0 \\
0 & T_{2}^{-1} C_{2} T_{2}
\end{array}\right) \\
& =\left(\begin{array}{cc}
T_{1}^{-1} & 0 \\
0 & T_{2}^{-1}
\end{array}\right)\left(\begin{array}{cc}
C_{1} & 0 \\
0 & C_{2}
\end{array}\right)\left(\begin{array}{cc}
T_{1} & 0 \\
0 & T_{2}
\end{array}\right)
\end{aligned}
$$

which (in an obvious shorthand notation) may be written in the form $\mathrm{H}=$ $\mathrm{T}^{-1} \mathrm{CT}$ if we note that

$$
\left(\begin{array}{cc}
T_{1}^{-1} & 0 \\
0 & T_{2}^{-1}
\end{array}\right)=\left(\begin{array}{cc}
T_{1} & 0 \\
0 & T_{2}
\end{array}\right)^{-1}
$$

We therefore have $\mathrm{H}=\mathrm{T}^{-1} \mathrm{CT}=\mathrm{T}^{-1} \mathrm{~S}^{-1} \mathrm{AST}=(\mathrm{ST})^{-1} \mathrm{~A}(\mathrm{ST})$ which shows that A is indeed similar to the direct sum of the hypercompanion matrices. In any case, we have proved the difficult part of the next very important theorem (see also Theorem 7.42).

Theorem 8.18 (Jordan Canonical Form) If $A \in M_{n}(\mathbb{C})$, then $A$ is similar over $\mathbb{C}$ to the direct sum of the hypercompanion matrices of all the elementary divisors of $\mathrm{xI}-\mathrm{A}$, and this direct sum is unique except for the order of the blocks. Moreover, the numbers appearing on the main diagonal of the Jordan form are precisely the eigenvalues of A . (Note that the field $\mathbb{C}$ can be replaced by an arbitrary field $\mathcal{F}$ if all the eigenvalues of A lie in $\mathcal{F}$.)

Proof Existence was proved in the above discussion, so we now consider uniqueness. According to our general prescription, given a matrix $A \in M_{n}(\mathbb{C})$, we would go through the following procedure to find its Jordan form. First we reduce the characteristic matrix xI - A to its unique Smith form, thus obtaining the similarity invariants of A. These similarity invariants are then factored (over $\mathbb{C}$ ) to obtain the elementary divisors of xI - A. Finally, the corresponding hypercompanion matrices are written down, and the Jordan form of A is just their direct sum.

All that remains is to prove the statement about the eigenvalues of A . To see this, recall that the eigenvalues of A are the roots of the characteristic polynomial $\operatorname{det}(x I-A)$. Suppose that $J=S^{-1} A S$ is the Jordan form of A. Then the eigenvalues of J are the roots of

$$
\operatorname{det}(x I-J)=\operatorname{det}\left(x I-S^{-1} A S\right)=\operatorname{det}\left[S^{-1}(x I-A) S\right]=\operatorname{det}(x I-A)
$$

so that A and J have the same eigenvalues. But J is an upper-triangular matrix, and hence the roots of $\operatorname{det}(\mathrm{xI}-\mathrm{J})$ are precisely the diagonal entries of J .

Example 8.9 Referring to Example 8.8, we regard C as a matrix over $\mathrm{M}_{6}(\mathbb{C})$. Then its elementary divisors are $\mathrm{e}_{1}(\mathrm{x})=(\mathrm{x}-1)^{2}, \mathrm{e}_{2}(\mathrm{x})=(\mathrm{x}+i)^{2}$ and $\mathrm{e}_{3}(\mathrm{x})=$ $(\mathrm{x}-i)^{2}$. The corresponding hypercompanion matrices are

$$
\begin{aligned}
& H_{1}=H\left(e_{1}(x)\right)=\left(\begin{array}{ll}
1 & 1 \\
0 & 1
\end{array}\right) \\
& H_{2}=H\left(e_{2}(x)\right)=\left(\begin{array}{rr}
-i & 1 \\
0 & -i
\end{array}\right) \\
& H_{3}=H\left(e_{3}(x)\right)=\left(\begin{array}{ll}
i & 1 \\
0 & i
\end{array}\right)
\end{aligned}
$$

and therefore $A$ is similar over $\mathbb{C}$ to its Jordan form $\mathrm{H}_{1} \oplus \mathrm{H}_{2} \oplus \mathrm{H}_{3}$. //

Our next theorem is really a corollary to Theorem 8.18 , but it is a sufficiently important result that we single it out by itself.

Theorem 8.19 The geometric multiplicity of an eigenvalue $\lambda_{i}$ (i.e., $\operatorname{dim} V_{\lambda_{i}}$ ) belonging to a matrix $A \in M_{n}(\mathbb{C})$ is the number of elementary divisors of the characteristic matrix xI - A that correspond to $\lambda_{i}$. In other words, the number of basic Jordan blocks (i.e., hypercompanion matrices) belonging to $\lambda_{\mathrm{i}}$ in the Jordan canonical form of A is the same as the geometric multiplicity of $\lambda_{\mathrm{i}}$.

Proof Suppose that there are $\mathrm{n}_{\mathrm{i}}$ elementary divisors belonging to $\lambda_{\mathrm{i}}$, and let $\left\{\mathrm{H}_{\mathrm{i} 1}, \ldots, \mathrm{H}_{\mathrm{in}_{\mathrm{i}}}\right\}$ be the corresponding hypercompanion matrices. By suitably numbering the eigenvalues, we may write the Jordan form of A as

$$
\mathrm{A}=\mathrm{H}_{11} \oplus \cdots \oplus \mathrm{H}_{1 \mathrm{n}_{1}} \oplus \cdots \oplus \mathrm{H}_{\mathrm{rl}} \oplus \cdots \oplus \mathrm{H}_{\mathrm{rn}_{\mathrm{r}}}
$$

where we assume that there are r distinct eigenvalues of A . For definiteness, let us arbitrarily consider the eigenvalue $\lambda_{1}$ and look at the matrix $\lambda_{1} \mathrm{I}-\mathrm{A}$. Since $\lambda_{1}-\lambda_{i} \neq 0$ for $i \neq 1$, this matrix takes the form

$$
\lambda_{1} \mathrm{I}-\mathrm{A}=\mathrm{B}_{11} \oplus \cdots \oplus \mathrm{~B}_{1 \mathrm{n}_{1}} \oplus \mathrm{~J}_{21} \oplus \cdots \oplus \mathrm{~J}_{2 \mathrm{n}_{2}} \oplus \cdots \oplus \mathrm{~J}_{\mathrm{r} 1} \oplus \cdots \oplus \mathrm{~J}_{\mathrm{rn}_{\mathrm{r}}}
$$

where each $\mathrm{B}_{\mathrm{ij}}$ is of the form

$$
\left(\begin{array}{rrrrr}
0 & -1 & 0 & \cdots & 0 \\
0 & 0 & -1 & \cdots & 0 \\
\vdots & \vdots & \vdots & & \vdots \\
0 & 0 & 0 & \cdots & -1 \\
0 & 0 & 0 & \cdots & 0
\end{array}\right)
$$

and each $\mathrm{J}_{\mathrm{ij}}$ looks like

$$
\left(\begin{array}{cccccc}
\lambda_{1}-\lambda_{i} & -1 & 0 & \cdots & 0 & 0 \\
0 & \lambda_{1}-\lambda_{i} & -1 & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & \cdots & \lambda_{1}-\lambda_{i} & -1 \\
0 & 0 & 0 & \cdots & 0 & \lambda_{1}-\lambda_{i}
\end{array}\right)
$$

It should be clear that each $\mathrm{J}_{\mathrm{ij}}$ is nonsingular (since they are all equivalent to the identity matrix of the appropriate size), and that each $\mathrm{B}_{\mathrm{ij}}$ has rank equal to one less than its size. Since $A$ is of size $n$, this means that the rank of $\lambda_{1} I-A$
is $\mathrm{n}-\mathrm{n}_{1}$ (just look at the number of linearly independent rows in $\lambda_{1} \mathrm{I}-\mathrm{A}$ ). But from Theorem 5.6 we have

$$
\operatorname{dim} V_{\lambda_{1}}=\operatorname{dim} \operatorname{Ker}\left(\lambda_{1} \mathrm{I}-\mathrm{A}\right)=\operatorname{nul}\left(\lambda_{1} \mathrm{I}-\mathrm{A}\right)=\mathrm{n}-\mathrm{r}\left(\lambda_{1} \mathrm{I}-\mathrm{A}\right)=\mathrm{n}_{1}
$$

In other words, the geometric multiplicity of $\lambda_{1}$ is equal to the number of hypercompanion matrices corresponding to $\lambda_{1}$ in the Jordan form of A. Since $\lambda_{1}$ could have been any of the eigenvalues, we are finished.

Example 8.10 Suppose $A \in M_{6}(\mathbb{C})$ has characteristic polynomial

$$
\Delta_{\mathrm{A}}(\mathrm{x})=(\mathrm{x}-2)^{4}(\mathrm{x}-3)^{2}
$$

and minimal polynomial

$$
\mathrm{m}(\mathrm{x})=(\mathrm{x}-2)^{2}(\mathrm{x}-3)^{2}
$$

Then A has eigenvalue $\lambda_{1}=2$ with multiplicity 4 , and $\lambda_{2}=3$ with multiplicity 2 , and these must lie along the diagonal of the Jordan canonical form. We know that (see the proof of the corollary to Theorem 8.10)

$$
\Delta_{\mathrm{A}}(\mathrm{x})=\mathrm{m}(\mathrm{x}) \mathrm{q}_{\mathrm{n}-1}(\mathrm{x}) \cdots \mathrm{q}_{1}(\mathrm{x})
$$

where $\mathrm{q}_{\mathrm{n}}(\mathrm{x})=\mathrm{m}(\mathrm{x}), \ldots, \mathrm{q}_{1}(\mathrm{x})$ are the similarity invariants of A , and that the elementary divisors of $x I-A$ are the powers of the prime factors of the $q_{i}(x)$. What we do not know however, is whether the set of elementary divisors of $\mathrm{xI}-\mathrm{A}$ is $\left\{(\mathrm{x}-2)^{2},(\mathrm{x}-3)^{2},(\mathrm{x}-2)^{2}\right\}$ or $\left\{(\mathrm{x}-2)^{2},(\mathrm{x}-3)^{2}, \mathrm{x}-2, \mathrm{x}-2\right\}$.

Using Theorem 8.18 , we then see that the only possible Jordan canonical forms are (up to the order of the blocks)

Note that in the first case, the geometric multiplicity of $\lambda_{1}=2$ is two, while in the second case, the geometric multiplicity of $\lambda_{1}=2$ is three. In both cases, the eigenspace corresponding to $\lambda_{2}=3$ is of dimension 1 . //

Example 8.11 Let us determine all possible Jordan canonical forms for the matrix $\mathrm{A} \in \mathbb{C}(3)$ given by

$$
\left(\begin{array}{rrr}
2 & a & b \\
0 & 2 & c \\
0 & 0 & -1
\end{array}\right) .
$$

The characteristic polynomial for A is easily seen to be

$$
\Delta_{\mathrm{A}}(\mathrm{x})=(\mathrm{x}-2)^{2}(\mathrm{x}+1)
$$

and hence (by Theorem 7.12) the minimal polynomial is either the same as $\Delta_{\mathrm{A}}(\mathrm{x})$, or is just $(\mathrm{x}-2)(\mathrm{x}+1)$. If $\mathrm{m}(\mathrm{x})=\Delta(\mathrm{x})$, then (using Theorem 8.18 again) the Jordan form must be

while in the second case, it must be

$$
\left(\begin{array}{ccc}
\boxed{2} & & \\
& \boxed{2} & \\
& & \boxed{-1}
\end{array}\right)
$$

If A is to be diagonalizable, then (either by Theorem 7.26 or the fact that the Jordan form in the second case is already diagonal) we must have the second case, and hence

$$
0=m(A)=(A-2 I)(A+I)=\left(\begin{array}{ccc}
0 & 3 a & a c \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right)
$$

so that A will be diagonalizable if and only if $\mathrm{a}=0 . / /$
As another application of Theorem 8.16 we have the following useful result. Note that here the field $\mathcal{F}$ can be either $\mathbb{R}$ or $\mathbb{C}$, and need not be algebraically closed in general.

