# Linear Algebra 

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## Introduction

References for these notes include Hoffman and Kunze [1], Koç [2], Lang [3, 4], and Roman [5], but I may not follow them closely.
Since in set theory the letter $\omega$ denotes the set $\{0,1,2, \ldots\}$ of natural numbers, I let $\mathbb{N}$ denote the set $\{1,2,3, \ldots\}$ of counting numbers. For notational convenience, each $n$ in $\mathbb{N}$ is the set $\{0, \ldots, n-1\}$, which has $n$ elements. The expressions $i<n$ and $i \in n$ are interchangeable.
An expression like

$$
\bigwedge_{i<n} \varphi(i)
$$

means $\varphi(i)$ holds whenever $i<n$; that is,

$$
i<n \Longrightarrow \varphi(i) .
$$

The notation $f: A \rightarrow B$ is to be read as a sentence, " $f$ is a function from $A$ to $B$."

## 1 Determinants

### 1.1 Matrix multiplication

The structures $\mathbb{C}, \mathbb{R}, \mathbb{Q}, \mathbb{Z}$, and $\mathbb{Z} /(n)$, where $n \in \mathbb{N}$, where

$$
\mathbb{N}=\{x \in \mathbb{Z}: x>0\},
$$

are commutative rings.
For us, a ring will be a structure $(R, \cdot, 1)$, where

1) $R$ is an abelian group, written additively,
2) • is a multiplication on $R$, that is, a binary operation on $R$ that distributes from each side over addition,
3) • is associative, and
4) 1 is a two-sided identity with respect to •.

We usually write $(R, \cdot, 1)$ as $R$.
A unit of a ring is an invertible element, that is, an element with a left inverse and a right inverse. When these one-sided inverses exist, they are equal. The units of a ring $R$ compose a multiplicative group, denoted by

$$
R^{\times}
$$

A ring is commutative if its multiplication is commutative. We gave examples above. For an example of a group of units, we note that, for all $n$ in $\mathbb{N}$,

$$
\left.\left|\mathbb{Z} /(n)^{\times}\right|=\mid x \in \mathbb{Z} /(n): \operatorname{gcd}(x, n)=1\right\} \mid=\phi(n)
$$

A commutative ring $R$ is a field if $R^{\times}=R \backslash\{0\}$. If $p$ is prime, then $\mathbb{Z} /(p)$ is the field $\mathbb{F}_{p}$, and

$$
\mathbb{F}_{p} \times \cong \mathbb{Z}_{p-1},
$$

where in general $\mathbb{Z}_{n}$ is the cyclic group of order $n$, and $\mathbb{Z} /(n)$ means ( $\mathbb{Z}_{n}, \cdot, 1$ ).
In this chapter, we shall work with an arbitrary commutative ring $K$. The definition of a module over $K$ is the same as the definition of a vector space, when $K$ is a field. An abelian group is a module over $\mathbb{Z}$.
If $(m, n) \in \mathbb{N} \times \mathbb{N}$, then $K^{m \times n}$ and $K^{n}$ are modules over $K$, and

$$
(X, \boldsymbol{y}) \mapsto X \boldsymbol{y}: K^{m \times n} \times K^{n} \rightarrow K^{m},
$$

defined as follows.
If $\Omega$ is a set, we denote by

$$
K^{\Omega}
$$

the $K$-module of functions from $\Omega$ to $K$. This defines $K^{n}$ when we understand $n$ as the $n$-element set $\{0, \ldots, n-1\}$. An arbitrary element of $K^{n}$ is one of

$$
\left(a^{0}, \ldots, a^{n-1}\right), \quad\left(a^{j}: j \in n\right), \quad \boldsymbol{a} .
$$

The superscripts are row numbers, when we think of $\boldsymbol{a}$ as the $1 \times n$ matrix

$$
\left(\begin{array}{c}
a^{0} \\
\vdots \\
a^{n-1}
\end{array}\right) .
$$

Many persons understand $K^{n}$ as $K^{[n]}$, where $[n]$ is the set $\{1, \ldots, n\}$ with $n$ elements. What is important is that the
entries of an element of $K^{n}$ be functions into $K$ from a linearly ordered set with $n$ elements.

An element $A$ of $K^{m \times n}$ is a matrix of $m$ rows and $n$ columns, having entries $a_{j}^{i}$ from $K$, where $i \in m$ and $j \in n$, so

$$
A=\left(\begin{array}{ccc}
a_{0}^{0} & \cdots & a_{n-1}^{0} \\
\vdots & \ddots & \vdots \\
a_{0}^{m-1} & \cdots & a_{n-1}^{m-1}
\end{array}\right)=\left(a_{j}^{i}\right)_{j \in n}^{i \in m}
$$

If one prefers, one may work instead with elements of $E^{[m] \times[n]}$, and one may write $a_{i j}$ for $a_{j}^{i}$. If also $\boldsymbol{b} \in K^{n}$, we define

$$
\begin{equation*}
A \boldsymbol{b}=\left(\sum_{j \in n} a_{j}^{i} b^{j}: i \in m\right) \tag{1.1}
\end{equation*}
$$

an element of $K^{m}$. As in (1.1) with $j$, when an index appears twice, once raised and once lowered, it is usually being summed over. When $j \in n$, we define

$$
\begin{equation*}
\mathbf{e}_{j}=\left(\delta_{j}^{i}: i \in n\right) \tag{1.2}
\end{equation*}
$$

in the module $K^{n}$, where

$$
\delta_{j}^{i}= \begin{cases}1, & \text { if } i=j  \tag{1.3}\\ 0, & \text { if } i \neq j\end{cases}
$$

Then

$$
\begin{equation*}
A \mathbf{e}_{j}=\left(\sum_{k \in n} a_{k}^{i} \delta_{j}^{k}: i \in n\right)=\left(a_{j}^{i}: i \in n\right)=\boldsymbol{a}_{j} \tag{1.4}
\end{equation*}
$$

this being column $j$ of $A$. If $\boldsymbol{b} \in K^{n}$, then

$$
\begin{equation*}
\boldsymbol{b}=\sum_{j \in n} b^{j} \mathbf{e}_{j} . \tag{1.5}
\end{equation*}
$$

We denote by

$$
\tau_{A}
$$

the function $\boldsymbol{x} \mapsto A \boldsymbol{x}$ from $K^{n}$ to $K^{m}$.
To say that a function $\varphi$ from $K^{n}$ to $K^{m}$ is a linear transformation means that $\varphi$ is a homomorphism of modules over $K$, that is,

$$
\varphi(\boldsymbol{b}+\boldsymbol{c})=\varphi(\boldsymbol{b})+\varphi(\boldsymbol{c}), \quad \varphi(d \cdot \boldsymbol{b})=d \cdot \varphi(\boldsymbol{b}) .
$$

The linear transformations from $K^{n}$ to $K^{m}$ compose a module over $K$ denoted by

$$
\mathscr{L}\left(K^{n}, K^{m}\right) .
$$

Theorem 1. $X \mapsto \tau_{X}: K^{m \times n} \cong \mathscr{L}\left(K^{n}, K^{m}\right)$.
Proof. We have to check that
(1) $\tau_{A} \in \mathscr{L}\left(K^{n}, K^{m}\right)$ for each $A$ in $K^{m \times n}$;
(2) $X \mapsto \tau_{X}$ is a homomorphism;
(3) if $\tau_{A}=0$, then $A=0$;
(4) every member of $\mathscr{L}\left(K^{n}, K^{m}\right)$ is $\tau_{A}$ for some $A$ in $K^{m \times n}$. Each step in the verification of the first two points uses the definition of a $K$-module or a property of $K$ as a commutative ring. If $\tau_{A}=0$, this means in each case $\mathbf{0}=A \mathbf{e}_{j}$, which is column $j$ of $A$ by (1.4); so $A=0$.
Finally, since each $\tau_{A}$ is linear, from (1.4) and (1.5) we have

$$
A \boldsymbol{b}=\sum_{j \in n} b^{j} \boldsymbol{a}_{j} .
$$

If $T \in \mathscr{L}\left(K^{n}, K^{m}\right)$, by defining

$$
T \mathbf{e}_{j}=\boldsymbol{a}_{j}
$$

we obtain $A$, and then

$$
T=\tau_{A} .
$$

If still $A \in K^{m \times n}$, and now also $C \in K^{n \times s}$, then we define

$$
\begin{equation*}
A C=\left(\sum_{j \in n} a_{j}^{i} c_{k}^{j}\right)_{k \in s}^{i \in m} \tag{1.6}
\end{equation*}
$$

an element of $K^{m \times s}$. We shall let $M$ denote the special case $K^{n \times n}$, which is closed under matrix multiplication. We have

$$
\mathrm{I} A=A=A \mathrm{I}
$$

where

$$
\begin{equation*}
\mathrm{I}=\left(\delta_{j}^{i}\right)_{j \in n}^{i \in n} . \tag{1.7}
\end{equation*}
$$

Theorem 2. When $A \in K^{m \times n}$ and $C \in K^{n \times s}$, then

$$
\tau_{A C}=\tau_{A} \circ \tau_{C}
$$

Thus for any matrices $A, B$, and $C$ for which either of the products $(A B) C$ and $A(B C)$ is defined, then both are defined, and they are equal. In particular, the structure $(M, \cdot, \mathrm{I})$ is a ring, and $X \mapsto \tau_{X}$ from $M$ to $\mathscr{L}\left(K^{n}, K^{n}\right)$ is an isomorphism of rings.

