## Chapter 2

## Matrices and Linear Algebra

### 2.1 Basics

Definition 2.1.1. A matrix is an $m \times n$ array of scalars from a given field $F$. The individual values in the matrix are called entries.

Examples.

$$
A=\left[\begin{array}{ccc}
2 & 1 & 3 \\
-1 & 2 & 4
\end{array}\right] \quad B=\left[\begin{array}{ll}
1 & 2 \\
3 & 4
\end{array}\right]
$$

The size of the array is-written as $m \times n$, where

$$
\begin{gathered}
m \times n \\
\nearrow \\
\text { number of rows } \\
\text { number of columns }
\end{gathered}
$$

## Notation

$$
A=\left[\begin{array}{cccc}
a_{11} & a_{12} & \ldots & a_{1 n} \\
a_{21} & a_{22} & \ldots & a_{2 n} \\
a_{n 1} & a_{n 2} & \ldots & a_{m n}
\end{array}\right] \begin{gathered}
\nwarrow \\
\nwarrow \\
\nwarrow \\
\text { columns } \\
\nearrow
\end{gathered}
$$

$A:=$ uppercase denotes a matrix
$a:=$ lower case denotes an entry of a matrix $a \in F$.
Special matrices
(1) If $m=n$, the matrix is called square. In this case we have
(1a) A matrix $A$ is said to be diagonal if

$$
a_{i j}=0 \quad i \neq j .
$$

(1b) A diagonal matrix $A$ may be denoted by $\operatorname{diag}\left(d_{1}, d_{2}, \ldots, d_{n}\right)$ where

$$
a_{i i}=d_{i} \quad a_{i j}=0 \quad j \neq i .
$$

The diagonal matrix $\operatorname{diag}(1,1, \ldots, 1)$ is called the identity matrix and is usually denoted by

$$
I_{n}=\left[\begin{array}{cccc}
1 & 0 & \ldots & 0 \\
0 & 1 & & \\
\vdots & & \ddots & \\
0 & & & 1
\end{array}\right]
$$

or simply $I$, when $n$ is assumed to be known. $0=\operatorname{diag}(0, \ldots, 0)$ is called the zero matrix.
(1c) A square matrix $L$ is said to be lower triangular if

$$
\ell_{i j}=0 \quad i<j .
$$

(1d) A square matrix $U$ is said to be upper triangular if

$$
u_{i j}=0 \quad i>j .
$$

(1e) A square matrix $A$ is called symmetric if

$$
a_{i j}=a_{j i} .
$$

(1f) A square matrix $A$ is called Hermitian if

$$
a_{i j}=\bar{a}_{j i} \quad(\bar{z}:=\text { complex conjugate of } z) .
$$

(1g) $E_{i j}$ has a 1 in the $(i, j)$ position and zeros in all other positions.
(2) A rectangular matrix $A$ is called nonnegative if

$$
a_{i j} \geq 0 \quad \text { all } \quad i, j .
$$

It is called positive if

$$
a_{i j}>0 \quad \text { all } \quad i, j .
$$

Each of these matrices has some special properties, which we will study during this course.

Definition 2.1.2. The set of all $m \times n$ matrices is denoted by $M_{m, n}(F)$, where $F$ is the underlying field (usually $R$ or $\mathbb{C}$ ). In the case where $m=n$ we write $M_{n}(F)$ to denote the matrices of size $n \times n$.

Theorem 2.1.1. $M_{m, n}$ is a vector space with basis given by $E_{i j}, 1 \leq i \leq$ $m, \quad 1 \leq j \leq n$.

## Equality, Addition, Multiplication

Definition 2.1.3. Two matrices $A$ and $B$ are equal if and only if they have the same size and

$$
a_{i j}=b_{i j} \quad \text { all } \quad i, j .
$$

Definition 2.1.4. If $A$ is any matrix and $\alpha \in F$ then the scalar multiplication $B=\alpha A$ is defined by

$$
b_{i j}=\alpha a_{i j} \quad \text { all } \quad i, j .
$$

Definition 2.1.5. If $A$ and $B$ are matrices of the same size then the sum $A$ and $B$ is defined by $C=A+B$, where

$$
c_{i j}=a_{i j}+b_{i j} \quad \text { all } \quad i, j
$$

We can also compute the difference $D=A-B$ by summing $A$ and $(-1) B$

$$
D=A-B=A+(-1) B .
$$

matrix subtraction.
Matrix addition "inherits" many properties from the field $F$.
Theorem 2.1.2. If $A, B, C \in M_{m, n}(F)$ and $\alpha, \beta \in F$, then
(1) $A+B=B+A$ commutivity
(2) $A+(B+C)=(A+B)+C \quad$ associativity
(3) $\alpha(A+B)=\alpha A+\alpha B \quad$ distributivity of a scalar
(4) If $B=0$ (a matrix of all zeros) then

$$
A+B=A+0=A
$$

(4) $(\alpha+\beta) A=\alpha A+\beta A$
(5) $\alpha(\beta A)=\alpha \beta A$
(6) $0 A=0$
(7) $\alpha 0=0$.