Theorem 8.20 Suppose $\mathrm{B}_{\mathrm{i}} \in \mathrm{M}_{\mathrm{n}_{\mathrm{i}}}(\mathcal{F})$ for $\mathrm{i}=1, \ldots, \mathrm{r}$ and let $\mathrm{A}=\mathrm{B}_{1} \oplus \cdots \oplus$ $B_{r} \in M_{n}(\mathcal{F})$ (so that $n=\sum_{1=1}^{r} n_{i}$ ). Then the set of elementary divisors of xI - A is the totality of elementary divisors of all the $\mathrm{xI}-\mathrm{B}_{\mathrm{i}}$ taken together.

Proof We prove the theorem for the special case of $\mathrm{A}=\mathrm{B}_{1} \oplus \mathrm{~B}_{2}$. The general case follows by an obvious induction argument. Let $S=\left\{\mathrm{e}_{1}(\mathrm{x}), \ldots, \mathrm{e}_{\mathrm{m}}(\mathrm{x})\right\}$ denote the totality of elementary divisors of $\mathrm{xI}-\mathrm{B}_{1}$ and $\mathrm{xI}-\mathrm{B}_{2}$ taken together. Thus, the elements of S are powers of prime polynomials. Following the method discussed at the end of Theorem 8.8, we multiply together the highest powers of all the distinct primes that appear in $S$ to obtain a polynomial which we denote by $q_{n}(x)$. Deleting from $S$ those $e_{i}(x)$ that we just used, we now multiply together the highest powers of all the remaining distinct primes to obtain $\mathrm{q}_{\mathrm{n}-1}(\mathrm{x})$. We continue this procedure until all the elements of S are exhausted, thereby obtaining the polynomials $\mathrm{q}_{\mathrm{k}+1}(\mathrm{x}), \ldots, \mathrm{q}_{\mathrm{n}}(\mathrm{x})$. Note that our construction guarantees that $\mathrm{q}_{\mathrm{j}}(\mathrm{x}) \mid \mathrm{q}_{\mathrm{j}+1}(\mathrm{x})$ for $\mathrm{j}=\mathrm{k}+1, \ldots, \mathrm{n}-1$. Since $f_{n}(x)=q_{1}(x) \cdots q_{n}(x)$, it should also be clear that

$$
\sum_{i=k+1}^{n} \operatorname{deg} q_{j}(x)=n_{1}+n_{2}=n
$$

Denote the companion matrix $\mathrm{C}\left(\mathrm{q}_{\mathrm{j}}(\mathrm{x})\right)$ by simply $\mathrm{C}_{\mathrm{j}}$, and define the matrix

$$
\mathrm{Q}=\mathrm{C}_{\mathrm{k}+1} \oplus \cdots \oplus \mathrm{C}_{\mathrm{n}} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})
$$

Then

$$
\mathrm{xI}-\mathrm{Q}=\left(\mathrm{xI}-\mathrm{C}_{\mathrm{k}+1}\right) \oplus \cdots \oplus\left(\mathrm{xI}-\mathrm{C}_{\mathrm{n}}\right) .
$$

But according to Theorem $8.14, \mathrm{xI}-\mathrm{C}_{\mathrm{j}} \approx \operatorname{diag}\left(1, \ldots, 1, \mathrm{q}_{\mathrm{j}}(\mathrm{x})\right)$, and hence

$$
\mathrm{xI}-\mathrm{Q} \approx \operatorname{diag}\left(1, \ldots, 1, \mathrm{q}_{\mathrm{k}+1}(\mathrm{x})\right) \oplus \cdots \oplus \operatorname{diag}\left(1, \ldots, 1, \mathrm{q}_{\mathrm{n}}(\mathrm{x})\right)
$$

Then (since the Smith form is unique) the nonunit similarity invariants of Q are just the $\mathrm{q}_{\mathrm{j}}(\mathrm{x})$ (for $\mathrm{j}=\mathrm{k}+1, \ldots, \mathrm{n}$ ), and hence (by definition of elementary divisor) the elementary divisors of $\mathrm{xI}-\mathrm{Q}$ are exactly the polynomials in S . Then by Theorem 8.16 , Q is similar to the direct sum of the companion matrices of all the polynomials in S .

On the other hand, Theorem 8.16 also tells us that $B_{1}$ and $B_{2}$ are each similar to the direct sum of the companion matrices of the elementary divisors of $\mathrm{xI}-\mathrm{B}_{1}$ and $\mathrm{xI}-\mathrm{B}_{2}$ respectively. Therefore $\mathrm{B}_{1} \oplus \mathrm{~B}_{2}=\mathrm{A}$ is similar to the direct sum of the companion matrices of all the polynomials in S . We now see that A is similar to Q , and hence (by Theorem 8.9, Corollary 1) xI - A and xI - Q have the same elementary divisors. Since the elementary divisors of xI - Q are just the polynomials in S , and S was defined to be the totality of elementary divisors of xI $-\mathrm{B}_{1}$ and $\mathrm{xI}-\mathrm{B}_{2}$, the proof is complete.

The notion of "uniqueness" in Theorem 8.18 is an assertion that the Jordan form is "uniquely defined" or "well-defined." Suppose $A \in M_{n}(\mathbb{C})$ has Jordan form $\mathrm{H}_{1} \oplus \cdots \oplus \mathrm{H}_{\mathrm{p}}$ where each $\mathrm{H}_{\mathrm{i}}$ is a basic Jordan block, and suppose that $\mathrm{G}_{1} \oplus \cdots \oplus \mathrm{G}_{\mathrm{q}}$ is any other matrix similar to $A$ which is a direct sum of basic Jordan blocks. Then it follows from Theorem 8.20 that the $\mathrm{G}_{\mathrm{i}}$ must, except for order, be exactly the same as the $\mathrm{H}_{\mathrm{i}}$ (see Exercise 8.6.4). We state this in the following corollary to Theorem 8.20.

Corollary (Uniqueness of the Jordan form) Suppose $A \in M_{n}(\mathbb{C})$, and let both $\mathrm{G}=\mathrm{G}_{1} \oplus \cdots \oplus \mathrm{G}_{\mathrm{p}}$ and $\mathrm{H}=\mathrm{H}_{1} \oplus \cdots \oplus \mathrm{H}_{\mathrm{q}}$ be similar to $A$, where each $\mathrm{G}_{\mathrm{i}}$ and $\mathrm{H}_{\mathrm{i}}$ is a basic Jordan block. Then $\mathrm{p}=\mathrm{q}$ and, except for order, the $\mathrm{G}_{\mathrm{i}}$ are the same as the $\mathrm{H}_{\mathrm{i}}$.

We saw in Section 7.5 that if a vector space V is the direct sum of T invariant subspaces $\mathrm{W}_{\mathrm{i}}$ (where $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ ), then the matrix representation A of T is the direct sum of the matrix representations of $\mathrm{T}_{\mathrm{i}}=\mathrm{T} \mid \mathrm{W}_{\mathrm{i}}$ (Theorem 7.20). Another common way of describing this decomposition of A is the following. We say that a matrix is reducible over $\mathcal{F}$ if it is similar to a block diagonal matrix with more than one block. In other words, $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ is reducible if there exists a nonsingular matrix $\mathrm{S} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ and matrices $\mathrm{B} \in \mathrm{M}_{\mathrm{p}}(\mathcal{F})$ and $\mathrm{C} \in$ $\mathrm{M}_{\mathrm{q}}(\mathcal{F})$ with $\mathrm{p}+\mathrm{q}=\mathrm{n}$ such that $\mathrm{S}^{-1} \mathrm{AS}=\mathrm{B} \oplus \mathrm{C}$. If A is not reducible, then we say that A is irreducible. A fundamental result is the following.

Theorem 8.21 A matrix $A \in M_{n}(\mathcal{F})$ is irreducible over $\mathcal{F}$ if and only if $A$ is nonderogatory and the characteristic polynomial $\Delta_{\mathrm{A}}(\mathrm{x})$ is a power of a prime polynomial. Alternatively, A is irreducible if and only if xI - A has only one elementary divisor.

Proof If A is irreducible, then xI - A can have only one elementary divisor (which is then necessarily a prime to some power) because (by Theorem 8.16) A is similar to the direct sum of the companion matrices of all the elementary divisors of $\mathrm{xI}-\mathrm{A}$. But these elementary divisors are the factors of the similarity invariants $\mathrm{q}_{\mathrm{k}}(\mathrm{x})$ where $\mathrm{q}_{\mathrm{k}}(\mathrm{x}) \mid \mathrm{q}_{\mathrm{k}+1}(\mathrm{x})$, and therefore it follows that

$$
\mathrm{q}_{1}(\mathrm{x})=\cdots=\mathrm{q}_{\mathrm{n}-1}(\mathrm{x})=1
$$

Hence A is nonderogatory (corollary to Theorem 8.10).

Now assume that A is nonderogatory and that $\Delta_{\mathrm{A}}(\mathrm{x})$ is a power of a prime polynomial. From Theorem 8.10 and its corollary we know that $q_{1}(x)=\cdots=$ $\mathrm{q}_{\mathrm{n}-1}(\mathrm{x})=1$, and hence $\mathrm{q}_{\mathrm{n}}(\mathrm{x})=\mathrm{m}(\mathrm{x})=\Delta_{\mathrm{A}}(\mathrm{x})$ is now the only elementary divisor of $\mathrm{xI}-\mathrm{A}$. If A were reducible, then (in the above notation) it would be similar over F to a matrix of the form $\mathrm{B} \oplus \mathrm{C}=\mathrm{S}^{-1} \mathrm{AS}$, and by Corollary 1 of Theorem 8.9, it would then follow that $\mathrm{xI}-\mathrm{A}$ has the same elementary divisors as $\mathrm{xI}-(\mathrm{B} \oplus \mathrm{C})=(\mathrm{xI}-\mathrm{B}) \oplus(\mathrm{xI}-\mathrm{C})$. Note that by the corollary to Theorem 8.8, xI -A and $\mathrm{S}^{-1}(\mathrm{xI}-\mathrm{A}) \mathrm{S}=\mathrm{xI}-\mathrm{S}^{-1} \mathrm{AS}$ have the same elementary divisors. But xI - B and xI - C necessarily have at least one elementary divisor each (since their characteristic polynomials are nonzero), and (by Theorem 8.20) the elementary divisors of $\mathrm{xI}-\mathrm{S}^{-1} \mathrm{AS}$ are the totality of the elementary divisors of $\mathrm{xI}-\mathrm{B}$ plus those of $\mathrm{xI}-\mathrm{C}$. This contradicts the fact that xI - A has only one elementary divisor, and therefore A must be irreducible.

For example, we see from Theorem 8.17 that the hypercompanion matrix $\mathrm{H}\left((\mathrm{x}-\mathrm{a})^{\mathrm{k}}\right)$ is always irreducible. One consequence of this is that the Jordan canonical form of a matrix is the "simplest" in the sense that there is no similarity transformation that will further reduce any of the blocks on the diagonal. Similarly, since any elementary divisor is a power of a prime polynomial, we see from Theorem 8.14 that the companion matrix of an elementary divisor is always irreducible. Thus the rational canonical form can not be further reduced either. Note that the rational canonical form of a matrix $A \in M_{n}(\mathbb{C})$ will have the same "shape" as the Jordan form of A. In other words, both forms will consist of the same number of blocks of the same size on the diagonal.