### 1.2 Determinants

We use the possibility of Gauss-Jordan elimination to motivate the so-called Leibniz formula (1.19) for the determinant.

### 1.2.1 Desired Properties

Let $M$ be the ring $K^{n \times n}$. We want to define a determinant function,

$$
X \mapsto \operatorname{det} X
$$

from $M$ to $K$ so that

$$
\begin{equation*}
\operatorname{det} X \in K^{\times} \Longleftrightarrow X \in M^{\times} . \tag{1.8}
\end{equation*}
$$

If $K$ is the two-element field $\mathbb{F}_{2}$, then (1.8) is equivalent to

$$
\operatorname{det} X= \begin{cases}1, & \text { if } X \in M^{\times}  \tag{1.9}\\ 0, & \text { otherwise }\end{cases}
$$

Moreover, with this definition,

$$
\begin{equation*}
\operatorname{det}(X Y)=\operatorname{det} X \operatorname{det} Y . \tag{1.10}
\end{equation*}
$$

However, over any $K$, we also want

$$
\begin{equation*}
\operatorname{det} X=f\left(x_{j}^{i}:(i, j) \in n \times n\right) \tag{1.11}
\end{equation*}
$$

for some polynomial $f$ (namely an element of the free abelian group generated by products of the variables $x_{j}^{i}$ ). In general then, (1.9) will fail. We still want (1.10) to hold, and this and (1.8) imply

$$
\begin{equation*}
\operatorname{det} \mathrm{I}=1 . \tag{1.12}
\end{equation*}
$$

### 1.2.2 Additional properties

In seeking a determinant function satisfying (1.8), (1.10), and (1.11), and therefore (1.12), we consider what we know about $M^{\times}$. An element $A$ of $M$ is in $M^{\times}$just in case $A$ is rowequivalent to I. This means, for some elementary matrices $E_{i}$,

$$
\begin{equation*}
A=E_{1} \cdots E_{n} \mathrm{I} . \tag{1.13}
\end{equation*}
$$

Thus, if (1.10) and (1.12) hold, then $\operatorname{det} A$ will determined by the $\operatorname{det} E_{i}$.

We recall that an elementary matrix is the result of applying to I an elementary row operation. If $\Phi$ is such, then

$$
\Phi(I) A=\Phi(A) .
$$

Here $\Phi$ does one of the following:

1) add to one row another row, scaled by some $a$ in $K$;
2) interchange two rows;
3) scale a row by an element $s$ of $K^{\times}$.

Let us denote the specific instance of $\Phi$ respectively by

$$
\Phi_{a}, \quad \Psi, \quad \Theta_{s}
$$

We do not specify the row or rows involved. We draw the following conclusions about determinants.

1. If (1.11) is to hold, then, for some single-variable polynomial $f$,

$$
\operatorname{det} \Phi_{a}(\mathrm{I})=f(a) .
$$

If also (1.10) is to hold, then, since

$$
\Phi_{a}(\mathrm{I}) \cdot \Phi_{b}(\mathrm{I})=\Phi_{a+b}(\mathrm{I}),
$$

we must have

$$
f(a) \cdot f(b)=f(a+b) .
$$

In particular, $f(x)^{n}=f(n x)$ for all $n$ in $\mathbb{N}$, and so, since $f \neq 0$, we must have

$$
\begin{equation*}
\operatorname{det} \Phi_{a}(\mathrm{I})=1 \tag{1.14}
\end{equation*}
$$

2. If, again, (1.10) is to hold, then, since

$$
\Psi(\mathrm{I}) \cdot \Psi(\mathrm{I})=\mathrm{I},
$$

we should have $\operatorname{det} \Psi(\mathrm{I})= \pm 1$; we choose

$$
\begin{equation*}
\operatorname{det} \Psi(\mathrm{I})=-1 \tag{1.15}
\end{equation*}
$$

3. If, again (1.11) is to hold, then, for some single-variable polynomial $g$,

$$
\operatorname{det} \Theta_{s}(\mathrm{I})=g(s)
$$

If also (1.10) is to hold, then, since

$$
\Theta_{s}(\mathrm{I}) \cdot \Theta_{t}(\mathrm{I})=\Theta_{s t}(\mathrm{I})
$$

we must have

$$
g(s) \cdot g(t)=g(s t)
$$

In particular, $g(x)^{n}=g\left(x^{n}\right)$, so $\operatorname{det} \Theta_{s}(\mathrm{I})$ must be a power of $s$; we choose

$$
\begin{equation*}
\operatorname{det} \Theta_{s}(\mathrm{I})=s \tag{1.16}
\end{equation*}
$$

The definitions, or choices, (1.14), (1.15), and (1.16) will follow if $X \mapsto \operatorname{det} X$ is an alternating multilinear form.

We can understand any module $K^{m \times n}$ as $\left(K^{m}\right)^{n}$ or $\left(K^{n}\right)^{m}$, treating an element $A$ as one of

$$
\left(\left(a_{j}^{i}: i \in m\right): j \in n\right), \quad\left(\left(a_{j}^{i}: j \in n\right): i \in m\right)
$$

Given a module $V$ over $K$ and $n$ in $\mathbb{N}$, we can form the module $V^{n}$. For each $k$ in $n$, we let $\pi_{k}$ the function from $V^{n}$ to $V$ given by

$$
\pi_{k}\left(\boldsymbol{x}_{j}: j \in n\right)=\boldsymbol{x}_{k}
$$

Suppose now

$$
\varphi: V^{n} \rightarrow K
$$

Given $k$ in $n$ and a function $j \mapsto \boldsymbol{a}_{j}$ from $n \backslash\{k\}$ to $V$, we let $\iota$ be the function from $V$ to $V^{n}$ given by the rule that, for each $j$ in $n$,

$$
\pi_{j}(\mathfrak{l}(\boldsymbol{x}))= \begin{cases}\boldsymbol{x}, & \text { if } j=k, \\ \boldsymbol{a}_{j}, & \text { if } j \in n \backslash\{k\} .\end{cases}
$$

If the function $\boldsymbol{x} \mapsto \varphi(\llcorner(\boldsymbol{x}))$ is always linear, then $\varphi$ itself is a multilinear form, specifically an $n$-linear form, on $V$. If, further, whenever $i<j<n$,

$$
\boldsymbol{x}_{i}=\boldsymbol{x}_{j} \Longrightarrow \varphi\left(\boldsymbol{x}_{k}: k \in n\right)=0,
$$

then $\varphi$ is alternating as a multilinear form.
We let the group of permutations of a set $\Omega$ be

$$
\operatorname{Sym}(\Omega) .
$$

If $\Omega$ is finite, then $\operatorname{Sym}(\Omega)$ is generated by transpositions. If $\sigma \in \operatorname{Sym}(n)$, we define

$$
\begin{equation*}
\operatorname{sgn}(\sigma)=(-1)^{\mid(i, j) \in n \times n: i<j \& \sigma(i)>\sigma(j)\} \mid}, \tag{1.17}
\end{equation*}
$$

one of the elements of $\mathbb{Z}^{\times}$.
Theorem 3. For every $n$ in $\mathbb{N}$, the function $\xi \mapsto \operatorname{sgn}(\xi)$ on $\operatorname{Sym}(n)$

1) is given by

$$
\begin{equation*}
\operatorname{sgn}(\sigma)=\prod_{i \in j \in n} \frac{\sigma(i)-\sigma(j)}{i-j}, \tag{1.18}
\end{equation*}
$$

2) is a homomorphism onto $\mathbb{Z}^{\times}$, and
3) takes every transposition to -1 .