Definition 2.1.6. If $x$ and $y \in R_{n}$,

$$
\begin{aligned}
x & =\left(x_{1} \ldots x_{n}\right) \\
y & =\left(y_{1} \ldots y_{n}\right) .
\end{aligned}
$$

Then the scalar or dot product of $x$ and $y$ is given by

$$
\langle x, y\rangle=\sum_{i=1}^{n} x_{i} y_{i}
$$

Remark 2.1.1. (i) Alternate notation for the scalar product: $\langle x, y\rangle=x \cdot y$. (ii) The dot product is defined only for vectors of the same length.

Example 2.1.1. Let $x=(1,0,3,-1)$ and $y=(0,2,-1,2)$ then $\langle x, y\rangle=$ $1(0)+0(2)+3(-1)-1(2)=-5$.
Definition 2.1.7. If $A$ is $m \times n$ and $B$ is $n \times p$. Let $r_{i}(A)$ denote the vector with entries given by the $i^{\text {th }}$ row of $A$, and let $c_{j}(B)$ denote the vector with entries given by the $j^{\text {th }}$ row of $B$. The product $C=A B$ is the $m \times p$ matrix defined by

$$
c_{i j}=\left\langle r_{i}(A), c_{j}(B)\right\rangle
$$

where $r_{i}(A)$ is the vector in $R_{n}$ consisting of the $i^{\text {th }}$ row of $A$ and similarly $c_{j}(B)$ is the vector formed from the $j^{\text {th }}$ column of $B$. Other notation for $C=A B$

$$
\begin{array}{ll}
c_{i j}=\sum_{k=1}^{n} a_{i k} b_{k j} & 1 \leq i \leq m \\
& 1 \leq j \leq p .
\end{array}
$$

Example 2.1.2. Let

$$
A=\left[\begin{array}{lll}
1 & 0 & 1 \\
3 & 2 & 1
\end{array}\right] \quad \text { and } \quad B=\left[\begin{array}{cc}
2 & 1 \\
3 & 0 \\
-1 & 1
\end{array}\right]
$$

Then

$$
A B=\left[\begin{array}{cc}
1 & 2 \\
11 & 4
\end{array}\right]
$$

## Properties of matrix multiplication

(1) If $A B$ exists, does it happen that $B A$ exists and $A B=B A$ ? The answer is usually no. First $A B$ and $B A$ exist if and only if $A \in$ $M_{m, n}(F)$ and $B \in M_{n, m}(F)$. Even if this is so the sizes of $A B$ and $B A$ are different ( $A B$ is $m \times m$ and $B A$ is $n \times n$ ) unless $m=n$. However even if $m=n$ we may have $A B \neq B A$. See the examples below. They may be different sizes and if they are the same size (i.e. $A$ and $B$ are square) the entries may be different

$$
\left.\begin{array}{rlrl}
A=[1,2] & B=\left[\begin{array}{c}
-1 \\
1
\end{array}\right] & A B & =[1] \\
A=\left[\begin{array}{ll}
1 & 2 \\
3 & 4
\end{array}\right] B=\left[\begin{array}{cc}
-1 & 1 \\
0 & 1
\end{array}\right] & A B & =\left[\begin{array}{cc}
-1 & -2 \\
1 & 2
\end{array}\right] \\
-1 & 3 \\
-3 & 7
\end{array}\right] \quad B A=\left[\begin{array}{cc}
2 & 2 \\
3 & 4
\end{array}\right]
$$

(2) If $A$ is square we define

$$
\begin{aligned}
& A^{1}=A, \quad A^{2}=A A, \quad A^{3}=A^{2} A=A A A \\
& A^{n}=A^{n-1} A=A \cdots A \quad(n \text { factors }) .
\end{aligned}
$$

(3) $I=\operatorname{diag}(1, \ldots, 1)$. If $A \in M_{m, n}(F)$ then

$$
\begin{aligned}
A I_{n} & =A \quad \text { and } \\
I_{m} A & =A .
\end{aligned}
$$

Theorem 2.1.3 (Matrix Multiplication Rules). Assume $A, B$, and $C$ are matrices for which all products below make sense. Then
(1) $A(B C)=(A B) C$
(2) $A(B \pm C)=A B \pm A C$ and $(A \pm B) C=A C \pm B C$
(3) $A I=A$ and $I A=A$
(4) $c(A B)=(c A) B$
(5) $A 0=0$ and $0 B=0$
(6) For A square

$$
A^{r} A^{s}=A^{s} A^{r} \quad \text { for all integers } \quad r, s \geq 1
$$

Fact: If $A C$ and $B C$ are equal, it does not follow that $A=B$. See Exercise 60.

Remark 2.1.2. We use an alternate notation for matrix entries. For any matrix $B$ denote the $(i, j)$-entry by $(B)_{i j}$.