In Sections 7.2 and 7.7 we proved several theorems that showed some of the relationships between eigenvalues and diagonalizability. Let us now relate what we have covered in this chapter to the question of diagonalizability. It is easiest to do this in the form of two simple theorems. The reader should note that the companion matrix of a linear polynomial $\mathrm{x}-\mathrm{a}_{0}$ is just the $1 \times 1$ matrix $\left(\mathrm{a}_{0}\right)$.

Theorem 8.22 A matrix $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ is similar over $\mathcal{F}$ to a diagonal matrix $\mathrm{D} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ if and only if all the elementary divisors of xI - A are linear.

Proof If the elementary divisors of xI - A are linear, then each of the corresponding companion matrices consists of a single scalar, and hence the rational canonical form of A will be diagonal (Theorem 8.16). Conversely, if A is similar to a diagonal matrix D , then $\mathrm{xI}-\mathrm{A}$ and $\mathrm{xI}-\mathrm{D}$ will have the same elementary divisors (Theorem 8.9, Corollary 1). Writing $D=D_{1} \oplus \cdots \oplus D_{n}$ where $D_{i}=\left(d_{i}\right)$ is just a $1 \times 1$ matrix, we see from Theorem 8.20 that the ele-
mentary divisors of xI-D are the linear polynomials $\left\{x-d_{1}, \ldots, x-d_{n}\right\}$ (since the elementary divisor of $\mathrm{xI}-\mathrm{D}_{\mathrm{i}}=\left(\mathrm{x}-\mathrm{d}_{\mathrm{i}}\right)$ is just $\left.\mathrm{x}-\mathrm{d}_{\mathrm{i}}\right)$.

Theorem 8.23 A matrix $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ is similar over $\mathcal{F}$ to a diagonal matrix $\mathrm{D} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ if and only if the minimal polynomial for A has distinct linear factors in $\mathcal{P}=\mathcal{F}[\mathrm{x}]$.

Proof Recall that the elementary divisors of a matrix in $\mathrm{M}_{\mathrm{n}}(\mathcal{P})$ are the powers of prime polynomials that factor the invariant factors $\mathrm{q}_{\mathrm{k}}(\mathrm{x})$, and furthermore, that $\mathrm{q}_{\mathrm{k}}(\mathrm{x}) \mid \mathrm{q}_{\mathrm{k}+1}(\mathrm{x})$. Then all the elementary divisors of such a matrix will be linear if and only if its invariant factor of highest degree has distinct linear factors in $\mathcal{P}$. But by Theorem 8.10, the minimal polynomial for $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{F})$ is just its similarity invariant of highest degree (i.e., the invariant factor of highest degree of $\mathrm{xI}-\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathcal{P})$ ). Then applying Theorem 8.22 , we see that A will be diagonalizable if and only if the minimal polynomial for A has distinct linear factors in $\mathcal{P}$.

While it is certainly not true that any $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathbb{C})$ is similar to a diagonal matrix, it is an interesting fact that A is similar to a matrix in which the offdiagonal entries are arbitrarily small. To see this, we first put A into its Jordan canonical form J. In other words, we have

$$
J=S^{-1} A S=\left(\begin{array}{ccccccc}
j_{11} & j_{12} & 0 & 0 & \cdots & 0 & 0 \\
0 & j_{22} & j_{23} & 0 & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & 0 & \cdots & j_{n-1 n-1} & j_{n-1 n} \\
0 & 0 & 0 & 0 & \cdots & 0 & j_{n n}
\end{array}\right) .
$$

If we now define the matrix $\mathrm{T}=\operatorname{diag}\left(1, \delta, \delta^{2}, \ldots, \delta^{\mathrm{n}-1}\right)$, then we leave it to the reader to show that

$$
\begin{aligned}
T^{-1} J T & =(S T)^{-1} A(S T) \\
& =\left(\begin{array}{ccccccc}
j_{11} & \delta j_{12} & 0 & 0 & \cdots & 0 & 0 \\
0 & j_{22} & \delta j_{23} & 0 & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & 0 & \cdots & j_{n-1 n-1} & \delta j_{n-1 n} \\
0 & 0 & 0 & 0 & \cdots & 0 & j_{n n}
\end{array}\right) .
\end{aligned}
$$

By choosing $\delta$ as small as desired, we obtain the form claimed.

## Exercises

1. If all the eigenvalues of $A \in M_{n}(\mathcal{F})$ lie in $\mathcal{F}$, show that the Jordan canonical form of A has the same "block structure" as its rational canonical form.
2. Prove Theorem 7.25 using the Jordan canonical form (Theorem 8.18).
3. Prove Theorem 7.26 using the Jordan canonical form.
4. Finish proving the corollary to Theorem 8.20.
5. State and prove a corollary to Theorem 8.16 that is the analogue of the corollary to Theorem 8.20.
6. (a) Suppose a matrix A has characteristic polynomial

$$
\Delta_{\mathrm{A}}(\mathrm{x})=(\mathrm{x}-2)^{4}(\mathrm{x}-3)^{3}
$$

and minimal polynomial

$$
\mathrm{m}(\mathrm{x})=(\mathrm{x}-2)^{2}(\mathrm{x}-3)^{2}
$$

What are the possible Jordan forms for A?
(b) Suppose A has characteristic polynomial $\Delta_{\mathrm{A}}(x)=(x-2)^{3}(x-5)^{2}$. What are the possible Jordan forms?
7. Find all possible Jordan forms for those matrices with characteristic and minimal polynomials given by:
(a) $\Delta(\mathrm{x})=(\mathrm{x}-2)^{4}(\mathrm{x}-3)^{2}$ and $\mathrm{m}(\mathrm{x})=(\mathrm{x}-2)^{2}(\mathrm{x}-3)^{2}$.
(b) $\Delta(\mathrm{x})=(\mathrm{x}-7)^{5}$ and $\mathrm{m}(\mathrm{x})=(\mathrm{x}-7)^{2}$.
(c) $\Delta(\mathrm{x})=(\mathrm{x}-2)^{7}$ and $\mathrm{m}(\mathrm{x})=(\mathrm{x}-2)^{3}$.
(d) $\Delta(\mathrm{x})=(\mathrm{x}-3)^{4}(\mathrm{x}-5)^{4}$ and $\mathrm{m}(\mathrm{x})=(\mathrm{x}-3)^{2}(\mathrm{x}-5)^{2}$.
8. Show that every complex matrix is similar to its transpose.
9. Is it true that all complex matrices $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathbb{C})$ with the property that $\mathrm{A}^{\mathrm{n}}=$ I but $\mathrm{A}^{\mathrm{k}} \neq \mathrm{I}$ for $\mathrm{k}<\mathrm{n}$ are similar? Explain.
10. (a) Is it true that an eigenvalue $\lambda$ of a matrix $A \in M_{n}(\mathbb{C})$ has multiplicity 1 if and only if $\lambda I-A$ has rank $n-1$ ?
(b) Suppose an eigenvalue $\lambda$ of $A \in M_{n}(\mathbb{C})$ is such that $r(\lambda I-A)=n-1$. Prove that either $\lambda$ has multiplicity 1 , or else $r(\lambda I-A)^{2}=n-2$.
11. Suppose $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathbb{C})$ is idempotent, i.e., $\mathrm{A}^{2}=\mathrm{A}$. What is the Jordan form of A?
12. Suppose $\mathrm{A} \in \mathrm{M}_{\mathrm{n}}(\mathbb{C})$ is such that $\mathrm{p}(\mathrm{A})=0$ where

$$
p(x)=(x-2)(x-3)(x-4)
$$

Prove or disprove the following statements:
(a) The minimal polynomial for A must be of degree 3 .
(b) A must be of size $\mathrm{n} \leq 3$.
(c) If $\mathrm{n}>3$, then the characteristic polynomial of A must have multiple roots.
(d) A is nonsingular.
(e) A must have 2, 3 and 4 as eigenvalues.
(f) If $\mathrm{n}=3$, then the minimal and characteristic polynomials of A must be the same.
(g) If $\mathrm{n}=3$ then, up to similarity, there are exactly 10 different choices for A .
13. Recall that $A \in \mathrm{M}_{\mathrm{n}}(\mathbb{C})$ is said to be nilpotent of index k if k is the smallest integer such that $\mathrm{A}^{\mathrm{k}}=0$.
(a) Describe the Jordan form of A.

Prove or disprove each of the following statements about $A \in M_{n}(\mathbb{C})$ :
(b) A is nilpotent if and only if every eigenvalue of A is zero.
(c) If $A$ is nilpotent, then $r(A)-r\left(A^{2}\right)$ is the number of elementary divisors of $A$.
(d) If $A$ is nilpotent, then $r(A)-r\left(A^{2}\right)$ is the number of $p x p$ Jordan blocks of A with $\mathrm{p}>1$.
(e) If A is nilpotent, then the nul(A) is the number of Jordan blocks of A (counting $1 \times 1$ blocks).
(f) If $A$ is nilpotent, then $\operatorname{nul}\left(\mathrm{A}^{\mathrm{k}+1}\right)-\operatorname{nul}\left(\mathrm{A}^{\mathrm{k}}\right)$ is the number of Jordan blocks of size greater than k .
14. Suppose $A \in M_{n}(\mathbb{C})$ has eigenvalue $\lambda$ of multiplicity m. Prove that the elementary divisors of A corresponding to $\lambda$ are all linear if and only if $r(\lambda I-A)=r\left((\lambda I-A)^{2}\right)$.
15. Prove or disprove the following statements about matrices $A, B \in M_{n}(\mathbb{C})$ :
(a) If either A or B is nonsingular, then AB and BA have the same minimal polynomial.
(b) If both $A$ and $B$ are singular and $A B \neq B A$, then $A B$ and $B A$ are not similar.
16. Suppose $A \in M_{n}(\mathbb{C})$, and let adj $A$ be as in Theorem 4.11. If $A$ is nonsingular, then $\left(\mathrm{SAS}^{-1}\right)^{-1}=\mathrm{SA}^{-1} \mathrm{~S}^{-1}$ implies that adj $\left(\mathrm{SAS}^{-1}\right)=\mathrm{S}(\operatorname{adj} \mathrm{A}) \mathrm{S}^{-1}$ by Theorem 4.11. By using "continuity" arguments, it is easy to show that this identity is true even if A is singular. Using this fact and the Jordan form, prove:
(a) If det $\mathrm{A}=0$ but $\operatorname{Tr}(\operatorname{adj} \mathrm{A}) \neq 0$, then 0 is an eigenvalue of A with multiplicity 1 .
(b) If $\operatorname{det} \mathrm{A}=0$ but $\operatorname{Tr}(\operatorname{adj} \mathrm{A}) \neq 0$, then $\mathrm{r}(\mathrm{A})=\mathrm{n}-1$.

### 8.7 CYCLIC SUBSPACES *

It is important to realize that the Jordan form can only be found in cases where the minimal polynomial is factorable into linear polynomials (for example, if the base field is algebraically closed). On the other hand, the rational canonical form is valid over non-algebraically closed fields. In order to properly present another way of looking at the rational canonical form, we first introduce cyclic subspaces. Again, we are seeking a criterion for deciding when two matrices are similar. The clue that we now follow up on was given earlier in Theorem 7.37.

Let $\mathrm{V} \neq 0$ be a finite-dimensional vector space over an arbitrary field F , and suppose $T \in L(V)$. We say that a nonzero $T$-invariant subspace $Z$ of $V$ is T-cyclic if there exists a nonzero $\mathrm{v} \in \mathrm{Z}$ and a positive integer $\mathrm{k} \geq 0$ such that Z is spanned by the set $\left\{\mathrm{v}, \mathrm{T}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{k}}(\mathrm{v})\right\}$. An equivalent way of defining T -cyclic subspaces is given in the following theorem.