Proof. 1. Since

$$
\prod_{i \in j \in n} \frac{\sigma(i)-\sigma(j)}{i-j}=\frac{\prod_{i \in j \in n}(\sigma(i)-\sigma(j))}{\prod_{i \in j \in n}(i-j)}= \pm 1,
$$

(1.17) follows from (1.18).
2. Note

$$
\begin{aligned}
\operatorname{sgn}(\tau \sigma) & =\prod_{i \in j \in n} \frac{\tau \sigma(i)-\tau \sigma(j)}{i-j} \\
& =\prod_{i \in j \in n}\left(\frac{\tau \sigma(i)-\tau \sigma(j)}{\sigma(i)-\sigma(j)} \cdot \frac{\sigma(i)-\sigma(j)}{i-j}\right) \\
& =\prod_{i \in j \in n} \frac{\tau(i)-\tau(j)}{i-j} \cdot \operatorname{sgn}(\sigma)=\operatorname{sgn}(\tau) \cdot \operatorname{sgn}(\sigma)
\end{aligned}
$$

3. Letting

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

since every transposition is $\sigma^{-1} \cdot \tau \cdot \sigma$ for some $\sigma$, it is enough to note that

$$
\operatorname{sgn}(\tau)=-1
$$

since

$$
\frac{\tau(i)-\tau(j)}{i-j} \begin{cases}>0, & \text { when }(i, j) \neq(0,1) \\ <0, & \text { when }(i, j)=(0,1)\end{cases}
$$

An element $\sigma$ of $\operatorname{Sym}(n)$ is even if $\operatorname{sgn}(\sigma)=1$; this means $\sigma$ is a product of an even number of transpositions. The even permutations compose the subgroup of $\operatorname{Sym}(n)$ denoted by

$$
\operatorname{Alt}(n)
$$

Theorem 4. For any module $V$ over $K$, for any $n$ in $\mathbb{N}$, for any $n$-linear form $\varphi$ on $V$, for each $\sigma$ in $\operatorname{Sym}(n)$,

$$
\varphi\left(\boldsymbol{x}_{\sigma(j)}: j \in n\right)=\operatorname{sgn}(\sigma) \cdot \varphi\left(\boldsymbol{x}_{j}: j \in n\right)
$$

Proof. Every permutation of a finite set being a product of transpositions, we need only prove the claim when $n=2$ and $\sigma$ is the nontrivial permutation of 2 . Assuming

$$
\boldsymbol{x}=\boldsymbol{y} \Longrightarrow \varphi(\boldsymbol{x}, \boldsymbol{y})=0
$$

we have $0=\varphi(\boldsymbol{x}+\boldsymbol{y}, \boldsymbol{x}+\boldsymbol{y})$, but the latter is

$$
\varphi(\boldsymbol{x}, \boldsymbol{x})+\varphi(\boldsymbol{x}, \boldsymbol{y})+\varphi(\boldsymbol{y}, \boldsymbol{x})+\varphi(\boldsymbol{y}, \boldsymbol{y})
$$

which reduces to $\varphi(\boldsymbol{x}, \boldsymbol{y})+\varphi(\boldsymbol{y}, \boldsymbol{x})$.
In particular, if $\sigma \in \operatorname{Alt}(n)$, then

$$
\varphi\left(\boldsymbol{x}_{\sigma(j)}: j \in n\right)=\varphi\left(\boldsymbol{x}_{j}: j \in n\right)
$$

### 1.2.3 Existence and uniqueness

Theorem 5. There is at most one alternating multilinear function $X \mapsto \operatorname{det} X$ from $M$ to $K$ that satisfies (1.12), and if it does exist, it satisfies satisfies (1.8) and (1.10).

Proof. The hypotheses ensure (1.14), (1.15), and (1.16), as well as (1.12). Then (1.10) holds when $X$ is elementary, and therefore it holds for all $X$, and also (1.8) holds by the analysis (1.13) and since every non-invertible matrix is row-equivalent to one with a zero row.

We now show that there is at least one function $X \mapsto \operatorname{det} X$ as desired. We define

$$
\begin{equation*}
\operatorname{det} X=\sum_{\sigma \in \operatorname{Sym}(n)} \operatorname{sgn}(\sigma) \prod_{i \in n} x_{\sigma(i)}^{i} \tag{1.19}
\end{equation*}
$$

Thus (1.11) holds.

Theorem 6. For all $A$ in $M$,

$$
\operatorname{det}\left(A^{\mathrm{t}}\right)=\operatorname{det} A
$$

Proof. Since $\operatorname{sgn}\left(\sigma^{-1}\right)=\operatorname{sgn}(\sigma)$, we compute

$$
\begin{aligned}
\operatorname{det}\left(A^{\mathrm{t}}\right) & =\sum_{\sigma \in \operatorname{Sym}(n)} \operatorname{sgn}(\sigma) \prod_{i \in n} a_{i}^{\sigma(i)} \\
& =\sum_{\sigma \in \operatorname{Sym}(n)} \operatorname{sgn}\left(\sigma^{-1}\right) \prod_{i \in n} a_{\sigma^{-1}(i)}^{i},
\end{aligned}
$$

which is $\operatorname{det} A$.
Theorem 7. The function given by (1.19) is n-linear and alternating, and satisfies (1.12).

Proof. By (1.7), since

$$
\prod_{i \in n} \delta_{\sigma(i)}^{i}=0 \Longleftrightarrow \sigma \neq \mathrm{id}_{n}
$$

(1.12) holds. For multilinearity, Suppose matrices $A, B$, and $C$ agree everywhere but in some row $k$, and $a_{j}^{k}=s \cdot b_{j}^{k}+t \cdot c_{j}^{k}$ for each $j$ in $n$, for some $s$ and $t$ in $K$. Then

$$
\begin{array}{r}
\operatorname{det} A=\sum_{\sigma \in \operatorname{Sym}(n)} \operatorname{sgn}(\sigma) \prod_{i \in n \backslash\{k\}} a_{\sigma(i)}^{i} \cdot\left(s \cdot b_{\sigma(k)}^{k}+t \cdot c_{\sigma(k)}^{k}\right) \\
=s \cdot \operatorname{det} B+t \cdot \operatorname{det} C .
\end{array}
$$

Finally, if $i<j<n$, and $\tau$ in $\operatorname{Sym}(n)$ transposes $i$ and $j$, then $\tau^{-1}=\tau$, and $\xi \mapsto \xi \circ \tau$ is a bijection between $\operatorname{Alt}(n)$ and $\operatorname{Sym}(n) \backslash \operatorname{Alt}(n)$, so

$$
\operatorname{det} A=\sum_{\sigma \in \operatorname{Alt}(n)}\left(\prod_{k \in n} a_{\sigma(k)}^{k}-\prod_{k \in n} a_{\sigma(k}^{\tau(k)}\right)
$$

If moreover $a_{k}^{i}=a_{k}^{j}$ for each $k$ in $n$, then $\operatorname{det} A=0$.

### 1.3 Inversion

We know from Theorems 5 and 7 that (1.8) holds. In particular, if $\operatorname{det} A \in K^{\times}$, then $A^{-1}$ exists in $M$. We can compute $A^{-1}$ by the reduction in (1.13); but we now develop another method.

As in (1.17), if $\tau$ is a bijection from a finite ordered set $S$ to a finite ordered set $T$, we can define

$$
\operatorname{sgn}(\tau)=(-1)^{\mid(i, j) \in S \times S: i<j \& \sigma(i)>\sigma(j)\} \mid}
$$

There is a unique isomorphism $\varphi$ from $S$ to $T$, and then

$$
\begin{gathered}
\varphi^{-1} \circ \tau \in \operatorname{Sym}(S) \\
\operatorname{sgn}(\tau)=\operatorname{sgn}\left(\varphi^{-1} \circ \tau\right)
\end{gathered}
$$

Suppose now $\sigma \in \operatorname{Sym}(n)$ and $k \in n$. Letting $S$ be $n \backslash\{k\}$ and $T$ be $n \backslash\{\sigma(k)\}$, we can define $\tau$ to be the restriction of $\sigma$ to $S$, so that $\tau$ is a bijection from $S$ to $T$. Then

$$
\frac{\operatorname{sgn}(\sigma)}{\operatorname{sgn}(\tau)}=(-1)^{|\{j \in n \backslash\{k\}: j>k \Longleftrightarrow \sigma(j)<\sigma(k)\}|}
$$

Theorem 8. In the notation above,

$$
\frac{\operatorname{sgn}(\sigma)}{\operatorname{sgn}(\tau)}=(-1)^{k+\sigma(k)}
$$

Proof. We may assume $k \leqslant \sigma(k)$. There are at least $\sigma(k)-k$ values of $j$ greater than $k$ and the condition

$$
\begin{equation*}
j>k \Longleftrightarrow \sigma(j)<\sigma(k) \tag{1.20}
\end{equation*}
$$

is satisfied. For every additional such value, there must be a value less than $k$ for which (1.20) is satisfied. This proves the claim.