Definition 2.1.8. Let $A \in M_{m, n}(F)$.
(i) Define the transpose of $A$, denoted by $A^{T}$, to be the $n \times m$ matrix with entries

$$
\left(A^{T}\right)_{i j}=a_{j i} .
$$

(ii) Define the adjoint of $A$, denoted by $A^{*}$, to be the $n \times m$ matrix with entries

$$
\left(A^{*}\right)_{i j}=\bar{a}_{j i} \quad \text { complex conjugate }
$$

## Example 2.1.3.

$$
A=\left[\begin{array}{lll}
1 & 2 & 3 \\
5 & 4 & 1
\end{array}\right] \quad A^{T}=\left[\begin{array}{ll}
1 & 5 \\
2 & 4 \\
3 & 1
\end{array}\right]
$$

In words... "The rows of $A$ become the columns of $A^{T}$, taken in the same order." The following results are easy to prove.

Theorem 2.1.4 (Laws of transposes). (1) $\left(A^{T}\right)^{T}=A$ and $\left(A^{*}\right)^{*}=A$
(2) $\begin{array}{ll}(A \pm B)^{T}=A^{T} \pm B^{T} \quad \text { (and for *) }\end{array}$
(3) $(c A)^{T}=c A^{T} \quad(c A)^{*}=\bar{c} A^{*}$
(4) $(A B)^{T}=B^{T} A^{T}$
(5) If $A$ is symmetric

$$
A=A^{T}
$$

(6) If $A$ is Hermitian

$$
A=A^{*} .
$$

More facts about symmetry.
Proof. (1) We know $\left(A^{T}\right)_{i j}=a_{j i}$. So $\left(\left(A^{T}\right)^{T}\right)_{i j}=a_{i j}$. Thus $\left(A^{T}\right)^{T}=A$.
(2) $(A \pm B)^{T}=a_{j i} \pm b_{j i}$. So $(A \pm B)^{T}=A^{T} \pm B^{T}$.

Proposition 2.1.1. (1) $A$ is symmetric if and only if $A^{T}$ is symmetric.
(1)* $A$ is Hermitian if and only if $A^{*}$ is Hermitian.
(2) If $A$ is symmetric, then $A^{2}$ is also symmetric.
(3) If $A$ is symmetric, then $A^{n}$ is also symmetric for all $n$.

Definition 2.1.9. A matrix is called skew-symmetric if

$$
A^{T}=-A
$$

Example 2.1.4. The matrix

$$
A=\left[\begin{array}{ccc}
0 & 1 & 2 \\
-1 & 0 & -3 \\
-2 & 3 & 0
\end{array}\right]
$$

is skew-symmetric.
Theorem 2.1.5. (1) If $A$ is skew symmetric, then $A$ is a square matrix and $a_{i i}=0, i=1, \ldots, n$.
(2) For any matrix $A \in M_{n}(F)$

$$
A-A^{T}
$$

is skew-symmetric while $A+A^{T}$ is symmetric.
(3) Every matrix $A \in M_{n}(F)$ can be uniquely written as the sum of $a$ skew-symmetric and symmetric matrix.
Proof. (1) If $A \in M_{m, n}(F)$, then $A^{T} \in M_{n, m}(F)$. So, if $A^{T}=-A$ we must have $m=n$. Also

$$
a_{i i}=-a_{i i}
$$

for $i=1, \ldots, n$. So $a_{i i}=0$ for all $i$.
(2) Since $\left(A-A^{T}\right)^{T}=A^{T}-A=-\left(A-A^{T}\right)$, it follows that $A-A^{T}$ is skew-symmetric.
(3) Let $A=B+C$ be a second such decomposition. Subtraction gives

$$
\frac{1}{2}\left(A+A^{T}\right)-B=C-\frac{1}{2}\left(A-A^{T}\right)
$$

The left matrix is symmetric while the right matrix is skew-symmetric. Hence both are the zero matrix.

$$
A=\frac{1}{2}\left(A+A^{T}\right)+\frac{1}{2}\left(A-A^{T}\right) .
$$

Examples. $A=\left[\begin{array}{cc}0 & -1 \\ 1 & 0\end{array}\right]$ is skew-symmetric. Let

$$
\begin{aligned}
B & =\left[\begin{array}{cc}
1 & 2 \\
-1 & 4
\end{array}\right] \\
B^{T} & =\left[\begin{array}{cc}
1 & -1 \\
2 & 4
\end{array}\right] \\
B-B^{T} & =\left[\begin{array}{cc}
0 & 3 \\
-3 & 0
\end{array}\right] \\
B+B^{T} & =\left[\begin{array}{ll}
2 & 1 \\
1 & 8
\end{array}\right] .
\end{aligned}
$$

Then

$$
B=\frac{1}{2}\left(B-B^{T}\right)+\frac{1}{2}\left(B+B^{T}\right) .
$$

An important observation about matrix multiplication is related to ideas from vector spaces. Indeed, two very important vector spaces are associated with matrices.

Definition 2.1.10. Let $A \in M_{m, n}(\mathbb{C})$.
(i)Denote by

$$
c_{j}(A):=j^{\text {th }} \text { column of } A
$$

$c_{j}(A) \in \mathbb{C}_{m}$. We call the subspace of $\mathbb{C}_{m}$ spanned by the columns of $A$ the column space of $A$. With $c_{1}(A), \ldots, c_{n}(A)$ denoting the columns of $A$
the column space is $\mathfrak{S}\left(c_{1}(A), \ldots, c_{n}(A)\right)$.
(ii) Similarly, we call the subspace of $\mathbb{C}_{n}$ spanned by the rows of $A$ the row space of $A$. With $r_{1}(A), \ldots, r_{m}(A)$ denoting the rows of $A$ the row space is therefore $\mathfrak{S}\left(r_{1}(A), \ldots, r_{m}(A)\right)$.