Theorem 8.24 Let $V$ be finite-dimensional and suppose $T \in L(V)$. A subspace $\mathrm{Z} \subset \mathrm{V}$ is T -cyclic if and only if there exists a nonzero $\mathrm{v} \in \mathrm{Z}$ such that every vector in $Z$ can be expressed in the form $f(T)(v)$ for some $f(x) \in$ $\mathcal{F}[\mathrm{x}]$.

Proof If Z is T-cyclic, then by definition, any $\mathrm{u} \in \mathrm{Z}$ may be written in terms of the set $\left\{\mathrm{v}, \mathrm{T}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{k}}(\mathrm{v})\right\}$ as

$$
\mathrm{u}=\mathrm{a}_{0} \mathrm{v}+\mathrm{a}_{1} \mathrm{~T}(\mathrm{v})+\cdots+\mathrm{a}_{\mathrm{k}} \mathrm{~T}^{\mathrm{k}}(\mathrm{v})=\left(\mathrm{a}_{0}+\mathrm{a}_{1} \mathrm{~T}+\cdots+\mathrm{a}_{\mathrm{k}} \mathrm{~T}^{\mathrm{k}}\right)(\mathrm{v})=\mathrm{f}(\mathrm{~T})(\mathrm{v})
$$

where $f(x)=a_{0}+a_{1} x+\cdots+a_{k} x^{k} \in \mathcal{F}[x]$. On the other hand, if every $u \in Z$ is of the form $f(T)(v)$, then $Z$ must be spanned by the set $\left\{v, T(v), T^{2}(v), \ldots\right\}$. But Z is finite-dimensional (since it is a subset of the finite-dimensional space V ), and hence there must exist a positive integer k such that Z is spanned by the set $\left\{\mathrm{v}, \mathrm{T}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{k}}(\mathrm{v})\right\}$.

Generalizing these definitions slightly, let $\mathrm{v} \in \mathrm{V}$ be nonzero. Then the set of all vectors of the form $f(T)(v)$ where $f(x)$ varies over all polynomials in $\mathcal{F}[\mathrm{x}]$ is a T -invariant subspace called the T -cyclic subspace of V generated by v . We denote this subspace by $\mathrm{Z}(\mathrm{v}, \mathrm{T})$. We also denote the restriction of T to $Z(v, T)$ by $T_{v}=T \mid Z(v, T)$. That $Z(v, T)$ is a subspace is easily seen since for any $\mathrm{f}, \mathrm{g} \in \mathcal{F}[\mathrm{x}]$ and $\mathrm{a}, \mathrm{b} \in \mathcal{F}$ we have

$$
\operatorname{af}(\mathrm{T})(\mathrm{v})+\mathrm{bg}(\mathrm{~T})(\mathrm{v})=[\mathrm{af}(\mathrm{~T})+\mathrm{bg}(\mathrm{~T})](\mathrm{v})=\mathrm{h}(\mathrm{~T})(\mathrm{v}) \in \mathrm{Z}(\mathrm{v}, \mathrm{~T})
$$

where $h(x)=a f(x)+b g(x) \in \mathcal{F}[x]$ (by Theorem 7.2). It should be clear that $Z(v, T)$ is T-invariant since any element of $Z(v, T)$ is of the form $f(T)(v)$, and hence

$$
\mathrm{T}[\mathrm{f}(\mathrm{~T})(\mathrm{v})]=[\mathrm{Tf}(\mathrm{~T})](\mathrm{v})=\mathrm{g}(\mathrm{~T})(\mathrm{v})
$$

where $\mathrm{g}(\mathrm{x})=\mathrm{xf}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$. In addition, $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ is T-cyclic by Theorem 8.24. In the particular case that $\mathrm{Z}(\mathrm{v}, \mathrm{T})=\mathrm{V}$, then v is called a cyclic vector for T .

Let us briefly refer to Section 7.4 where we proved the existence of a unique monic polynomial $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ of least degree such that $\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{v})=0$. This polynomial was called the minimal polynomial of the vector v . The existence of $m_{v}(x)$ was based on the fact that $V$ was of dimension $n$, and hence for any $\mathrm{v} \in \mathrm{V}$, the $\mathrm{n}+1$ vectors $\left\{\mathrm{v}, \mathrm{T}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{n}}(\mathrm{v})\right\}$ must be linearly dependent. This showed that $\operatorname{deg} \mathrm{m}_{\mathrm{v}}(\mathrm{x}) \leq \mathrm{n}$. Since $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ generates the ideal $\mathrm{N}_{\mathrm{T}}(\mathrm{v})$, it follows that $m_{v}(x) \mid f(x)$ for any $f(x) \in N_{T}(v)$, i.e., where $f(x)$ is such that $f(T)(v)=0$. Let us now show how this approach can be reformulated in terms of T-cyclic subspaces.

Using Theorem 8.24 , we see that for any nonzero $\mathrm{v} \in \mathrm{V}$ we may define $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ as that finite-dimensional T -invariant subspace of V spanned by the linearly independent set $\left\{\mathrm{v}, \mathrm{T}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{d}-1}(\mathrm{v})\right\}$, where the integer $\mathrm{d} \geq 1$ is defined as the smallest integer such that the set $\left\{\mathrm{v}, \mathrm{T}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{d}}(\mathrm{v})\right\}$ is linearly
dependent. This means that $T^{d}(v)$ must be a linear combination of the vectors $v, T(v), \ldots, T^{d-1}(v)$, and hence is of the form

$$
\mathrm{T}^{\mathrm{d}}(\mathrm{v})=\mathrm{a}_{0} \mathrm{v}+\cdots+\mathrm{a}_{\mathrm{d}-1} \mathrm{~T}^{\mathrm{d}-1}(\mathrm{v})
$$

for some set of scalars $\left\{a_{i}\right\}$. Defining the polynomial

$$
m_{v}(x)=x^{d}-a_{d-1} x^{d-1}-\cdots-a_{0}
$$

we see that $\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{v})=0$, where $\operatorname{deg} \mathrm{m}_{\mathrm{v}}(\mathrm{x})=\mathrm{d}$. All that really remains is to show that if $\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$ is such that $\mathrm{f}(\mathrm{T})(\mathrm{v})=0$, then $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mid \mathrm{f}(\mathrm{x})$. This will prove that $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ is the polynomial of least degree with the property that $\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{v})=0$.

From the division algorithm, there exists $\mathrm{g}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$ such that

$$
f(x)=m_{v}(x) g(x)+r(x)
$$

where either $r(x)=0$ or $\operatorname{deg} r(x)<\operatorname{deg} m_{v}(x)$. Substituting $T$ and applying this to v we have (using $\left.\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{v})=0\right)$

$$
0=\mathrm{f}(\mathrm{~T})(\mathrm{v})=\mathrm{g}(\mathrm{~T}) \mathrm{m}_{\mathrm{v}}(\mathrm{~T})(\mathrm{v})+\mathrm{r}(\mathrm{~T})(\mathrm{v})=\mathrm{r}(\mathrm{~T})(\mathrm{v}) .
$$

But if $r(x) \neq 0$ with $\operatorname{deg} r(x)<\operatorname{deg} \mathrm{m}_{\mathrm{v}}(\mathrm{x})$, then (since $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ is T-invariant) $\mathrm{r}(\mathrm{T})(\mathrm{v})$ is a linear combination of elements in the set $\left\{\mathrm{v}, \mathrm{T}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{d}-1}(\mathrm{v})\right\}$, and hence the equation $r(T)(v)=0$ contradicts the assumed linear independence of this set. Therefore we must have $\mathrm{r}(\mathrm{x})=0$, and hence $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mid \mathrm{f}(\mathrm{x})$.

Lastly, we note that $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ is in fact the unique monic polynomial of least degree such that $m_{v}(T)(v)=0$. Indeed, if $m^{\prime}(x)$ is also of least degree such that $\mathrm{m}^{\prime}(\mathrm{T})(\mathrm{v})=0$, then the fact that $\operatorname{deg} \mathrm{m}^{\prime}(\mathrm{x})=\operatorname{deg} \mathrm{m}_{\mathrm{v}}(\mathrm{x})$ together with the result of the previous paragraph tells us that $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mid \mathrm{m}^{\prime}(\mathrm{x})$. Thus $\mathrm{m}^{\prime}(\mathrm{x})=\alpha \mathrm{m}_{\mathrm{v}}(\mathrm{x})$ for some $\alpha \in \mathrm{F}$, and choosing $\alpha=1$ shows that $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ is the unique monic polynomial of least degree such that $\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{v})=0$.

We summarize this discussion in the following theorem.
Theorem 8.25 Let $v \in V$ be nonzero and suppose $T \in L(V)$. Then there exists a unique monic polynomial $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ of least degree such that $\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{v})=$ 0 . Moreover, for any polynomial $\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$ with $\mathrm{f}(\mathrm{T})(\mathrm{v})=0$ we have $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mid \mathrm{f}(\mathrm{x})$.

Corollary If $\mathrm{m}(\mathrm{x})$ is the minimal polynomial for T on V , then $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mid \mathrm{m}(\mathrm{x})$ for every nonzero $\mathrm{v} \in \mathrm{V}$.

Proof By definition of minimal polynomial we know that $\mathrm{m}(\mathrm{T})=0$ on V , so that in particular we have $\mathrm{m}(\mathrm{T})(\mathrm{v})=0$. But now Theorem 8.25 shows that $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mid \mathrm{m}(\mathrm{x})$.

For ease of reference, we bring together Theorems 8.24 and 8.25 in the next basic result.

Theorem 8.26 Let $v \in V$ be nonzero, suppose $T \in L(V)$, and let

$$
m_{v}(x)=x^{d}-a_{d-1} x^{d-1}-\cdots-a_{0}
$$

be the minimal polynomial of $v$. Then $\left\{v, T(v), \ldots, T^{d-1}(v)\right\}$ is a basis for the $T$-cyclic subspace $\mathrm{Z}(\mathrm{v}, \mathrm{T})$, and hence $\operatorname{dim} \mathrm{Z}(\mathrm{v}, \mathrm{T})=\operatorname{deg} \mathrm{m}_{\mathrm{v}}(\mathrm{x})=\mathrm{d}$.

Proof From the way that $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ was constructed, the vector $\mathrm{T}^{\mathrm{d}}(\mathrm{v})$ is the first vector in the sequence $\left\{v, T(v), T^{2}(v), \ldots\right\}$ that is a linear combination of the preceding vectors. This means that the set $\mathrm{S}=\left\{\mathrm{v}, \mathrm{T}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{d}-1}(\mathrm{v})\right\}$ is linearly independent. We must now show that $f(T)(v)$ is a linear combination of the elements of $S$ for every $f(x) \in \mathcal{F}[x]$.

Since $\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{v})=0$ we have $T^{d}(v)=\sum_{i=0}^{d-1} a_{i} T^{i}(v)$. Therefore

$$
T^{d+1}(v)=\sum_{i=0}^{d-2} a_{i} T^{i+1}(v)+a_{d-1} T^{d}(v)=\sum_{i=0}^{d-2} a_{i} T^{i+1}(v)+a_{d-1} \sum_{i=0}^{d-2} a_{i} T^{i}(v)
$$

This shows that $T^{d+1}(v)$ is a linear combination of the elements of $S$. We can clearly continue this process for any $T^{k}(v)$ with $k \geq d$, and therefore $f(T)(v)$ is a linear combination of $v, T(v), \ldots, T^{d-1}(v)$ for every $f(x) \in \mathcal{F}[x]$. Thus $S$ is a basis for the T-cyclic subspace of V generated by v .

The following example will be used in the proof of the elementary divisor theorem given in the next section.