For any $(k, \ell)$ in $n \times n$, assuming $n>1$, we let $\hat{A}_{\ell}^{k}$ be the matrix that we obtain from $A$ by deleting row $k$ and column $\ell$. Formally,

$$
\hat{A}_{\ell}^{k}=\left(a_{[j, \ell]}^{[i, k]}\right)_{j \in n-1}^{i \in n-1},
$$

where

$$
[i, k]= \begin{cases}i, & \text { if } i<k \\ i+1, & \text { if } k \leqslant i\end{cases}
$$

Theorem 9. For any $k$ in $n$,

$$
\operatorname{det} X=\sum_{j \in n}(-1)^{k+j} x_{j}^{k} \operatorname{det} \hat{X}_{j}^{k}
$$

Proof. We group the terms in (1.19), which are indexed by $\sigma$ in $\operatorname{Sym}(n)$, according to the value of $\sigma(k)$ :

$$
\begin{aligned}
\operatorname{det} X & =\sum_{j \in n} \sum_{\substack{\sigma \in \operatorname{Sym}(n) \\
\sigma(k)=j}} \operatorname{sgn}(\sigma) \prod_{i \in n} x_{\sigma(i)}^{i} \\
& =\sum_{j \in n} x_{j}^{k} \sum_{\substack{\sigma \in \operatorname{Sym}(n) \\
\sigma(k)=j}} \operatorname{sgn}(\sigma) \prod_{\substack{i \in n \\
i \neq k}} x_{\sigma(i)}^{i} \\
& =\sum_{j \in n}(-1)^{k+j} x_{j}^{k} \operatorname{det} \hat{X}_{j}^{k}
\end{aligned}
$$

by Theorem 8.
We now define the operation $X \mapsto \operatorname{adj}(X)$ on $M$ by

$$
\operatorname{adj}(A)=\left((-1)^{i+j} \operatorname{det} \hat{A}_{i}^{j}\right)_{j \in n}^{i \in n}
$$

This is the adjugate of $A$.

Theorem 10. For all $A$ in $M$,

$$
A \operatorname{adj}(A)=\operatorname{det} A \cdot \mathrm{I} .
$$

Proof. By Theorem 9, if $A \operatorname{adj}(A)=B$, then $b_{j}^{i}$ is the determinant of the matrix that we obtain from $A$ by replacing row $j$ with row $i$. This determinant is

- $\operatorname{det} A$, if $i=j$;
- 0 , if $i \neq j$, since $X \mapsto \operatorname{det} X$ is alternating.

Theorem 11. If $\operatorname{det} A \in K^{\times}$, then

$$
A^{-1}=(\operatorname{det} A)^{-1} \cdot \operatorname{adj}(A) .
$$

Proof. Assuming $\operatorname{det} A \in K^{\times}$, if we denote $(\operatorname{det} A)^{-1} \cdot \operatorname{adj}(A)$ by $B$, then by Theorem 10,

$$
A B=\mathrm{I} .
$$

Since $A^{-1}$ does exist, we have

$$
A^{-1}=A^{-1}(A B)=\left(A^{-1} A\right) B=\mathrm{I} B=B .
$$

## 2 Polynomials

### 2.1 Characteristic values

We henceforth suppose $K$ is a field; still $M$ is $K^{n \times n}$. For any $A$ in $M$, an element $\lambda$ of $K$ is a characteristic value or eigenvalue of $A$ if, for some $\boldsymbol{b}$ in $K^{n}$,

$$
\begin{equation*}
A \boldsymbol{b}=\lambda \cdot \boldsymbol{b} . \tag{2.1}
\end{equation*}
$$

In this case, $\boldsymbol{b}$ is a characteristic vector or eigenvector of $A$ associated with $\lambda$. Rewriting (2.1) as

$$
(A-\lambda \cdot \mathrm{I}) \boldsymbol{b}=\mathbf{0}
$$

shows that the characteristic values of $A$ are precisely the zeroes of the polynomial

$$
\operatorname{det}(A-x \cdot \mathrm{I}),
$$

which is called the characteristic polynomial of $A$.
If $\lambda$ is indeed a characteristic value of $A$, then the null-space of $A-\lambda \cdot \mathrm{I}$ is the characteristic space or eigenspace of $A$ associated with $\lambda$.

Theorem 12. Eigenvectors corresponding to distinct eigenvalues of any matrix are linearly independent.

Proof. We prove the claim by induction on the number of eigenvectors. The empty set of eigenvectors is trivially linearly
independent. Suppose ( $\left.\boldsymbol{v}_{i}: i<k\right)$ is linearly independent, each $\boldsymbol{v}_{i}$ being an eigenvector of $A$ with associated eigenvalue $\lambda_{i}$, the $\lambda_{i}$ being distinct. Let $\boldsymbol{v}_{k}$ be a an eigenvector associated with a new eigenvalue, $\lambda_{k}$. If

$$
\begin{equation*}
\sum_{i \leqslant k} x^{i} \boldsymbol{v}_{i}=\mathbf{0} \tag{2.2}
\end{equation*}
$$

then

$$
\begin{aligned}
0=\left(A-\lambda_{k} \cdot \mathrm{I}\right) \sum_{i<m+1} x^{i} \boldsymbol{v}_{i} & =\sum_{i \leqslant k}\left(\lambda_{i}-\lambda_{k}\right) x^{i} \boldsymbol{v}_{i} \\
& =\sum_{i<k}\left(\lambda_{i}-\lambda_{k}\right) x^{i} \boldsymbol{v}_{i}
\end{aligned}
$$

so $x^{i}=0$ when $i<k$, and then also $x^{k}=0$ by (2.2).
If $A$ in $M$ has $n$ linearly independent eigenvectors $\boldsymbol{b}_{i}$, each associated with an eigenvalue $\lambda_{i}$ (possibly not distinct), then the eigenvectors are the columns of an element $B$ of $M^{\times}$, and

$$
A B=B L
$$

where

$$
\ell_{j}^{i}= \begin{cases}\lambda_{i}, & \text { if } i=j \\ 0, & \text { if } i \neq j\end{cases}
$$

or in short

$$
L=\operatorname{diag}\left(\lambda_{i}: i \in n\right)
$$

a diagonal matrix. Thus

$$
B^{-1} A B=\operatorname{diag}\left(\lambda_{i}: i \in n\right)
$$

and in particular $A$ is diagonalizable.

It will be useful to recall that every matrix $B$ in $M^{\times}$is the change-of-basis matrix from the basis ( $j: j \in n$ ) of $K^{n}$ consisting of the columns of $B$ to the standard basis of $K^{n}$.
Every matrix of the form $P^{-1} A P$ for some $P$ in $M^{\times}$is similar to $A$ (in group theory one says conjugate). Similarity of matrices is an equivalence relation, as is row-equivalence (mentioned first on page 9); but they are different relations. We want to characterize the diagonalizable matrices.
A matrix $A$ in $M$ is triangular if

$$
\begin{equation*}
\bigwedge_{j<i<n} a_{j}^{i}=0 . \tag{2.3}
\end{equation*}
$$

A matrix similar to a triangular matrix is triangularizable.
Theorem 13. A matrix $A$ in $M$ is triangularizable just in case, for some $B$ in $M^{\times}$,

$$
\begin{equation*}
\bigwedge_{j \in n} A \boldsymbol{b}_{j} \in \operatorname{span}\left\{\boldsymbol{b}_{0}, \ldots, \boldsymbol{b}_{j}\right\} \tag{2.4}
\end{equation*}
$$

and in this case $B^{-1} A B$ is triangular.
Proof. The condition (2.3) on $A$ for being triangular means precisely

$$
\begin{equation*}
\bigwedge_{j \in n} A \mathbf{e}_{j}=\sum_{i=0}^{j} a_{j}^{i} \mathbf{e}_{i}, \tag{2.5}
\end{equation*}
$$

and thus that I is a matrix $B$ as in the statement of the theorem. If $B^{-1} A B$ is triangular, then putting this matrix in place of $A$ in (2.5) yields (2.4). Conversely, if $B$ is as in the statement, then we can write (2.4) as

$$
\bigwedge_{j \in n} A B \mathbf{e}_{j} \in \operatorname{span}\left\{B \mathbf{e}_{0}, \ldots, B \mathbf{e}_{j}\right\},
$$

and then

$$
\bigwedge_{j \in n} B^{-1} A B \mathbf{e}_{j} \in \operatorname{span}\left\{\mathbf{e}_{0}, \ldots, \mathbf{e}_{j}\right\},
$$

so $B^{-1} A B$ is triangular.
Theorem 14. Every matrix in $M$ is triangularizable over an algebraically closed field.