Let $x \in \mathbb{C}_{n}$, which we view as the $n \times 1$ matrix $x=\left[x_{1} \ldots x_{n}\right]^{T}$. The product $A x$ is defined and

$$
A x=\sum_{j=1}^{n} x_{j} c_{j}(A)
$$

That is to say, $A x \in \mathfrak{S}\left(c_{1}(A), \ldots, c_{n}(A)\right)=$ column space of $A$.
Definition 2.1.11. Let $A \in M_{n}(F)$. The matrix $A$ is said to be invertible if there is a matrix $B \in M_{n}(F)$ such that

$$
A B=B A=I
$$

In this case $B$ is called the inverse of $A$, and the notation for the inverse is $A^{-1}$.

## Examples.

(i) Let

$$
\begin{aligned}
A & =\left[\begin{array}{cc}
1 & 3 \\
-1 & 2
\end{array}\right] \\
A^{-1} & =\frac{1}{5}\left[\begin{array}{cc}
2 & -3 \\
1 & 1
\end{array}\right] .
\end{aligned}
$$

(ii) For $n=3$ we have

$$
A=\left[\begin{array}{ccc}
1 & 2 & -1 \\
-1 & 3 & -1 \\
-2 & 3 & -1
\end{array}\right] \quad A^{-1}=\left[\begin{array}{ccc}
0 & 1 & -1 \\
-1 & 3 & -2 \\
-3 & 7 & -5
\end{array}\right]
$$

A square matrix need not have an inverse, as will be discussed in the next section. As examples, the two matrices below do not have inverses

$$
A=\left[\begin{array}{cc}
1 & -2 \\
-1 & 2
\end{array}\right] \quad B=\left[\begin{array}{lll}
1 & 0 & 1 \\
0 & 2 & 1 \\
1 & 2 & 2
\end{array}\right]
$$

### 2.2 Linear Systems

The solutions of linear systems is likely the single largest application of matrix theory. Indeed, most reasonable problems of the sciences and economics that have the need to solve problems of several variable almost without exception are reduced to component parts where one of them is the solution of a linear system. Of course the entire solution process may have the linear system solver as a relatively small component, but an essential one. Even the solution of nonlinear problems, especially, employ linear systems to great and crucial advantage.

To be precise, we suppose that the coefficients $a_{i j}, \quad 1 \leq i \leq m$ and $1 \leq$ $j \leq n$ and the data $b_{j}, 1 \leq j \leq m$ are known. We define the linear system for the $n$ unknowns $x_{1}, \ldots, x_{n}$ to be

$$
\begin{align*}
a_{11} x_{1}+a_{12} x_{2}+\cdots+a_{1 n} x_{n} & =b_{1} \\
a_{21} x_{1}+a_{22} x_{2}+\cdots+a_{2 n} x_{n} & =b_{2}  \tag{*}\\
a_{m 1} x_{1}+a_{m 2} x_{2}+\cdots+a_{m n} x_{n} & =b_{m}
\end{align*}
$$

The solution set is defined to be the subset of $R_{n}$ of vectors $\left(x_{1}, \ldots, x_{n}\right)$ that satisfy each of the $m$ equations of the system. The question of how to solve a linear system includes a vast literature of theoretical and computation methods. Certain systems form the model of what to do. In the systems below we note that the first one has three highly coupled (interrelated) variables.

$$
\begin{aligned}
3 x_{1}-2 x_{2}+4 x_{3} & =7 \\
x_{1}-6 x_{2}-2 x_{3} & =0 \\
-x_{1}+3 x_{2}+6 x_{3} & =-2
\end{aligned}
$$

The second system is more tractable because there appears even to the untrained eye a clear and direct method of solution.

$$
\begin{aligned}
3 x_{1}-2 x_{2}-x_{3} & =7 \\
x_{2}-2 x_{3} & =1 \\
2 x_{3} & =-2
\end{aligned}
$$

Indeed, we can see right off that $x_{3}=-1$. Substituting this value into the second equation we obtain $x_{2}=1-2=-1$. Substituting both $x_{2}$ and $x_{3}$ into the first equation, we obtain $2 x_{1}-2(-1)-(-1)=7$, gives $x_{1}=2$. The
solution set is the vector $(2,-1,-1)$. The virtue of the second system is that the unknowns can be determined one-by-one, back substituting those already found into the next equation until all unknowns are determined. So if we can convert the given system of the first kind to one of the second kind, we can determine the solution.

This procedure for solving linear systems is therefore the applications of operations to effect the gradual elimination of unknowns from the equations until a new system results that can be solved by direct means. The operations allowed in this process must have precisely one important property: They must not change the solution set by either adding to it or subtracting from it. There are exactly three such operations needed to reduce any set of linear equations so that it can be solved directly.
(E1) Interchange two equations.
(E2) Multiply any equation by a nonzero constant.
(E3) Add a multiple of one equation to another.
This can be summarized in the following theorem
Theorem 2.2.1. Given the linear system (*). The set of equation operations E1, E2, and E3 on the equations of $\left(^{*}\right)$ does not alter the solution set of the system (*).