Example 8.12 Suppose that the minimal polynomial of $v$ is given by $m_{v}(x)=$ $\mathrm{p}(\mathrm{x})^{\mathrm{n}}$ where $\mathrm{p}(\mathrm{x})$ is a monic prime polynomial of degree d. Defining $\mathrm{W}=$ $\mathrm{Z}(\mathrm{v}, \mathrm{T})$, we will show that $\mathrm{p}(\mathrm{T})^{\mathrm{s}}(\mathrm{W})$ is a T -cyclic subspace generated by $\mathrm{p}(\mathrm{T})(\mathrm{v})$, and is of dimension $\mathrm{d}(\mathrm{n}-\mathrm{s})$ if $\mathrm{s}<\mathrm{n}$, and dimension 0 if $\mathrm{s} \geq \mathrm{n}$. It should be clear that $p(T)^{s}(W)$ is a T-cyclic subspace since every element of W is of the form $f(T)(v)$ for some $f(x) \in \mathcal{F}[x]$ and $W$ is T-invariant.

Since $p(x)$ is of degree $d$, we see that $\operatorname{deg} m_{v}(x)=\operatorname{deg} p(x)^{n}=\operatorname{dn}$ (see Theorem 6.2(b)). From Theorem 8.26, we then follows that W has the basis $\left\{\mathrm{v}, \mathrm{T}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{dn}-1}(\mathrm{v})\right\}$. This means that any $\mathrm{w} \in \mathrm{W}$ may be written as

$$
\mathrm{w}=\mathrm{a}_{0} \mathrm{v}+\mathrm{a}_{1} \mathrm{~T}(\mathrm{v})+\cdots+\mathrm{a}_{\mathrm{dn}-1} \mathrm{~T}^{\mathrm{dn}-1}(\mathrm{v})
$$

for some set of scalars $\mathrm{a}_{\mathrm{i}}$. Applying $\mathrm{p}(\mathrm{T})^{\mathrm{s}}$ to w we have

$$
\mathrm{p}(\mathrm{~T})^{\mathrm{s}}(\mathrm{w})=\mathrm{a}_{0} \mathrm{p}(\mathrm{~T})^{\mathrm{s}}(\mathrm{v})+\cdots+\mathrm{a}_{\mathrm{i}}\left[\mathrm{~T}^{\mathrm{i}} \mathrm{p}(\mathrm{~T})^{\mathrm{s}}\right](\mathrm{v})+\cdots+\mathrm{a}_{\mathrm{dn}-1}\left[\mathrm{~T}^{\mathrm{dn}-1} \mathrm{p}(\mathrm{~T})^{\mathrm{s}}\right](\mathrm{v}) .
$$

But $m_{v}(T)(v)=p(T)^{n}(v)=0$ where $\operatorname{deg} m_{v}(x)=d n$, and $\operatorname{deg} p(x)^{s}=d s$. Therefore, if $s \geq n$ we automatically have $p(T)^{s}(w)=0$ so that $p(T)^{s}(W)$ is of dimension 0 . If $\mathrm{s}<\mathrm{n}$, then the maximum value of i in the expression for $\mathrm{p}(\mathrm{T})^{\mathrm{s}}(\mathrm{w})$ comes from the requirement that $\mathrm{i}+\mathrm{ds}<\mathrm{dn}$ which is equivalent to $\mathrm{i}<\mathrm{d}(\mathrm{n}-\mathrm{s})$. This leaves us with

$$
\mathrm{p}(\mathrm{~T})^{\mathrm{s}}(\mathrm{w})=\mathrm{a}_{0}\left[\mathrm{p}(\mathrm{~T})^{\mathrm{s}}(\mathrm{v})\right]+\cdots+\mathrm{a}_{\mathrm{d}(\mathrm{n}-\mathrm{s})-1} \mathrm{~T}^{\mathrm{d}(\mathrm{n}-\mathrm{s})-1}\left[\mathrm{p}(\mathrm{~T})^{\mathrm{s}}(\mathrm{v})\right]
$$

and we now see that any element in $p(T)^{s}(W)$ is a linear combination of the terms $\mathrm{a}_{\mathrm{i}} \mathrm{T}^{\mathrm{i}}\left[\mathrm{p}(\mathrm{T})^{\mathrm{s}}(\mathrm{v})\right]$ for $\mathrm{i}=0, \ldots, \mathrm{~d}(\mathrm{n}-\mathrm{s})-1$. Therefore if $\mathrm{s}<\mathrm{n}$, this shows that $p(T)^{s}(W)$ is a $T$-cyclic subspace of dimension $d(n-s)$ generated by $\mathrm{p}(\mathrm{T})^{\mathrm{s}}(\mathrm{v})$. //

In Section 7.4 we showed that the minimal polynomial for T was the unique monic generator of the ideal $\mathrm{N}_{\mathrm{T}}=\cap_{\mathrm{v} \in \mathrm{V}} \mathrm{N}_{\mathrm{T}}(\mathrm{v})$. If we restrict ourselves to the subspace $Z(v, T)$ of $V$ then, as we now show, it is true that the minimal polynomial $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ of v is actually the minimal polynomial for $\mathrm{T}_{\mathrm{v}}=\mathrm{T} \mid \mathrm{Z}(\mathrm{v}, \mathrm{T})$.

Theorem 8.27 Let $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ be the T -cyclic subspace of V generated by v . Then $m_{v}(x)$ is equal to the minimal polynomial for $T_{v}=T \mid Z(v, T)$.

Proof Since $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ is spanned by $\left\{\mathrm{v}, \mathrm{T}(\mathrm{v}), \mathrm{T}^{2}(\mathrm{v}), \ldots, \mathrm{T}^{\mathrm{d}-1}(\mathrm{v})\right\}$, the fact that $\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{v})=0$ means that $\mathrm{m}_{\mathrm{v}}(\mathrm{T})=0$ on $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ (by Theorem 7.2). If $\mathrm{p}(\mathrm{x})$ is the minimal polynomial for $T_{v}$, then Theorem 7.4 tells us that $p(x) \mid m_{v}(x)$. On the other hand, from Theorem $7.17(a)$, we see that $p(T)(v)=p\left(T_{v}\right)(v)=0$ since $p(x)$ is the minimal polynomial for $T_{v}$. Therefore, Theorem 8.25 shows us that $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mid \mathrm{p}(\mathrm{x})$. Since both $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ and $\mathrm{p}(\mathrm{x})$ are monic, this implies that $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ $=\mathrm{p}(\mathrm{x})$.

Theorem 8.27 also gives us another proof of the corollary to Theorem 8.25. Thus, since $\mathrm{m}_{\mathrm{v}}(\mathrm{x})=\mathrm{p}(\mathrm{x})$ (i.e., the minimal polynomial for $\mathrm{T}_{\mathrm{v}}$ ), Theorem 7.17(b) shows that $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mid \mathrm{m}(\mathrm{x})$. Moreover, we have the next result that ties together these concepts with the structure of quotient spaces.

Theorem 8.28 Suppose $T \in L(V)$, let W be a T -invariant subspace of V and let $\bar{T} \in A(\overline{\mathrm{~V}})$ be the induced linear operator on $\overline{\mathrm{V}}=\mathrm{V} / \mathrm{W}$ (see Theorem 7.35). Then the minimal polynomial $\overline{\mathrm{m}}_{\mathrm{v}}(\mathrm{x})$ for $\overline{\mathrm{v}} \in \mathrm{V} / \mathrm{W}$ divides the minimal polynomial $\mathrm{m}(\mathrm{x})$ for T .

Proof From the corollary to Theorem 8.25 we have $\overline{\mathrm{m}}_{\mathrm{v}}(\mathrm{x}) \mid \overline{\mathrm{m}}(\mathrm{x})$ where $\overline{\mathrm{m}}(\mathrm{x})$ is the minimal polynomial for $\overline{\mathrm{T}}$. But $\overline{\mathrm{m}}(\mathrm{x}) \mid \mathrm{m}(\mathrm{x})$ by Theorem 7.35 .

Corollary Using the same notation as in Theorems 8.25 and 8.28 , if the minimal polynomial for $T$ is of the form $p(x)^{n}$ where $p(x)$ is a monic prime polynomial, then for any $\mathrm{v} \in \mathrm{V}$ we have $\mathrm{m}_{\mathrm{v}}(\mathrm{x})=\mathrm{p}(\mathrm{x})^{\mathrm{n}_{1}}$ and $\overline{\mathrm{m}}_{\mathrm{v}}(\mathrm{x})=\mathrm{p}(\mathrm{x})^{\mathrm{n}_{2}}$ for some $\mathrm{n}_{1}, \mathrm{n}_{2} \leq \mathrm{n}$.

Proof From the above results we know that $\mathrm{m}_{\mathrm{v}}(\mathrm{x}) \mid \mathrm{p}(\mathrm{x})^{\mathrm{n}}$ and $\overline{\mathrm{m}}_{\mathrm{v}}(\mathrm{x}) \mid \mathrm{p}(\mathrm{x})^{\mathrm{n}}$. The corollary then follows from this along with the unique factorization theorem (Theorem 6.6) and the fact that $\mathrm{p}(\mathrm{x})$ is monic and prime.

In the discussion that followed Theorem 7.16 we showed that the (unique) minimal polynomial $\mathrm{m}(\mathrm{x})$ for $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is also the minimal polynomial $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$ for some $\mathrm{v} \in \mathrm{V}$. (This is because each basis vector $\mathrm{v}_{\mathrm{i}}$ for V has its own minimal polynomial $\mathrm{m}_{\mathrm{i}}(\mathrm{x})$, and the least common multiple of the $\mathrm{m}_{\mathrm{i}}(\mathrm{x})$ is both the minimal polynomial for some vector $\mathrm{v} \in \mathrm{V}$ and the minimal polynomial for T.) Now suppose that v also happens to be a cyclic vector for T , i.e., $\mathrm{Z}(\mathrm{v}, \mathrm{T})=$ V. By Theorem 8.26 we know that

$$
\operatorname{dim} V=\operatorname{dim} Z(v, T)=\operatorname{deg} m_{v}(x)=\operatorname{deg} m(x)
$$

However, the characteristic polynomial $\Delta_{\mathrm{T}}(\mathrm{x})$ for T must always be of degree equal to $\operatorname{dim} \mathrm{V}$, and hence the corollary to Theorem 7.42 (or Theorems 7.11 and 7.12) shows us that $\mathrm{m}(\mathrm{x})=\Delta_{\mathrm{T}}(\mathrm{x})$.

On the other hand, suppose that the characteristic polynomial $\Delta_{T}(x)$ of $T$ is equal to the minimal polynomial $m(x)$ for $T$. Then if $v \in V$ is such that $m_{v}(x)$ $=\mathrm{m}(\mathrm{x})$ we have

$$
\operatorname{dim} V=\operatorname{deg} \Delta_{\mathrm{T}}(\mathrm{x})=\operatorname{deg} \mathrm{m}(\mathrm{x})=\operatorname{deg} \mathrm{m}_{\mathrm{v}}(\mathrm{x})
$$

Applying Theorem 8.26 again, we see that $\operatorname{dim} \mathrm{Z}(\mathrm{v}, \mathrm{T})=\operatorname{deg} \mathrm{m}_{\mathrm{v}}(\mathrm{x})=\operatorname{dim} \mathrm{V}$, and hence v is a cyclic vector for T . We have thus proven the following useful result.

Theorem 8.29 Let $V$ be finite-dimensional and suppose $T \in L(V)$. Then $T$ has a cyclic vector if and only if the characteristic and minimal polynomials for T are identical. Thus the matrix representation of T is nonderogatory.