Proof. Given $A$ in $M$, assuming $K$ is algebraically closed, so that the characteristic polynomial of $A$ has at least one zero, and therefore $A$ has at least one eigenvector, we extend this to a basis of $K^{n}$ that satisfies (2.4). Doing this will be enough, by Theorem 13 .

We use induction on $n$. The claim is trivial when $n=1$. Suppose it holds when $n=m$. Now let $n=m+1$ and $A \in M$. There is a basis $\left(\boldsymbol{p}_{0}, \ldots, \boldsymbol{p}_{m}\right)$ of $K^{n}$ such that $\boldsymbol{p}_{0}$ is an eigenvector. Thus the basis satisfies the first conjunct of (2.4). We could satisfy the remaining conjuncts, by the inductive hypothesis, if we had

$$
\bigwedge_{j=1}^{m} A \boldsymbol{p}_{j} \in \operatorname{span}\left\{\boldsymbol{p}_{1}, \ldots, \boldsymbol{p}_{m}\right\} .
$$

However, we may not actually have this. Nonetheless, there are matrices $B$ and $C$ such that

$$
\begin{equation*}
\boldsymbol{\tau}_{B}\left(\sum_{i=0}^{m} x^{i} \boldsymbol{p}_{i}\right)=x_{0} \boldsymbol{p}_{0}, \quad \boldsymbol{\tau}_{C}\left(\sum_{i=0}^{m} x^{i} \boldsymbol{p}_{i}\right)=\sum_{i=1}^{m} x^{i} \boldsymbol{p}_{i} . \tag{2.6}
\end{equation*}
$$

In words,

- $\boldsymbol{\tau}_{C}$ is an endomorphism of $\operatorname{span}\left\{\boldsymbol{p}_{1}, \ldots, \boldsymbol{p}_{m}\right\}$, and therefore so is $\tau_{C A}$;
- $\boldsymbol{\tau}_{B}$ is a homomorphism from $K^{n}$ to $\operatorname{span}\left\{\boldsymbol{p}_{0}\right\}$, and therefore so is $\tau_{B A}$.
Now span $\left\{\boldsymbol{p}_{1}, \ldots, \boldsymbol{p}_{m}\right\}$ has a basis $\left(\boldsymbol{v}_{1} \ldots, \boldsymbol{v}_{m}\right)$ such that

$$
\bigwedge_{j=1}^{m} C A \boldsymbol{v}_{j} \in \operatorname{span}\left\{\boldsymbol{v}_{1} \ldots, \boldsymbol{v}_{j}\right\}
$$

by inductive hypothesis. Therefore now

$$
\bigwedge_{j=1}^{m}(B A+C A) \boldsymbol{v}_{j} \in \operatorname{span}\left\{\boldsymbol{v}_{0}, \ldots, \boldsymbol{v}_{j}\right\}
$$

From all of (2.6),

$$
\tau_{B}+\tau_{C}=\operatorname{id}_{K^{n}}
$$

and so

$$
\bigwedge_{j=1}^{m} A \boldsymbol{v}_{j} \in \operatorname{span}\left\{\boldsymbol{v}_{0}, \ldots, \boldsymbol{v}_{j}\right\}
$$

Finally, since $\boldsymbol{v}_{0}$ is an eigenvector of $A$,

$$
\bigwedge_{j=0}^{m} A \boldsymbol{v}_{j} \in \operatorname{span}\left\{\boldsymbol{v}_{0}, \ldots, \boldsymbol{v}_{j}\right\}
$$

Thus we have (2.4). This completes the induction.
We can write out the foregoing proof entirely in terms of matrices as follows. We have

$$
P^{-1} A P=\left(\begin{array}{c|c}
\lambda & \boldsymbol{a} \\
\hline \mathbf{0} & D
\end{array}\right)
$$

for some $m \times m$ matrix $D$, where $\lambda$ is the eigenvalue associated with $\boldsymbol{p}_{0}$. We choose $B$ and $C$ so that

$$
P^{-1} B P=\left(\begin{array}{l|l}
1 & \mathbf{0} \\
\hline \mathbf{0} & 0
\end{array}\right), \quad \quad P^{-1} C P=\left(\begin{array}{c|c}
0 & \mathbf{0} \\
\hline \mathbf{0} & \mathrm{I}
\end{array}\right) .
$$

Then

$$
\begin{array}{r}
P^{-1} C A P=P^{-1} C P P^{-1} A P=\left(\begin{array}{c|c}
0 & \mathbf{0} \\
\hline \mathbf{0} & \mathrm{I}
\end{array}\right)\left(\begin{array}{l|l}
\lambda & \boldsymbol{a} \\
\hline \mathbf{0} & D
\end{array}\right) \\
=\left(\begin{array}{l|l}
0 & \mathbf{0} \\
\hline \mathbf{0} & D
\end{array}\right)
\end{array}
$$

and

$$
P^{-1} B A P=\left(\begin{array}{c|c}
1 & 0 \\
\hline \mathbf{0} & 0
\end{array}\right)\left(\begin{array}{c|c}
\lambda & \boldsymbol{a} \\
\hline \mathbf{0} & D
\end{array}\right)=\left(\begin{array}{c|c}
\lambda & \boldsymbol{a} \\
\hline \mathbf{0} & 0
\end{array}\right) .
$$

Therefore

$$
\begin{gathered}
P^{-1} B A P+P^{-1} C A P=P^{-1} A P \\
B A+C A=A .
\end{gathered}
$$

By inductive hypothesis, for some $Q, Q^{-1} D Q$ is a triangular matrix $T$. Then

$$
\left(\begin{array}{c|c}
1 & \mathbf{0} \\
\hline \mathbf{0} & Q
\end{array}\right)^{-1} P^{-1} C A P\left(\begin{array}{c|c}
1 & \mathbf{0} \\
\hline \mathbf{0} & Q
\end{array}\right)=\left(\begin{array}{c|c}
0 & \mathbf{0} \\
\hline \mathbf{0} & T
\end{array}\right),
$$

while

$$
\begin{aligned}
& \left(\begin{array}{c|c}
1 & \mathbf{0} \\
\hline \mathbf{0} & Q
\end{array}\right)^{-1} \quad P^{-1} B A P\left(\begin{array}{c|c}
1 & \mathbf{0} \\
\hline \mathbf{0} & Q
\end{array}\right) \\
& \quad=\left(\begin{array}{c|c}
1 & \mathbf{0} \\
\hline \mathbf{0} & Q^{-1}
\end{array}\right)\left(\begin{array}{c|c}
\lambda & \boldsymbol{a} Q \\
\hline \mathbf{0} & 0
\end{array}\right)=\left(\begin{array}{c|c}
\lambda & \boldsymbol{a} Q \\
\hline \mathbf{0} & 0
\end{array}\right)
\end{aligned}
$$

and therefore

$$
\left(\begin{array}{c|c}
1 & \mathbf{0} \\
\hline \mathbf{0} & Q
\end{array}\right)^{-1} P^{-1} A P\left(\begin{array}{c|c}
1 & \mathbf{0} \\
\hline \mathbf{0} & Q
\end{array}\right)=\left(\begin{array}{c|c}
\lambda & \boldsymbol{a} Q \\
\hline \mathbf{0} & T
\end{array}\right),
$$

a triangular matrix.