We leave this result to the exercises. Our main intent is to convert these operations into corresponding operations for matrices. Before we do this we clarify which linear systems can have a soltution. First, the system can be converted to matrix form by setting $A$ equal to the $m \times n$ matrix of coefficients, $b$ equal to the $m \times 1$ vector of data, and $x$ equal to the $n \times 1$ vector of unknowns. Then the system (*) can be written as

$$
A x=b
$$

In this way we see that with $c_{i}(A)$ denoting the $i^{\text {th }}$ column of $A$, the system is expressible as

$$
x_{1} c_{1}(A)+\cdots+x_{n} c_{n}(A)=b
$$

From this equation it is clear that the system has a solution if and only if the vector $b$ is in $\mathfrak{S}\left(c_{1}(A), \cdots, c_{n}(A)\right)$. This is summarized in the following theorem.

Theorem 2.2.2. A necessary and sufficient condition that $A x=b$ has $a$ solution is that $b \in \mathfrak{S}\left(c_{1}(A) \ldots c_{n}(A)\right)$.

In the general matrix product $C=A B$, we note that the column space of $C \subset$ column space of $A$. In the following definition we regard the matrix $A$ as a function acting upon vectors in one vector space with range in another vector space. This is entirely similar to the domain-range idea of function theory.

Definition 2.2.1. The range of $A=\left\{A x \mid x \in R_{n}\left(\right.\right.$ or $\left.\left.\mathbb{C}_{n}\right)\right\}$.
It follows directly from our discussion above that the range of $A$ equals $\mathfrak{S}\left(c_{1}(A), \ldots, c_{n}(A)\right)$.
Row operations: To solve $A x=b$ we use a process called Gaussian elimination, which is based on row operations.

Type 1: Interchange two rows. (Notation: $R_{i} \longleftrightarrow R_{j}$ )
Type 2: Multiply a row by a nonzero constant. (Notation: $c R_{i} \rightarrow R_{i}$ )
Type 3: Add a multiple of one row to another row. (Notation: $c R_{i}+R_{j} \rightarrow$ $R_{j}$ )

Gaussian elimination is the process of reducing a matrix to its RREF using these row operations. Each of these operations is the respective analogue of the equation operations described above, and each can be realized by left matrix multiplication. We have the following.
Type 1


Notation: $R_{i} \leftrightarrow R_{j}$
Type 2

$$
E_{2}=\left[\begin{array}{ccccccc}
1 & & & \vdots & & & \\
& \ddots & & \vdots & & & \\
& & 1 & \vdots & & & \\
\ldots & \ldots & \ldots & c & \ldots & \ldots & \ldots \\
& & & \vdots & 1 & & \\
& & & \vdots & & \ddots & \\
& & & \vdots & & & 1
\end{array}\right] \quad \text { row } i
$$

Notation: $c R_{i}$
Type 3


Notation: $c R_{i}+R_{j}$, the abbreviated form of $c R_{i}+R_{j} \rightarrow R_{j}$

Example 2.2.1. The operations

$$
\left[\begin{array}{ccc}
2 & 1 & 0 \\
0 & 2 & 1 \\
-1 & 0 & 2
\end{array}\right] \underset{\rightarrow}{\longrightarrow} R_{1} \underset{\rightarrow}{\longleftrightarrow}\left[\begin{array}{ccc}
0 & 2 & 1 \\
2 & 1 & 0 \\
-1 & 0 & 2
\end{array}\right] \underset{\rightarrow}{4 R_{3}}\left[\begin{array}{ccc}
0 & 2 & 1 \\
2 & 1 & 0 \\
-4 & 0 & 8
\end{array}\right]
$$

can also be realized as

$$
\begin{aligned}
& R_{1} \longleftrightarrow R_{2}:\left[\begin{array}{lll}
0 & 1 & 0 \\
1 & 0 & 0 \\
0 & 0 & 1
\end{array}\right]\left[\begin{array}{ccc}
2 & 1 & 0 \\
0 & 2 & 1 \\
-1 & 0 & 2
\end{array}\right]=\left[\begin{array}{ccc}
0 & 2 & 1 \\
2 & 1 & 0 \\
-1 & 0 & 2
\end{array}\right] \\
& 4 R_{3}: \quad\left[\begin{array}{lll}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & 4
\end{array}\right]\left[\begin{array}{ccc}
0 & 2 & 1 \\
2 & 1 & 0 \\
-1 & 0 & 2
\end{array}\right]=\left[\begin{array}{ccc}
0 & 2 & 1 \\
2 & 1 & 0 \\
-4 & 0 & 8
\end{array}\right]
\end{aligned}
$$