In view of Theorem 8.12, our next result should have been expected.
Theorem 8.30 Let $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ be a T -cyclic subspace of V , let $\mathrm{T}_{\mathrm{v}}=\mathrm{T} \mid \mathrm{Z}(\mathrm{v}, \mathrm{T})$ and suppose that the minimal polynomial for v is given by

$$
m_{v}(x)=x^{d}-a_{d-1} x^{d-1}-\cdots-a_{0} .
$$

Then the matrix representation of $T_{v}$ relative to the basis $v, T(v), \ldots, T^{d-1}(v)$ for $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ is the companion matrix

$$
C\left(m_{v}(x)\right)=\left(\begin{array}{cccccc}
0 & 0 & 0 & \cdots & 0 & a_{0} \\
1 & 0 & 0 & \cdots & 0 & a_{1} \\
0 & 1 & 0 & \cdots & 0 & a_{2} \\
\vdots & \vdots & \vdots & & \vdots & \vdots \\
0 & 0 & 0 & \cdots & 0 & a_{d-2} \\
0 & 0 & 0 & \cdots & 1 & a_{d-1}
\end{array}\right)
$$

Proof Simply look at $\mathrm{T}_{\mathrm{v}}$ applied to each of the basis vectors of $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ and note that $\mathrm{m}_{\mathrm{v}}(\mathrm{T})(\mathrm{v})=0$ implies that $\mathrm{T}^{\mathrm{d}}(\mathrm{v})=\mathrm{a}_{0} \mathrm{v}+\cdots+\mathrm{a}_{\mathrm{d}-1} \mathrm{~T}^{\mathrm{d}-1}(\mathrm{v})$. This yields

$$
\begin{aligned}
T_{v}(v) & =0 v+T(v) \\
T_{v}(T(v)) & =0 v+0 T(v)+T^{2}(v) \\
& \vdots \\
T_{v}\left(T^{d-2}(v)\right) & =0 v+\cdots+T^{d-1}(v) \\
T_{v}\left(T^{d-1}(v)\right) & =T^{d}(v)=a_{0} v+\cdots+a_{d-1} T^{d-1}(v)
\end{aligned}
$$

As usual, the ith column of the matrix representation of $\mathrm{T}_{\mathrm{v}}$ is just the image under $\mathrm{T}_{\mathrm{v}}$ of the i th basis vector of $\mathrm{Z}(\mathrm{v}, \mathrm{T})$ (see Theorem 5.11).

## Exercises

1. If $T \in L(V)$ and $v \in V$, prove that $Z(v, T)$ is the intersection of all $T$ invariant subspaces containing v .
2. Suppose $T \in L(V)$, and let $u, v \in V$ have relatively prime minimal polynomials $\mathrm{m}_{\mathrm{u}}(\mathrm{x})$ and $\mathrm{m}_{\mathrm{v}}(\mathrm{x})$. Show that $\mathrm{m}_{\mathrm{u}}(\mathrm{x}) \mathrm{m}_{\mathrm{v}}(\mathrm{x})$ is the minimal polynomial of $u+v$.
3. Prove that $Z(u, T)=Z(v, T)$ if and only if $g(T)(u)=v$ where $g(x)$ is relatively prime to $\mathrm{m}_{\mathrm{u}}(\mathrm{x})$.

### 8.8 THE ELEMENTARY DIVISOR THEREOM *

The reader should recall from Section 7.5 that if the matrix representation A of an operator $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ is the direct sum of smaller matrices (in the appropriate basis for V ), then V is just the direct sum of T -invariant subspaces (see Theorem 7.20). If we translate Theorem 8.16 (the rational canonical form) into the corresponding result on the underlying space V , then we obtain the elementary divisor theorem.

Theorem 8.31 (Elementary Divisor Theorem) Let $\mathrm{V} \neq\{0\}$ be finitedimensional over an arbitrary field $\mathcal{F}$, and suppose $\mathrm{T} \in \mathrm{L}(\mathrm{V})$. Then there exist vectors $\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{r}}$ in V such that:
(a) $\mathrm{V}=\mathrm{Z}\left(\mathrm{v}_{1}, \mathrm{~T}\right) \oplus \cdots \oplus \mathrm{Z}\left(\mathrm{v}_{\mathrm{r}}, \mathrm{T}\right)$.
(b) Each $\mathrm{v}_{\mathrm{i}}$ has minimal polynomial $\mathrm{p}_{\mathrm{i}}(\mathrm{x})^{\mathrm{n}_{\mathrm{i}}}$ where $\mathrm{p}_{\mathrm{i}}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$ is a monic prime.
(c) The number r of terms in the decomposition of V is uniquely determined, as is the set of minimal polynomials $p_{i}(x)^{n_{i}}$.

Proof This is easy to prove from Theorem 8.16 (the rational canonical form) and what we know about companion matrices and cyclic subspaces (particularly Theorem 8.30). The details are left to the reader.

From Theorem 8.26 we see that $\operatorname{dim} Z\left(v_{i}, T\right)=\operatorname{deg} p_{i}(x)^{n_{i}}$, and hence from the corollary to Theorem 2.15 we have

$$
\operatorname{dim} V=\sum_{i=1}^{r} \operatorname{deg} p_{i}(x)^{n_{i}}
$$

The polynomials $p_{i}(x)^{n_{i}}$ defined in Theorem 8.31 are just the elementary divisors of $\mathrm{x} 1-\mathrm{T}$. For example, suppose that $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ and $\mathrm{x} 1-\mathrm{T}$ has the ele-
mentary divisors $x+1,(x-1)^{2}, x+1, x^{2}+1$ over the field $R$. This means that V is a vector space over $\mathbb{R}$ with

$$
\mathrm{V}=\mathrm{Z}\left(\mathrm{v}_{1}, \mathrm{~T}\right) \oplus \mathrm{Z}\left(\mathrm{v}_{2}, \mathrm{~T}\right) \oplus \mathrm{Z}\left(\mathrm{v}_{3}, \mathrm{~T}\right) \oplus \mathrm{Z}\left(\mathrm{v}_{4}, \mathrm{~T}\right)
$$

and the minimal polynomials of $\mathrm{v}_{1}, \mathrm{v}_{2}, \mathrm{v}_{3}, \mathrm{v}_{4}$ are $\mathrm{x}+1,(\mathrm{x}-1)^{2}, \mathrm{x}+1, \mathrm{x}^{2}+1$ respectively. Furthermore, $T=T_{1} \oplus T_{2} \oplus T_{3} \oplus T_{4}$ where $T_{i}=T \mid Z\left(v_{i}, T\right)$ and the minimal polynomial for $T_{i}$ is just the corresponding minimal polynomial of $v_{i}$ (Theorem 8.27). Note that if the field were $\mathbb{C}$ instead of $\mathbb{R}$, then $x^{2}+1$ would not be prime, and hence could not be an elementary divisor of $\mathrm{x} 1-\mathrm{T}$.

It is important to realize that Theorem 8.31 only claims the uniqueness of the set of elementary divisors of $\mathrm{x} 1-\mathrm{T}$. Thus the vectors $\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{r}}$ and corresponding subspaces $\mathrm{Z}\left(\mathrm{v}_{1}, \mathrm{~T}\right), \ldots, \mathrm{Z}\left(\mathrm{v}_{\mathrm{r}}, \mathrm{T}\right)$ are themselves not uniquely determined by T . In addition, we have seen that the elementary divisors are unique only up to a rearrangement.

It is also possible to prove Theorem 8.31 without using Theorem 8.16 or any of the formalism developed in Sections $8.2-8.7$. We now present this alternative approach as a difficult but instructive application of quotient spaces, noting that it is not needed for anything else in this book. We begin with a special case that takes care of most of the proof. Afterwards, we will show how Theorem 8.31 follows from Theorem 8.32. It should also be pointed out that Theorem 8.32 also follows from the rational canonical form (Theorem 8.16).

Theorem 8.32 Let $T \in L(V)$ have minimal polynomial $p(x)^{n}$ where $p(x)$ is a monic prime polynomial. Then there exist vectors $\mathrm{v}_{1}, \ldots, \mathrm{v}_{\mathrm{r}} \in \mathrm{V}$ such that

$$
\mathrm{V}=\mathrm{Z}\left(\mathrm{v}_{1}, \mathrm{~T}\right) \oplus \cdots \oplus \mathrm{Z}\left(\mathrm{v}_{\mathrm{r}}, \mathrm{~T}\right)
$$

In addition, each $\mathrm{v}_{\mathrm{i}}$ has corresponding minimal polynomial (i.e., order) given by $p(x)^{n_{i}}$ where $n=n_{1} \geq n_{2} \geq \cdots \geq n_{r}$. Furthermore, any other decomposition of V into the direct sum of T-cyclic subspaces has the same number r of components and the same set of minimal polynomials (i.e., orders).

Proof Throughout this (quite long) proof, we will use the term "order" rather than "minimal polynomial" for the sake of clarity. Furthermore, we will refer to the $\mathbf{T}$-order of a vector rather than simply the order when there is a possible ambiguity with respect to the operator being referred to.

We proceed by induction on the dimension of V . First, if $\operatorname{dim} \mathrm{V}=1$, then $\mathrm{T}(\mathrm{V})=\mathrm{V}$ and hence $\mathrm{f}(\mathrm{T})(\mathrm{V})=\mathrm{V}$ for any $\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$. Therefore V is T -cyclic
and the theorem holds in this case. Now assume $\operatorname{dim} \mathrm{V}>1$, and suppose that the theorem holds for all vector spaces of dimension less than that of V .

Since $p(x)^{n}$ is the minimal polynomial for $T$, we know that $p(T)^{n}(v)=0$ for all $v \in V$. In particular, there must exist a $v_{1} \in V$ such that $p(T)^{n}\left(v_{1}\right)=0$ but $\mathrm{p}(\mathrm{T})^{\mathrm{n}-1}\left(\mathrm{v}_{1}\right) \neq 0$ (or else $\mathrm{p}(\mathrm{x})^{\mathrm{n}-1}$ would be the minimal polynomial for T ). This means that $p(x)^{n}$ must be the T-order of $v_{1}$ (since the minimal polynomial of $\mathrm{v}_{1}$ is unique and monic). Now let $\mathrm{Z}_{1}=\mathrm{Z}\left(\mathrm{v}_{1}, \mathrm{~T}\right)$ be the T -invariant T-cyclic subspace of V generated by $\mathrm{v}_{1}$. We also define $\overline{\mathrm{V}}=\mathrm{V} / \mathrm{Z}_{1}$ along with the induced operator $\overline{\mathrm{T}} \in \mathrm{A}(\overline{\mathrm{V}})$. Then by Theorem 7.35 we know that the minimal polynomial for $\overline{\mathrm{T}}$ divides the minimal polynomial $\mathrm{p}(\mathrm{x})^{\mathrm{n}}$ for T , and hence the minimal polynomial for $\overline{\mathrm{T}}$ is $\mathrm{p}(\mathrm{x})^{\mathrm{n}_{2}}$ where $\mathrm{n}_{2} \leq \mathrm{n}$. This means that $\overline{\mathrm{V}}$ and $\overline{\mathrm{T}}$ satisfy the hypotheses of the theorem, and hence by our induction hypothesis (since $\operatorname{dim} \overline{\mathrm{V}}<\operatorname{dim} \mathrm{V}$ ), $\overline{\mathrm{V}}$ must be the direct sum of T-cyclic subspaces. We thus write

$$
\overline{\mathrm{V}}=\mathrm{Z}\left(\overline{\mathrm{v}}_{2}, \overline{\mathrm{~T}}\right) \oplus \cdots \oplus \mathrm{Z}\left(\overline{\mathrm{v}}_{\mathrm{r}}, \overline{\mathrm{~T}}\right)
$$

where each $\bar{v}_{i}$ has corresponding $\overline{\mathrm{T}}$-order $\mathrm{p}(\mathrm{x})^{\mathrm{n}_{\mathrm{i}}}$ with $\mathrm{n} \geq \mathrm{n}_{2} \geq \cdots \geq \mathrm{n}_{\mathrm{r}}$. It is important to remember that each of these $\bar{v}_{i}$ is a coset of $Z_{1}$ in $V$, and thus may be written as $\bar{v}_{i}=u_{i}+Z_{1}$ for some $u_{i} \in V$. This means that every element of $\bar{v}_{i}$ is of the form $u_{i}+z_{i}$ for some $z_{i} \in Z_{1}$.