### 2.2 Polynomial functions of matrices

Although $K$ is a field, the ring $M$ is not commutative when $n>1$. However, it has commutative sub-rings. Indeed, for every $A$ in $M$, the smallest sub-ring of $M$ that contains $A$ is commutative. We may denote this sub-ring by

$$
K[A] .
$$

This is also a vector space over $K$, spanned by the powers I, $A, A^{2}, A^{3}, \ldots$, of $A$. Thus

$$
K[A]=\{f(A): f \in K[x]\},
$$

where, if

$$
\begin{equation*}
f(x)=\sum_{i=0}^{m} f_{i} x^{i} \tag{2.7}
\end{equation*}
$$

in $K[x]$ (and $x^{i}$ is now the power $\prod_{k \in i} x$ ), we define

$$
f(A)=\sum_{i=0}^{n} f_{i} A^{i} .
$$

If $f(A)$ is the zero matrix, we may say $A$ is a zero of $f$. However, theorems about zeroes in fields may not apply here. For example, since $K[A]$ may have zero divisors, the number of zeroes of $f$ in $M$ may exceed $\operatorname{deg} f$. Indeed, $A$ itself may be a zero divisor, as for example when

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

since then $A^{2}$ is the zero matrix. In this case every scalar multiple $b \cdot A$ of $A$ is a zero in $K[A]$ of the polynomial $x^{2}$.

### 2.3 Cayley-Hamilton Theorem

Given $A$ in $M$, we are going to want to know that $A$ is a zero of some nonzero polynomial over $K$. Suppose

$$
\begin{equation*}
f(x)=\operatorname{det}(x \cdot I-A) \tag{2.8}
\end{equation*}
$$

the characteristic polynomial of $A$. The equation remains correct automatically when we replace $x$ with an element of $K$ or of any field that includes $K$. Note however that, for a matrix $B$ in $M$, while $f(B) \in M$, we have

$$
\operatorname{det}(B \mathrm{I}-A) \in K
$$

Since $A \mathrm{I}-A$ is the zero matrix, we have $\operatorname{det}(A I-A)=0$. This observation is not enough to ensure that $f(A)$ is the zero matrix. Nonetheless, we shall show that it is, in two ways.

Theorem 15 (Cayley-Hamilton). Over any field, every matrix is a zero of its characteristic polynomial.

First proof. By Theorem 10, with $f$ as in (2.8) we have

$$
\begin{equation*}
f(x) \cdot \mathrm{I}=(x \cdot \mathrm{I}-A) \operatorname{adj}(x \cdot \mathrm{I}-A) . \tag{2.9}
\end{equation*}
$$

Moreover,

$$
f(x) \cdot \mathrm{I}=\sum_{j=0}^{n} x^{j} \cdot F_{j}
$$

where, in the notation of (2.7),

$$
F=f_{j} \cdot \mathrm{I}
$$

Likewise, for some matrices $B_{j}$ in $M$,

$$
\operatorname{adj}(x \cdot \mathrm{I}-A)=\sum_{j=0}^{n-1} x^{j} \cdot B_{j} .
$$

Thus (2.9) becomes

$$
\begin{equation*}
\sum_{j=0}^{n} x^{j} \cdot f_{j} \cdot \mathrm{I}=(x \cdot \mathrm{I}-A) \sum_{j=0}^{n-1} x^{j} \cdot B_{j} \tag{2.10}
\end{equation*}
$$

This then will be true when $x$ is replaced by an element of $M$ that commutes with $A$. Since $A$ is such an element, and the right member of (2.10) becomes 0 when $x$ is replaced with $A$, the same is true for the left member; but this just means $f(A)=0$.

Second proof. Letting $f$ be the characteristic polynomial of $A$ in $M$ as in (2.8), we want to show $f(A)=0$. Since the determinant function is multiplicative, for every $P$ in $M^{\times}$,

$$
\begin{aligned}
\operatorname{det}(x \cdot \mathrm{I}-A) & =\operatorname{det}\left(P^{-1} \cdot(x \cdot \mathrm{I}-A) \cdot P\right) \\
& =\operatorname{det}\left(x \cdot \mathrm{I}-P^{-1} A P\right)
\end{aligned}
$$

By Theorem 14, for some matrix $P, P^{-1} A P$ is a triangular matrix. It does not matter that entries of $P$ may come from the algebraic closure of $K$, possibly not $K$ itself. We may assume $A$ is triangular. The characteristic polynomial of $A$ is now

$$
\prod_{i<n}\left(x-a_{i}^{i}\right) .
$$

Since the product is independent of the order of the factors, so is the product $\prod_{i<n}\left(A-a_{i}^{i} \cdot \mathrm{I}\right)$. We have to show that this product is 0 . Column $j$ of the product is

$$
\prod_{i<n}\left(A-a_{i}^{i} \cdot \mathrm{I}\right) \mathbf{e}_{j} .
$$

However, by (2.5),

$$
\begin{equation*}
\left(A-a_{j}^{j} \cdot \mathrm{I}\right) \mathbf{e}_{j}=A \mathbf{e}_{j}-a_{j}^{j} \mathbf{e}_{j}=\sum_{i<j} a_{i}^{j} \mathbf{e}_{i} \tag{2.11}
\end{equation*}
$$

and in particular

$$
\left(A-a_{j}^{j} \cdot \mathrm{I}\right) \mathbf{e}_{j} \in \operatorname{span}\left\{\mathbf{e}_{i}: i<j\right\}
$$

By induction then,

$$
\prod_{i \leqslant j}\left(A-a_{i}^{i} \cdot \mathrm{I}\right) \mathbf{e}_{j}=\mathbf{0}
$$

Finally

$$
\prod_{i \in n}\left(A-a_{i}^{i} \cdot \mathrm{I}\right) \mathbf{e}_{j}=\prod_{j<i<n}\left(A-a_{i}^{i} \cdot \mathrm{I}\right) \prod_{i \leqslant j}\left(A-a_{i}^{i} \cdot \mathrm{I}\right) \mathbf{e}_{j}=\mathbf{0}
$$

### 2.4 Minimal polynomial

Theorem 16. For any $A$ in $M$, the subset

$$
\{f \in K[x]: f(A)=0\}
$$

of $K[x]$ is a nonzero ideal.
Proof. The set is easily an ideal. It is nontrivial for containing the characteristic polynomial of $A$; alternatively, since $\operatorname{dim} M=n^{2}$, there must be some coefficients $f_{i}$, not all 0 , for which

$$
f_{0}+f_{1} \cdot A+\cdots+f_{n^{2}} \cdot A^{n^{2}}=0
$$

Since $K[x]$ is a principal-ideal domain, the ideal of the theorem has a monic generator, called the minimal polynomial of $A$. This polynomial therefore is a factor of the characteristic polynomial of $A$. In particular, every zero in $K$ of the minimal polynomial is a zero of the characteristic polynomial.

Theorem 17. In a field, every zero of the characteristic polynomial of a square matrix over the field is a zero of the minimal polynomial. Hence every irreducible factor of the characteristic polynomial is a factor of the minimal polynomial.

Proof. A zero of the characteristic polynomial of $A$ is just an eigenvalue of $A$. Let $\lambda$ be an eigenvalue, with corresponding eigenvector $\boldsymbol{b}$. Thus

$$
\begin{aligned}
A \boldsymbol{b} & =\lambda \boldsymbol{b}, \\
A^{j} \boldsymbol{b} & =\lambda^{j} \boldsymbol{b}, \\
f(A) \boldsymbol{b} & =f(\lambda) \boldsymbol{b}
\end{aligned}
$$

for all $f(x)$ in $K[x]$. In particular,

$$
f(A)=0 \Longrightarrow f(\lambda)=0 .
$$

If $f$ is the minimal polynomial of $A$, then $f(A)=0$, so $f(\lambda)=$ 0.

Theorem 18. A square matrix is diagonalizable if and only if its minimal polynomial is the product of distinct linear factors.