The operations

$$
\left[\begin{array}{ccc}
2 & 1 & 0 \\
0 & 2 & 1 \\
-1 & 0 & 2
\end{array}\right] \begin{gathered}
-3 R_{1}+R_{2} \\
\rightarrow \\
2 R_{1}+R_{3}
\end{gathered}\left[\begin{array}{ccc}
2 & 1 & 0 \\
-6 & -1 & 1 \\
3 & 2 & 2
\end{array}\right]
$$

can be realized by the left matrix multiplications

$$
\left[\begin{array}{lll}
1 & 0 & 0 \\
0 & 1 & 0 \\
2 & 0 & 1
\end{array}\right]\left[\begin{array}{ccc}
1 & 0 & 0 \\
-3 & 1 & 0 \\
0 & 0 & 1
\end{array}\right]\left[\begin{array}{ccc}
2 & 1 & 0 \\
0 & 2 & 1 \\
-1 & 0 & 2
\end{array}\right]=\left[\begin{array}{ccc}
2 & 1 & 0 \\
-6 & -1 & 1 \\
3 & 2 & 2
\end{array}\right]
$$

Note there are two matrix multiplications them, one for each Type 3 elementary operation.

Row-reduced echelon form. To each $A \in M_{m, n}(E)$ there is a canonical form also in $M_{m, n}(E)$ which may be obtained by row operations. Called the RREF, it has the following properties.
(a) Each nonzero row has a 1 as the first nonzero entry (:= leading one).
(b) All column entries above and below a leading one are zero.
(c) All zero rows are at the bottom.
(d) The leading one of one row is to the left of leading ones of all lower rows.

## Example 2.2.2.

$$
B=\left[\begin{array}{ccccc}
1 & 2 & 0 & 0 & -1 \\
0 & 0 & 1 & 0 & 3 \\
0 & 0 & 0 & 1 & 0 \\
0 & 0 & 0 & 0 & 0
\end{array}\right] \quad \text { is in RREF. }
$$

Theorem 2.2.3. Let $A \in M_{m, n}(F)$. Then the RREF is necessarily unique.
We defer the proof of this result. Let $A \in M_{m, n}(F)$. Recall that the row space of $A$ is the subspace of $R_{n}$ (or $\left.\mathbb{C}_{n}\right)$ spanned by the rows of $A$. In symbols the row space is

$$
\mathfrak{S}\left(r_{1}(A), \ldots, r_{m}(A)\right)
$$

Proposition 2.2.1. For $A \in M_{m, n}(F)$ the rows of its $R R E F$ span the rows space of $A$.

Proof. First, we know the nonzero rows of the RREF are linearly independent. And all row operations are linear combinations of the rows. Therefore the row space generated from the RREF is contained in the row space of $A$. If the containment is proper. That is there is a row of $A$ that is linearly independent from the row space of the RREF, this is a contradiction because every row of $A$ can be obtained by the inverse row operations from the RREF.

Proposition 2.2.2. If $A \in M_{m, n}(F)$ and a row operation is applied to $A$, then linearly dependent columns of $A$ remain linearly dependent and linearly independent columns of $A$ remain linearly independent.

Proposition 2.2.3. The number of linearly independent columns of $A \in$ $M_{m, n}(F)$ is the same as the number of leading ones in the RREF of $A$.

Proof. Let $S=\left\{i_{1} \ldots i_{k}\right\}$ be the columns of the RREF of $A$ having a leading one. These columns of the RREF are linearly independent Thus these columns were originally linearly independent. If another column is linearly independent, this column of the RREF is linearly dependent on the columns with a leading one. This is a contradiction to the above proposition.

Proof of Theorem 2.2.3. By the way the RREF is constructed, left-to-right, and top-to-bottom, it should be apparent that if the right most row of the RREF is removed, there results the RREF of the $m \times(n-1)$ matrix formed from $A$ by deleting the $n^{\text {th }}$ column. Similarly, if the bottom row of the RREF is removed there results a new matrix in RREF form, though not simply related to the original matrix $A$.

To prove that the RREF is unique, we proceed by a double induction, first on the number of columns. We take it as given that for an $m \times 1$ matrix the RREF is unique. It is either the zero $m \times 1$ matrix, which would be the case if $A$ was zero or the matrix with a 1 in the first row and zeros in
the other rows. Assume therefore that the RREF is unique if the number of columns is less than $n$. Assume there are two RREF forms, $B_{1}$ and $B_{2}$ for $A$. Now the RREF of $A$ is therefore unique through the $(n-1)^{\text {st }}$ columns. The only difference between the RREF's $B_{1}$ and $B_{2}$ must occur in the $n^{\text {th }}$ column. Now proceed by induction on the number of nonzero rows. Assume that $A \neq 0$. If $A$ has just one row, the RREF of $A$ is simply the scalar multiple of $A$ that makes the first nonzero column entry a one. Thus it is unique. If $A=0$, the RREF is also zero. Assume now that the RREF is unique for matrices with less than $m$ rows. By the comments above that the only difference between the RREF's $B_{1}$ and $B_{2}$ can occur at the ( $m, n$ )-entry. That is $\left(B_{1}\right)_{m, n} \neq\left(B_{2}\right)_{m, n}$. They are therefore not leading ones. (Why?) There is a leading one in the $m^{\text {th }}$ row, however, because it is a non zero row. Because the row spaces of $B_{1}$ and $B_{2}$ are identical, this results in a contradiction, and therefore the $(m, n)$-entries must be equal. Finally, $B_{1}=B_{2}$. This completes the induction. (Alternatively, the two systems pertaining to the RREF's must have the same solution set to the system $A x=0$. With $\left(B_{1}\right)_{m, n} \neq\left(B_{2}\right)_{m, n}$, it is easy to see that the solution sets to $B_{1} x=0$ and $B_{2} x=0$ must differ.)