We now claim that there exists a vector $\mathrm{v}_{2}$ in the coset $\overline{\mathrm{v}}_{2}$ such that the T order of $\mathrm{v}_{2}$ is just the $\overline{\mathrm{T}}$-order $\mathrm{p}(\mathrm{x})^{\mathrm{n}_{2}}$ of $\overline{\mathrm{v}}_{2}$. To see this, let $\mathrm{w} \in \overline{\mathrm{v}}_{2}$ be arbitrary so that we may write $\mathrm{w}=\mathrm{u}_{2}+\mathrm{z}_{2}$ for some $\mathrm{u}_{2} \in \mathrm{~V}$ and $\mathrm{z}_{2} \in \mathrm{Z}_{1} \subset \mathrm{~V}$. Since $\mathrm{p}(\overline{\mathrm{T}})^{\mathrm{n}_{2}}\left(\overline{\mathrm{v}}_{2}\right)=\overline{0}=\mathrm{Z}_{1}$, we have (see Theorem 7.35)

$$
\mathrm{Z}_{1}=\mathrm{p}(\overline{\mathrm{~T}})^{\mathrm{n}_{2}}\left(\overline{\mathrm{v}}_{2}\right)=\mathrm{p}(\overline{\mathrm{~T}})^{\mathrm{n}_{2}}\left(\mathrm{u}_{2}+\mathrm{Z}_{1}\right)=\mathrm{p}(\mathrm{~T})^{\mathrm{n}_{2}}\left(\mathrm{u}_{2}\right)+\mathrm{Z}_{1}
$$

and hence $\mathrm{p}(\mathrm{T})^{\mathrm{n}_{2}}\left(\mathrm{u}_{2}\right) \in \mathrm{Z}_{1}$. Using the fact that $\mathrm{Z}_{1}$ is T -invariant, we see that

$$
\mathrm{p}(\mathrm{~T})^{\mathrm{n}_{2}}(\mathrm{w})=\mathrm{p}(\mathrm{~T})^{\mathrm{n}_{2}}\left(\mathrm{u}_{2}\right)+\mathrm{p}(\mathrm{~T})^{\mathrm{n}_{2}}\left(\mathrm{Z}_{2}\right) \in \mathrm{Z}_{1} .
$$

Using the definition of $Z_{1}$ as the $T$-cyclic subspace generated by $\mathrm{v}_{1}$, this last result implies that there exists a polynomial $\mathrm{f}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$ such that

$$
\begin{equation*}
p(T)^{n_{2}}(w)=f(T)\left(v_{1}\right) . \tag{1}
\end{equation*}
$$

But $\mathrm{p}(\mathrm{x})^{\mathrm{n}}$ is the minimal polynomial for T , and hence (1) implies that

$$
0=p(T)^{\mathrm{n}}(\mathrm{w})=\mathrm{p}(\mathrm{~T})^{\mathrm{n}-\mathrm{n}_{2}} \mathrm{p}(\mathrm{~T})^{\mathrm{n}_{2}}(\mathrm{w})=\mathrm{p}(\mathrm{~T})^{\mathrm{n}-\mathrm{n}_{2}} \mathrm{f}(\mathrm{~T})\left(\mathrm{v}_{1}\right) .
$$

Since we showed that $p(x)^{n}$ is also the T-order of $v_{1}$, Theorem 8.25 tells us that $\mathrm{p}(\mathrm{x})^{\mathrm{n}}$ divides $\mathrm{p}(\mathrm{x})^{\mathrm{n}}{ }^{-\mathrm{n}_{2}} \mathrm{f}(\mathrm{x})$, and hence there exists a polynomial $\mathrm{g}(\mathrm{x}) \in$ $\mathcal{F}[\mathrm{x}]$ such that $\mathrm{p}(\mathrm{x})^{\mathrm{n}-\mathrm{n}_{2}} \mathrm{f}(\mathrm{x})=\mathrm{p}(\mathrm{x})^{\mathrm{n}} \mathrm{g}(\mathrm{x})$. Rearranging, this may be written as $p(x)^{n-n_{2}}\left[f(x)-p(x)^{n_{2}} g(x)\right]=0$. Since $\mathcal{F}[x]$ is an integral domain, this implies (see Theorem 6.2, Corollary 2)

$$
\begin{equation*}
f(x)=p(x)^{n_{2}} g(x) . \tag{2}
\end{equation*}
$$

We now define

$$
\begin{equation*}
v_{2}=w-g(T)\left(v_{1}\right) . \tag{3}
\end{equation*}
$$

By definition of $Z_{1}$ we see that $w-v_{2}=g(T)\left(v_{1}\right) \in Z_{1}$, and therefore (see Theorem 7.30)

$$
\mathrm{v}_{2} \in \mathrm{w}+\mathrm{Z}_{1}=\mathrm{u}_{2}+\mathrm{z}_{2}+\mathrm{Z}_{1}=\mathrm{u}_{2}+\mathrm{Z}_{1}=\overline{\mathrm{v}}_{2}
$$

Since $\bar{v}_{2}=u_{2}+Z_{1}$ and $\mathrm{v}_{2} \in \overline{\mathrm{v}}_{2}$, it follows that $\mathrm{v}_{2}=\mathrm{u}_{2}+\mathrm{z}$ for some $\mathrm{z} \in \mathrm{Z}_{1}$. Now suppose that $h(x)$ is any polynomial such that $h(T)\left(v_{2}\right)=0$. Then

$$
0=\mathrm{h}(\mathrm{~T})\left(\mathrm{v}_{2}\right)=\mathrm{h}(\mathrm{~T})\left(\mathrm{u}_{2}+\mathrm{z}\right)=\mathrm{h}(\mathrm{~T})\left(\mathrm{u}_{2}\right)+\mathrm{h}(\mathrm{~T})(\mathrm{z})
$$

so that $\mathrm{h}(\mathrm{T})\left(\mathrm{u}_{2}\right)=-\mathrm{h}(\mathrm{T})(\mathrm{z}) \in \mathrm{Z}_{1}$ (since $\mathrm{Z}_{1}$ is T-invariant). We then have

$$
\mathrm{h}(\overline{\mathrm{~T}})\left(\overline{\mathrm{v}}_{2}\right)=\mathrm{h}(\overline{\mathrm{~T}})\left(\mathrm{u}_{2}+\mathrm{Z}_{1}\right)=\mathrm{h}(\mathrm{~T})\left(\mathrm{u}_{2}\right)+\mathrm{Z}_{1}=\mathrm{Z}_{1}=\overline{0} .
$$

According to Theorem 8.25, this then means that the $\overline{\mathrm{T}}$-order of $\overline{\mathrm{v}}_{2}$ divides $h(x)$. In particular, choosing $h(x)$ to be the T-order of $v_{2}$, we see that the Torder of $\mathrm{v}_{2}$ is some multiple of the $\overline{\mathrm{T}}$-order of $\overline{\mathrm{v}}_{2}$. In other words, the T-order of $\mathrm{v}_{2}$ must equal $\mathrm{p}(\mathrm{x})^{\mathrm{n}_{2}} \mathrm{q}(\mathrm{x})$ for some polynomial $\mathrm{q}(\mathrm{x}) \in \mathcal{F}[\mathrm{x}]$. However, from (3), (1) and (2) we have

$$
\begin{aligned}
p(T)^{n_{2}}\left(v_{2}\right) & =p(T)^{n_{2}}\left[w-g(T)\left(v_{1}\right)\right] \\
& =p(T)^{n_{2}}(w)-p(T)^{n_{2}} g(T)\left(v_{1}\right) \\
& =f(T)\left(v_{1}\right)-f(T)\left(v_{1}\right) \\
& =0 .
\end{aligned}
$$

This shows that in fact the T-order of $v_{2}$ is equal to $p(x)^{n_{2}}$ as claimed.
In an exactly analogous manner, we see that there exist vectors $v_{3}, \ldots, v_{r}$ in $V$ with $v_{i} \in \bar{v}_{i}$ and such that the T-order of $v_{i}$ is equal to the $\bar{T}$-order $p(x)^{n_{i}}$ of $\bar{v}_{i}$. For each $i=1, \ldots, r$ we then define the $T$-cyclic subspaces $Z_{i}=Z\left(v_{i}, T\right)$
where $\mathrm{Z}_{1}$ was defined near the beginning of the proof. We must show that $\mathrm{V}=$ $\mathrm{Z}_{1} \oplus \cdots \oplus \mathrm{Z}_{\mathrm{r}}$.

Let $\operatorname{deg} p(x)=d$ so that $\operatorname{deg} p(x)^{n_{i}}=d_{n_{i}}$ (see Theorem 6.2(b)). Since $p(x)^{n_{i}}$ is the T-order of $\mathrm{v}_{\mathrm{i}}$, Theorem 8.26 shows that $\mathrm{Z}\left(\mathrm{v}_{\mathrm{i}}, \mathrm{T}\right)$ has basis

$$
\left\{\mathrm{v}_{\mathrm{i}}, \mathrm{~T}\left(\mathrm{v}_{\mathrm{i}}\right), \ldots, \mathrm{T}^{\mathrm{dn}_{\mathrm{i}}-1}\left(\mathrm{v}_{\mathrm{i}}\right)\right\}
$$

Similarly, $p(x)^{n_{i}}$ is also the $\overline{\mathrm{T}}$-order of $\overline{\mathrm{v}}_{\mathrm{i}}$ for $\mathrm{i}=2, \ldots, r$ and hence $\mathrm{Z}\left(\overline{\mathrm{v}}_{\mathrm{i}}, \overline{\mathrm{T}}\right)$ has the basis

$$
\left\{\overline{\mathrm{v}}_{\mathrm{i}}, \overline{\mathrm{~T}}\left(\overline{\mathrm{v}}_{\mathrm{i}}\right), \ldots, \overline{\mathrm{T}}^{\mathrm{dn}_{\mathrm{i}}-1}\left(\overline{\mathrm{v}}_{\mathrm{i}}\right)\right\} .
$$

Since $\overline{\mathrm{V}}=\mathrm{Z}\left(\overline{\mathrm{v}}_{2}, \overline{\mathrm{~T}}\right) \oplus \cdots \oplus \mathrm{Z}\left(\overline{\mathrm{V}}_{\mathrm{r}}, \overline{\mathrm{T}}\right)$, we see from Theorem 2.15 that $\overline{\mathrm{V}}$ has basis

$$
\left\{\overline{\mathrm{v}}_{2}, \ldots, \overline{\mathrm{~T}}^{\mathrm{dn}_{2}-1}\left(\overline{\mathrm{v}}_{2}\right), \ldots, \overline{\mathrm{v}}_{\mathrm{r}}, \ldots, \overline{\mathrm{~T}}^{\mathrm{d}_{\mathrm{r}}-1}\left(\overline{\mathrm{v}}_{\mathrm{r}}\right)\right\}
$$

Recall that $\bar{v}_{i}=u_{i}+Z_{1}$ and $v_{i} \in \bar{v}_{i}$. This means that $v_{i}=u_{i}+z_{i}$ for some $\mathrm{z}_{\mathrm{i}} \in \mathrm{Z}_{1}$ so that

$$
\overline{\mathrm{v}}_{\mathrm{i}}=\mathrm{v}_{\mathrm{i}}-\mathrm{z}_{\mathrm{i}}+\mathrm{Z}_{1}=\mathrm{v}_{\mathrm{i}}+\mathrm{Z}_{1}
$$

and hence (see the proof of Theorem 7.35)

$$
\overline{\mathrm{T}}^{\mathrm{m}}\left(\overline{\mathrm{v}}_{\mathrm{i}}\right)=\overline{\mathrm{T}}^{\mathrm{m}}\left(\mathrm{v}_{\mathrm{i}}+\mathrm{Z}_{1}\right)=\overline{\mathrm{T}}^{\overline{\mathrm{m}}}\left(\mathrm{v}_{\mathrm{i}}+\mathrm{Z}_{1}\right)=\mathrm{T}^{\mathrm{m}}\left(\mathrm{v}_{\mathrm{i}}\right)+\mathrm{Z}_{1} .
$$

Using this result in the terms for the basis of $\overline{\mathrm{V}}$, Theorem 7.34 shows that V has the basis (where we recall that $Z_{1}$ is just $Z\left(v_{1}, T\right)$ )

$$
\left\{\mathrm{v}_{1}, \ldots, \mathrm{~T}^{\mathrm{dn}_{1}-1}\left(\mathrm{v}_{1}\right), \mathrm{v}_{2}, \ldots, \mathrm{~T}^{\mathrm{dn}_{2}-1}\left(\mathrm{v}_{2}\right), \ldots, \mathrm{v}_{\mathrm{r}}, \ldots, \mathrm{~T}^{\mathrm{dn}_{\mathrm{r}}-1}\left(\mathrm{v}_{\mathrm{r}}\right)\right\}
$$

Therefore, by Theorem 2.15, V must be the direct sum of the $\mathrm{Z}_{\mathrm{i}}=\mathrm{Z}\left(\mathrm{v}_{\mathrm{i}}, \mathrm{T}\right)$ for $\mathrm{i}=1, \ldots, \mathrm{r}$. This completes the first part of the proof.