Proof. Suppose $A$ in $M$ is diagonalizable, so that, for some $B$ in $M^{\times}$, for some $\lambda_{j}$ in $K$,

$$
A B=B \operatorname{diag}\left(\lambda_{j}: j \in n\right) .
$$

Letting column $j$ of $B$ be $\boldsymbol{b}_{j}$, we have

$$
\begin{gathered}
A \boldsymbol{b}_{j}=\lambda_{j} \boldsymbol{b}_{j} \\
\left(A-\lambda_{j} \cdot \mathrm{I}\right) \boldsymbol{b}_{j}=\mathbf{0}
\end{gathered}
$$

Letting $m=\left|\left\{\lambda_{j}: j \in n\right\}\right|$, we may assume

$$
\left\{\lambda_{j}: j \in n\right\}=\left\{\lambda_{i}: i \in m\right\} .
$$

For all $j$ in $n$, we have

$$
\prod_{i \in m}\left(A-\lambda_{i} \cdot \mathrm{I}\right) \boldsymbol{b}_{j}=\mathbf{0}
$$

The $\boldsymbol{b}_{j}$ being linearly independent, letting

$$
\begin{equation*}
f(x)=\prod_{i \in m}\left(x-\lambda_{i}\right) \tag{2.12}
\end{equation*}
$$

we conclude $f(A)=0$, so the minimal polynomial of $A$ is a factor of $f(x)$. (It is the same as $f(x)$, since the $\lambda_{i}$ are eigenvectors of $A$, and each of these must be a zero of the minimal polynomial, by Theorem 17.)

Suppose conversely $f(x)$ as given by (2.12), where again the $\lambda_{i}$ are all distinct, is the minimal polynomial of $A$. In particular, $f(A)=0$. If $j \in m$, we can define $g_{j}(x)$ in $K[x]$ by

$$
\begin{equation*}
\left(x-\lambda_{j}\right) g_{j}(x)=f(x) \tag{2.13}
\end{equation*}
$$

The $\lambda_{j}$ being distinct, the greatest common divisor of the $g_{j}(x)$ in $K[x]$ is unity. Since $K[x]$ is a Euclidean domain, by Bézout's Lemma there are $q_{j}(x)$ in $K[x]$ such that

$$
\sum_{j \in m} g_{j}(x) q_{j}(x)=1
$$

Then

$$
\sum_{j \in m} g_{j}(A) q_{j}(A)=\mathrm{I}
$$

Thus for every $\boldsymbol{v}$ in $K^{n}$, when we define

$$
\begin{equation*}
g_{j}(A) q_{j}(A) \boldsymbol{v}=\boldsymbol{w}_{j} \tag{2.14}
\end{equation*}
$$

we have

$$
\sum_{j \in m} \boldsymbol{w}_{j}=\boldsymbol{v}
$$

But then since $f(A)=0$, from (2.13) and (2.14) we have

$$
\mathbf{0}=f(A) q_{j}(A) \boldsymbol{v}=\left(A-\lambda_{j}\right) \boldsymbol{w}_{j}
$$

so that $\boldsymbol{w}_{j}$ belongs to the eigenspace associated with $\lambda_{j}$. In particular, by Theorem 12, there must be $n$ linearly independent eigenvectors, so $A$ is diagonalizable.

## 3 Jordan Normal Form

The presentation here is based mainly on Lang [3].

### 3.1 Cyclic spaces

Supposing $\lambda$ is an eigenvalue of the $n \times n$ matrix $A$, we let

$$
\begin{equation*}
B_{\lambda}=A-\lambda \cdot \mathrm{I} . \tag{3.1}
\end{equation*}
$$

If $\boldsymbol{v}_{0}$ is a corresponding eigenvector, this means

$$
\boldsymbol{v}_{0} \neq \mathbf{0}, \quad B_{\lambda} \boldsymbol{v}_{0}=\mathbf{0} .
$$

If possible now, let $B_{\lambda} \boldsymbol{v}_{1}=\boldsymbol{v}_{0}$. Then

$$
A \boldsymbol{v}_{1}=\lambda \boldsymbol{v}_{1}+\boldsymbol{v}_{0}, \quad B_{\lambda}^{2} \boldsymbol{v}_{1}=\mathbf{0}
$$

Suppose, in this way, for some $s$, when $0<k<s$,

$$
A \boldsymbol{v}_{k}=\lambda \boldsymbol{v}_{k}+\boldsymbol{v}_{k-1}, \quad B_{\lambda}{ }^{k+1} \boldsymbol{v}_{k}=\mathbf{0} .
$$

Then defining $P$ as the $n \times s$ matrix

$$
\left(\boldsymbol{v}_{0}|\cdots| \boldsymbol{v}_{s-1}\right)
$$

we have

$$
\begin{align*}
A P= & \left(A \boldsymbol{v}_{0}|\cdots| A \boldsymbol{v}_{s-1}\right) \\
& =\left(\lambda \boldsymbol{v}_{0}\left|\boldsymbol{v}_{0}+\lambda \boldsymbol{v}_{1}\right| \cdots \mid \boldsymbol{v}_{s-2}+\lambda \boldsymbol{v}_{s-1}\right)=P J \tag{3.2}
\end{align*}
$$

where $J$ is the $s \times s$ matrix

$$
\left(\begin{array}{ccccc}
\lambda & 1 & 0 & \ldots & 0 \\
0 & \lambda & 1 & \ddots & \vdots \\
0 & 0 & \ddots & \ddots & 0 \\
\vdots & & \ddots & \lambda & 1 \\
0 & \ldots & \ldots & 0 & \lambda
\end{array}\right) .
$$

Theorem 19. The columns of the matrix $P$ just defined are linearly independent.

Proof. Writing $\boldsymbol{v}$ for $\boldsymbol{v}_{s-1}$ and $B$ for $B_{\lambda}$, we have

$$
P=\left(B^{s-1} \boldsymbol{v}|\cdots| B^{0} \boldsymbol{v} .\right) .
$$

We show the columns are linearly independent. Suppose for some scalars $c^{i}$,

$$
\sum_{i<s} c^{i} \cdot B^{s-i} \boldsymbol{v}=\mathbf{0}
$$

Then $f(B) \boldsymbol{v}=\mathbf{0}$, where

$$
f(x)=\sum_{i<s} c^{i} x^{s-i}
$$

However, also $g(B) \boldsymbol{v}=\mathbf{0}$, where

$$
g(x)=x^{s}
$$

Letting $h$ be the greatest common factor of $f$ and $g$, we have

$$
h(B) \boldsymbol{v}=\mathbf{0}
$$

Also, $h(x)=x^{r}$ for some $r$, where $r \leqslant s$. When $r<s$, we have

$$
B^{r} \boldsymbol{v}=\boldsymbol{v}_{s-1-r}
$$

which is not 0 . Thus $h(x)=x^{s}$, and therefore $f=0$.

In the proof, $\operatorname{span}\left\{\boldsymbol{v}_{k}: k \in s\right\}$ is a $B$-cyclic subspace of $K^{n}$, because it is, for some one vector $\boldsymbol{v}$, spanned by the vectors $B^{k} \boldsymbol{v}$. The space is then $B$-invariant, because closed under multiplication by $B$.

### 3.2 Direct sums

Suppose $V$ is a vector space over $K$, and for some $m$ in $\mathbb{N}$, and for each $j$ in $n, V_{j}$ is a subspace of $V$. If the homomorphism

$$
\left(v_{i}: i<n\right) \mapsto \sum_{i<n} v_{i}
$$

from $\prod_{i<n} V_{i}$ to $V$ is surjective, then $V$ is the sum of the subspaces $V_{i}$, and we may write

$$
V=V_{0}+\cdots+V_{n-1}=\sum_{i<n} V_{i}
$$

If, further, the homomorphism is injective, then $V$ is the direct sum of the $V_{i}$, and we may write

$$
V=V_{0} \oplus \cdots \oplus V_{n-1}=\bigoplus_{i<n} V_{i}
$$

Given $B$ in $M$, we shall understand

$$
\operatorname{ker} B=\left\{\boldsymbol{x} \in K^{n}: B \boldsymbol{x}=\mathbf{0}\right\}
$$

Lemma 1. If $f$ and $g$ in $K[x]$ are co-prime, then for all $A$ in M,

$$
\operatorname{ker}(f(A) g(A))=\operatorname{ker} f(A) \oplus \operatorname{ker} g(A)
$$

Proof. By Bézout's Lemma for some $q$ and $r$ in $K[x]$,

$$
\begin{gathered}
q f+r g=1 \\
q(A) f(A)+r(A) g(A)=\mathrm{I}
\end{gathered}
$$