Definition 2.2.2. Let $A \in M_{m, n}$ and $b \in R_{m}$ (or $\mathbb{C}_{n}$ ). Define

$$
[A \mid b]=\left[\begin{array}{cccc}
a_{11} & \ldots & a_{1 n} & b_{1} \\
a_{21} & \ldots & a_{2 n} & b_{2} \\
a_{m 1} & \ldots & a_{m n} & b_{m}
\end{array}\right]
$$

$[A \mid b]$ is called the augmented matrix of $A$ by $b .[A \mid b] \in M_{m, n+1}(F)$. The augmented matrix is a useful notation for finding the solution of systems using row operations.

Identical to other definitions for solutions of equations, the equivalence of two systems is defined via the idea of equality of the solution set.

Definition 2.2.3. Two linear systems $A x=b$ and $B x=c$ are called equivalent if one can be converted to the other by elementary equation operations.

It is easy to see that this implies the following
Theorem 2.2.4. Two linear systems $A x=b$ and $B x=c$ are equivalent if and only if both $[A \mid b]$ and $[B \mid c]$ have the same row reduced echelon form.

We leave the prove to the reader. (See Exercise 23.) Note that the solution set need not be a single vector; it can be null or infinite.

### 2.3 Rank

Definition 2.3.1. The rank of any matrix $A$, denote by $r(A)$, is the dimension of its column space.

Proposition 2.3.1. (i) The rank of $A$ equals the number of nonzero rows of the RREF of A, i.e. the number of leading ones.
(ii) $r(A)=r\left(A^{T}\right)$.

Proof. (i) Follows from previous results.
(ii) The number of linearly independent rows equals the number of linearly independent columns. The number of linearly independent rows is the number of linearly independent columns of $A^{T}$-by definition. Hence $r(A)=r\left(A^{T}\right)$.

Proposition 2.3.2. Let $A \in M_{m, n}(\mathbb{C})$ and $b \in \mathbb{C}_{m}$. Then $A x=b$ has $a$ solution if and only if $r(A)=r([A \mid b])$, where $[A \mid b]$ is the augmented matrix.
Remark 2.3.1. Solutions may exist and may not. However, even if a solution exists, it may not be unique. Indeed if it is not unique, there is an infinity of solutions.
Definition 2.3.2. When $A x=b$ has a solution we say the system is consistent.

Naturally, in practical applications we want our systems to be consistent. When they are not, this can be an indicator that something is wrong with the underlying physical model. In mathematics, we also want consistent systems; they are usually far more interesting and offer richer environments for study.

In addition to the column and row spaces, another space of great importance is the so-called null space, the set of vectors $x \in R_{n}$ for which $A x=0$. In contrast, when solving the simple single variable linear equation $a x=b$ with $a \neq 0$ we know there is always a unique solution $x=b / a$. In solving even the simplest higher dimensional systems, the picture is not as clear.
Definition 2.3.3. Let $A \in M_{m, n}(F)$. The null space of $A$ is defined to be

$$
\operatorname{Null}(A)=\left\{x \in R_{n} \mid A x=0\right\}
$$

It is a simple consequence of the linearity of matrix multiplication that $\operatorname{Null}(A)$ is a linear subspace of $R_{n}$. That is to say, $\operatorname{Null}(A)$ is closed under vector addition and scalar multiplication. In fact, $A(x+y)=A x+A y=$ $0+0=0$, if $x, y \in \operatorname{Null}(A)$. Also, $A(\alpha x)=\alpha A x=0$, if $x \in \operatorname{Null}(A)$. We state this formally as

Theorem 2.3.1. Let $A \in M_{m, n}(F)$. Then $\operatorname{Null}(A)$ is a subspace of $R_{n}$ while the range of $A$ is in $R_{m}$.

Having such solutions gives valuable information about the solution set of the linear system $A x=b$. For, if we have found a solution, $x$, and have any vector $z \in \operatorname{Null}(A)$, then $x+z$ is a solution of the same linear system. Indeed, what is easy to see is that if $u$ and $v$ are both solutions to $A x=b$, then $A(u-v)=A u-A v=0$, or what is the same $x-y \in \operatorname{Null}(A)$. This means that to find all solutions to $A x=b$, we need only find a single solution and the null space. We summarize this as the following theorem.