We now turn to the uniqueness of the direct sum expansion of $V$. Note that we have just shown that $\mathrm{V}=\mathrm{Z}_{1} \oplus \cdots \oplus \mathrm{Z}_{\mathrm{r}}$ where each $\mathrm{Z}_{\mathrm{i}}=\mathrm{Z}\left(\mathrm{v}_{\mathrm{i}}, \mathrm{T}\right)$ is a T cyclic subspace of V . In addition, the minimal polynomial (i.e., order) of $v_{i}$ is $p(x)^{n_{i}}$ where $p(x)$ is a monic prime polynomial of degree $d$, and $p(x)^{n}$ is the minimal polynomial for $T$. Let us assume that we also have the decomposition $\mathrm{V}=\mathrm{Z}_{1}^{\prime} \oplus \cdots \oplus \mathrm{Z}_{\mathrm{s}}^{\prime}$ where $\mathrm{Z}_{\mathrm{i}}^{\prime}=\mathrm{Z}\left(\mathrm{v}_{\mathrm{i}}, \mathrm{T}\right)$ is a T -cyclic subspace of V , and $\mathrm{v}_{\mathrm{i}}^{\prime}$ has minimal polynomial $p(x)^{m_{i}}$ with $m_{1} \geq \cdots \geq m_{s}$. (Both $v_{i}$ and $v_{i}^{\prime}$ have orders that are powers of the same polynomial $\mathrm{p}(\mathrm{x})$ by the corollary to Theorem 8.25.) We must show that $\mathrm{s}=\mathrm{r}$ and that $\mathrm{m}_{\mathrm{i}}=\mathrm{n}_{\mathrm{i}}$ for $\mathrm{i}=1, \ldots$, r .

Suppose that $n_{i} \neq m_{i}$ for at least one $i$, and let $k$ be the first integer such that $n_{k} \neq m_{k}$ while $n_{j}=m_{j}$ for $j=1, \ldots, k-1$. We may arbitrarily take $n_{k}>$
$\mathrm{m}_{\mathrm{k}}$. Since $\mathrm{V}=\mathrm{Z}_{1}^{\prime} \oplus \cdots \oplus \mathrm{Z}_{\mathrm{s}}^{\prime}$, any $\mathrm{u} \in \mathrm{V}$ may be written in the form $\mathrm{u}=$ $\mathrm{u}^{\prime}{ }_{1}+\cdots+\mathrm{u}^{\prime}{ }_{\mathrm{s}}$ where $\mathrm{u}^{\prime}{ }_{\mathrm{i}} \in \mathrm{Z}_{\mathrm{i}}{ }_{\mathrm{i}}$. Furthermore, since $\mathrm{p}(\mathrm{T})^{\mathrm{m}_{\mathrm{i}}}$ is linear, we see that

$$
\mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{i}}}(\mathrm{u})=\mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{i}}}\left(\mathrm{u}_{1}^{\prime}\right)+\cdots+\mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{i}}}\left(\mathrm{u}_{\mathrm{s}}^{\prime}\right)
$$

and hence we may write

$$
\mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{i}}}(\mathrm{~V})=\mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{i}}}\left(\mathrm{Z}^{\prime}{ }^{\prime}\right) \oplus \cdots \oplus \mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{i}}}\left(\mathrm{Z}_{\mathrm{s}}^{\prime}\right)
$$

Using the definition of the T-cyclic subspace $\mathrm{Z}_{\mathrm{k}}$ along with the fact that $p(T)^{m_{k}}\left(v_{k}^{\prime}\right)=0$, it is easy to see that $p(T)^{m_{k}}\left(Z_{k}^{\prime}\right)=0$. But the inequality $m_{k} \geq$ $m_{k+1} \geq \cdots \geq m_{s}$ implies that $p(T)^{m_{k}}\left(Z_{i}^{\prime}\right)=0$ for $i=k, k+1, \ldots, s$ and hence we have

$$
\mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{k}}}(\mathrm{~V})=\mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{k}}}\left(\mathrm{Z}_{\mathrm{t}}^{\prime}\right) \oplus \cdots \oplus \mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{k}}}\left(\mathrm{Z}_{\mathrm{k}-1}^{\prime}\right) .
$$

From Example 8.12, we see that $\mathrm{p}(\mathrm{T})^{\mathrm{m}_{\mathrm{i}}}\left(\mathrm{Z}_{\mathrm{j}}^{\prime}\right)$ is of dimension $\mathrm{d}\left(\mathrm{m}_{\mathrm{j}}-\mathrm{m}_{\mathrm{i}}\right)$ for $\mathrm{m}_{\mathrm{i}} \leq \mathrm{m}_{\mathrm{j}}$. This gives us (see the corollary to Theorem 2.15)

$$
\operatorname{dim} \mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{k}}}(\mathrm{~V})=\mathrm{d}\left(\mathrm{~m}_{1}-\mathrm{m}_{\mathrm{k}}\right)+\cdots+\mathrm{d}\left(\mathrm{~m}_{\mathrm{k}-1}-\mathrm{m}_{\mathrm{k}}\right) .
$$

On the other hand, we have $\mathrm{V}=\mathrm{Z}_{1} \oplus \cdots \oplus \mathrm{Z}_{\mathrm{r}}$, and since $\mathrm{k} \leq \mathrm{r}$ it follows that $\mathrm{Z}_{1} \oplus \cdots \oplus \mathrm{Z}_{\mathrm{k}} \subset \mathrm{V}$. Therefore

$$
\mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{k}}}(\mathrm{~V}) \supset \mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{k}}}\left(\mathrm{Z}_{1}\right) \oplus \cdots \oplus \mathrm{p}(\mathrm{~T})^{\mathrm{m}_{\mathrm{k}}}\left(\mathrm{Z}_{\mathrm{k}}\right)
$$

and hence, since $\operatorname{dim} p(T)^{m_{i}}\left(Z_{j}\right)=d\left(n_{j}-m_{i}\right)$ for $m_{i} \leq n_{j}$, we have

$$
\operatorname{dim} p(T)^{m_{k}}(V) \geq d\left(n_{1}-m_{k}\right)+\cdots+d\left(n_{k-1}-m_{k}\right)+d\left(n_{k}-m_{k}\right) .
$$

However, $\mathrm{n}_{\mathrm{i}}=\mathrm{m}_{\mathrm{i}}$ for $\mathrm{i}=1, \ldots, \mathrm{k}-1$ and $\mathrm{n}_{\mathrm{k}}>\mathrm{m}_{\mathrm{k}}$. We thus have a contradiction in the value of $\operatorname{dim} p(T)^{m_{k}}(V)$, and hence $n_{i}=m_{i}$ for every $i=1, \ldots, r$ (since if $\mathrm{r}<\mathrm{s}$ for example, then we would have $0=\mathrm{n}_{\mathrm{s}} \neq \mathrm{m}_{\mathrm{s}}$ for every $\mathrm{s}>\mathrm{r}$ ). This completes the entire proof of Theorem 8.32.

In order to prove Theorem 8.31 now, we must remove the requirement in Theorem 8.32 that $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ have minimal polynomial $\mathrm{p}(\mathrm{x})^{\mathrm{n}}$. For any finitedimensional $\mathrm{V} \neq\{0\}$, we know that any $\mathrm{T} \in \mathrm{L}(\mathrm{V})$ has a minimal polynomial $m(x)$ (Theorem 7.4). From the unique factorization theorem (Theorem 6.6), we know that any polynomial can be factored into a product of prime polynomials. We can thus always write $\mathrm{m}(\mathrm{x})=\mathrm{p}_{1}(\mathrm{x})^{\mathrm{n}_{1}} \cdots \mathrm{p}_{\mathrm{r}}(\mathrm{x})^{\mathrm{n}_{\mathrm{r}}}$ where each
$\mathrm{p}_{\mathrm{i}}(\mathrm{x})$ is prime. Hence, from the primary decomposition theorem (Theorem 7.23), we then see that V is the direct sum of T -invariant subspaces $\mathrm{W}_{\mathrm{i}}=$ $\operatorname{Ker} \mathrm{p}_{\mathrm{i}}(\mathrm{T})^{\mathrm{n}_{\mathrm{i}}}$ for $\mathrm{i}=1, \ldots, r$ such that minimal polynomial of $\mathrm{T}_{\mathrm{i}}=\mathrm{T} \mid \mathrm{W}_{\mathrm{i}}$ is $p_{i}(x)^{n_{i}}$.

Applying Theorem 8.32 to each space $\mathrm{W}_{\mathrm{i}}$ and operator $\mathrm{T}_{\mathrm{i}} \in \mathrm{L}\left(\mathrm{W}_{\mathrm{i}}\right)$, we see that there exist vectors $\mathrm{w}_{\mathrm{i}_{\mathrm{k}}} \in \mathrm{W}_{\mathrm{i}}$ for $\mathrm{k}=1, \ldots, \mathrm{r}_{\mathrm{i}}$ such that $\mathrm{W}_{\mathrm{i}}$ is the direct sum of the $\mathrm{Z}\left(\mathrm{w}_{\mathrm{i}_{\mathrm{k}}}, \mathrm{T}_{\mathrm{i}}\right)$. Moreover, since each $\mathrm{W}_{\mathrm{i}}$ is T-invariant, each of the $\mathrm{T}_{\mathrm{i}}{ }^{-}$ cyclic subspaces $\mathrm{Z}\left(\mathrm{w}_{\mathrm{i}_{\mathrm{k}}}, \mathrm{T}_{\mathrm{i}}\right)$ is also T -cyclic, and the minimal polynomial of each generator $\mathrm{w}_{\mathrm{i}_{\mathrm{k}}}$ is a power of $\mathrm{p}_{\mathrm{i}}(\mathrm{x})$. This discussion completes the proof of Theorem 8.31.

Finally, we remark that it is possible to prove the rational form of a matrix from this version (i.e., proof) of the elementary divisor theorem. However, we feel that at this point it is not terribly instructive to do so, and hence the interested reader will have to find this approach in one of the books listed in the bibliography.

## Exercise

Prove Theorem 8.31 using the rational canonical form.

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