For all $\boldsymbol{v}$ in $K^{n}$ then,

$$
q(A) f(A) \boldsymbol{v}+r(A) g(A) \boldsymbol{v}=\boldsymbol{v}
$$

Suppose now

$$
\boldsymbol{w} \in \operatorname{ker}(f(A) g(A))
$$

Then

$$
r(A) g(A) \boldsymbol{w} \in \operatorname{ker} f(A), \quad q(A) f(A) \boldsymbol{w} \in \operatorname{ker} g(A)
$$

and so

$$
\boldsymbol{w} \in \operatorname{ker} f(A)+\operatorname{ker} g(A)
$$

Conversely, suppose

$$
\boldsymbol{u} \in \operatorname{ker} f(A), \quad \boldsymbol{v} \in \operatorname{ker} g(A)
$$

Then

$$
\begin{aligned}
\boldsymbol{u}=q(A) f(A) \boldsymbol{u}+r(A) g & (A) \boldsymbol{u} \\
& =r(A) g(A) \boldsymbol{u}=r(A) g(A)(\boldsymbol{u}+\boldsymbol{v})
\end{aligned}
$$

and likewise

$$
\boldsymbol{v}=q(A) f(A)(\boldsymbol{u}+\boldsymbol{v})
$$

This shows $(\boldsymbol{u}, \boldsymbol{v}) \mapsto \boldsymbol{u}+\boldsymbol{v}$ is injective.
Theorem 20. If each of some $f$ in $K[x]$ is prime to the others, then for all $A$ in $M$,

$$
\operatorname{ker} \prod_{f} f(A)=\bigoplus_{f} \operatorname{ker} f(A)
$$

### 3.3 Kernels

Suppose $A$ in $M$ has characteristic polynomial $f$, and $K$ is algebraically closed. Then

$$
f=\prod_{j<m}\left(x-\lambda_{j}\right)^{r_{j}}
$$

for some $\lambda_{j}$ in $K$ and $r_{j}$ in $\mathbb{N}$. By the Cayley-Hamilton Theorem,

$$
\operatorname{ker}(f(A))=K^{n} .
$$

Letting

$$
B_{j}=A-\lambda_{j} \cdot \mathrm{I},
$$

we have now, by Theorem 20,

$$
K^{n}=\bigoplus_{j<m} \operatorname{ker}\left(B_{j}{ }^{r_{j}}\right) .
$$

Theorem 21. For all $B$ in $M$, for all $s$ in $\mathbb{N}, \operatorname{ker}\left(B^{s}\right)$ is the direct sum of $B$-cyclic subspaces.

Proof. We shall prove the claim for every $B$-invariant subspace of $\operatorname{ker}\left(B^{s}\right)$. We use induction on the dimension of the subspace. If the dimension is 0 , the claim is vacuously true. Suppose $V$ is a $B$-invariant subspace of $\operatorname{ker}\left(B^{s}\right)$ having positive dimension. Then

$$
V \nsubseteq \operatorname{ker}\left(B^{0}\right), \quad \operatorname{ker}\left(B^{0}\right) \subseteq \ldots \subseteq \operatorname{ker}\left(B^{s}\right), \quad V \subseteq \operatorname{ker}\left(B^{s}\right),
$$

so for some $r$,

$$
V \nsubseteq \operatorname{ker}\left(B^{r-1}\right), \quad V \subseteq \operatorname{ker}\left(B^{r}\right)
$$

Then

$$
V B \subseteq V \cap \operatorname{ker}\left(B^{r-1}\right) \subset V
$$

This shows

$$
V B \subset V .
$$

As an inductive hypothesis, we assume

$$
V B=\bigoplus_{i<m} W_{i},
$$

where each $W_{i}$ is $B$-cyclic. Then for some $\boldsymbol{w}_{i}$ in $V$, for some $r_{i}$ in $\mathbb{N}$,

$$
\begin{equation*}
W_{i}=\operatorname{span}\left\{B^{j} \boldsymbol{w}_{i}: j<r_{i}\right\}, \quad \mathbf{0}=B^{r_{i}} \boldsymbol{w}_{i} . \tag{3.5}
\end{equation*}
$$

For some $\boldsymbol{v}_{i}$ in $V$,

$$
\begin{equation*}
\boldsymbol{w}_{i}=B \boldsymbol{v}_{i} . \tag{3.6}
\end{equation*}
$$

Now let

$$
V_{i}=\operatorname{span}\left\{B^{j} \boldsymbol{v}_{i}: i \leqslant r_{i}\right\} .
$$

Then $V_{i}$ is a $B$-cyclic space, since $B^{r_{i}+1} \boldsymbol{v}_{i}=\mathbf{0}$. We shall show that the sum of the $V_{i}$ is direct. An arbitrary element of $V_{i}$ is $f_{i}(B) \boldsymbol{v}_{i}$ for some $f_{i}$ in $K[x]$ such that

$$
\begin{equation*}
\operatorname{deg} f_{i} \leqslant r_{i} \tag{3.7}
\end{equation*}
$$

Suppose

$$
\mathbf{0}=\sum_{i<m} f_{i}(B) \boldsymbol{v}_{i} .
$$

Then by (3.6),

$$
\begin{equation*}
\mathbf{0}=\sum_{i<m} f_{i}(B) \boldsymbol{w}_{i} . \tag{3.8}
\end{equation*}
$$

But then by (3.4),

$$
\mathbf{0}=f_{i}(B) \boldsymbol{w}_{i},
$$

so by $(3 \cdot 7)$, and (3.5), and Theorem 19 ,

$$
f_{i}=c_{i} x^{r_{i}}
$$

for some $c_{i}$ in $K$. In this case, we can write (3.8) as

$$
\mathbf{0}=\sum_{i<m} c_{i} B^{r_{i}-1} \boldsymbol{w}_{i}
$$

which implies that each $c_{i}$ is 0 . Thus $f_{i}=0$.
Now we can let

$$
V^{\prime}=\bigoplus_{i<m} V_{i}
$$

Then $V^{\prime} \subseteq V$. By construction, $V_{i} B=W_{i}$, so

$$
V^{\prime} B=W=V B
$$

Therefore

$$
V=V^{\prime}+\operatorname{ker} B
$$

Each element of ker $B$ constitutes a basis of a one-dimensional $B$-cyclic space. Then $V$ is the direct sum of some of these spaces, along with the $V_{i}$, as desired.

In the notation of (3.3), there are $n_{j}$ in $\mathbb{N}$, and then there are $\boldsymbol{v}_{j k}$ in $\operatorname{ker}\left(B_{j}{ }^{r_{j}}\right)$ and $s_{j k}$ in $\mathbb{N}$ such that

$$
B_{j}^{s_{j k}-1} \boldsymbol{v}_{j k} \neq \mathbf{0}, \quad B_{j}^{s_{j k}} \boldsymbol{v}_{j k}=\mathbf{0}
$$

and

$$
\operatorname{ker}\left(B_{j}{ }^{r_{j}}\right)=\bigoplus_{k<n_{j}} \operatorname{span}\left\{B_{j}{ }^{i} \boldsymbol{v}_{j k}: i<s_{j k}\right\}
$$

Now we may let

$$
P=\left(P_{0}|\cdots| P_{m-1}\right)
$$

where, for each $j$ in $m$,

$$
P_{j}=\left(P_{j 0}|\cdots| P_{j, n_{j}-1}\right),
$$

where, for each $k$ in $n_{j}$,

$$
P_{j k}=\left(B_{j}^{s_{j k}-1} \boldsymbol{v}_{j k}|\cdots| \boldsymbol{v}_{j, k}\right) .
$$

Then $P A P^{-1}$ is a Jordan normal form for $A$. Indeed, by the considerations yielding (3.2),

$$
P A P^{-1}=\operatorname{diag}\left(\Lambda_{0}, \ldots, \Lambda_{m-1}\right)
$$

where, for each $j$ in $m$,

$$
\Lambda_{j}=\operatorname{diag}\left(\Lambda_{j 0}, \ldots, \Lambda_{j, n_{j}-1}\right)
$$

where, for each $k$ in $n_{j}, \Lambda_{j, k}$ is the $s_{j k} \times s_{j k}$ matrix

$$
\left(\begin{array}{ccccc}
\lambda_{j} & 1 & 0 & \ldots & 0 \\
0 & \lambda_{j} & 1 & \ddots & \vdots \\
0 & 0 & \ddots & \ddots & 0 \\
\vdots & & \ddots & \lambda_{j} & 1 \\
0 & \ldots \ldots & 0 & \lambda_{j}
\end{array}\right)
$$

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