Theorem 2.3.2. Let $A \in M_{m, n}(F)$ with null space $\operatorname{Null}(A)$. Let $x$ be any nonzero solution to $A x=b$. Then the set $x+\operatorname{Null}(A)$ is the entire solution set to $A x=b$.
Example 2.3.1. Find the null space of $A=\left[\begin{array}{cc}1 & 3 \\ -3 & -9\end{array}\right]$.
Solution. Solve $A x=0$. The RREF for $A$ is $\left[\begin{array}{ll}1 & 3 \\ 0 & 0\end{array}\right]$. Solving $x_{1}+3 x_{2}=0$, take $x_{2}=t$, a "free" parameter and solve for $x_{1}$ to get $x_{1}=-3 t$. Thus every solution to $A x=0$ can be written in the form

$$
x=\left[\begin{array}{c}
-3 t \\
t
\end{array}\right]=t\left[\begin{array}{c}
-3 \\
1
\end{array}\right] \quad t \in R
$$

Expressed this way we see that $\operatorname{Null}(A)=\left\{\left.t\left[\begin{array}{c}-3 \\ 1\end{array}\right] \right\rvert\, t \in R\right\}$, a subspace of $R_{2}$ of dimension 1 .

Theorem 2.3.3 (Fundamental theorem on rank). $A \in M_{m, n}(F)$. The following are equivalent
(a) $r(A)=k$.
(b) There exist exactly $k$ linearly independent columns of $A$.
(c) There exist exactly $k$ linearly independent rows of $A$.
(d) The dimension of the column space of $A$ is $k($ i.e. $\operatorname{dim}($ Range $A)=k)$.
(e) There exists a set $S$ of exactly $k$ vectors in $R_{m}$ for which $A x=b$ has a solution for each $b \in \mathfrak{S}(S)$.
(f) The null space of $A$ has dimension $n-k$.

Proof. The equivalence of (a), (b), (c) and (d) follow from previous considerations. To establish (e), let $S=\left\{c_{\ell_{1}}, c_{\ell_{2}}, \ldots, c_{\ell_{k}}\right\}$ denote the linearly independent column vectors of $A$. Let $T=\left\{e_{\ell_{1}}, e_{\ell_{2}}, \ldots, e_{\ell_{k}}\right\} \subset R_{n}$ be the standard vectors. Then $A e_{\ell_{j}}=c_{\ell_{j}}$. If $b \in \mathfrak{S}(S)$, then $b=a_{1} c_{\ell_{1}}+a_{2} c_{\ell_{2}}+$ $\cdots+a_{k} c_{\ell_{k}}$. A solution to $A x=b$ is given by $x=a_{1} e_{\ell_{1}}+a_{2} e_{\ell_{2}}+\cdots+a_{k} e_{\ell_{k}}$. Conversely, if (e) holds, then the set $S$ must be linearly independent for otherwise $S$ could be reduced to $k-1$ or fewer vectors. Similarly if $A$ has $k+1$ linearly independent columns then set $S$ can be expanded. Therefore, the column space of $A$ must have exactly $k$ vectors.

To prove (f) we assume that $S=\left\{v_{1}, \ldots, v_{k}\right\}$ is a basis for the column space of $A$. Let $T=\left\{w_{1}, \ldots, w_{k}\right\} \subset R_{n}$ for which $A w_{i}=v_{i}, i=1, \ldots, k$. By our extension theorem, we select $n-k$ vectors $w_{k+1}, \ldots, w_{n}$ such that $U=\left\{w_{1}, \ldots, w_{k}, w_{k+1}, \ldots, w_{n}\right\}$ is a basis of $R_{n}$. We must have that $A w_{k+1} \in \mathfrak{S}(S)$. Hence there are scalars $b_{1}, \ldots, b_{k}$ such that

$$
A w_{k+1}=A\left(b_{1} w_{1}+\cdots+b_{k} w_{k}\right)
$$

and thus $w_{k+1}^{\prime}=w_{k+1}-\left(b_{1} w_{1}+\cdots+b_{k} w_{k}\right)$ is in the null space of $A$. Repeat this process for each $w_{k+j}, j=1, \ldots, n-k$. We generate a total of $n-k$ vectors $\left\{w_{k+1}^{\prime}, \ldots, w_{n}^{\prime}\right\}$ in this manner. This set must be linearly independent. (Why?) Therefore, the dimension of the null space must be at least $n-k$. Now we consider a new basis which consists of the original vectors and the $n-k$ vectors $\left\{w_{k+1}^{\prime}, w_{k+2}^{\prime}, \ldots, w_{n}^{\prime}\right\}$ for which $A w=0$. We assert that the dimension of the null space is exactly $n-k$. For if $z \in R_{n}$ is a vector for which $A z=0$, then $z$ can be uniquely written as a component $z_{1}$ from $\mathfrak{S}(T)$ and a component $z_{2}$ from $\mathfrak{S}\left(\left\{w_{k+1}^{\prime}, \ldots, w_{n}^{\prime}\right\}\right)$. But $A z_{1} \neq 0$ and $A z_{2}=0$. Therefore $A z=0$ is impossible unless the component $z_{1}=0$.

Conversely, if ( f ) holds we take a basis for the null space $T=\left\{u_{1}, u_{2}, \ldots, u_{n-k}\right\}$ and extend the basis

$$
T^{\prime}=T \cup\left\{u_{n-k+1}, \ldots, u_{n}\right\}
$$

to $R_{n}$. Next argue similarly to above that

$$
A u_{n-k+1}, A u_{n-k+2}, \ldots, A u_{n}
$$

must be linearly independent, for otherwise there is yet another linearly independent vector that can be added to its basis, a contradiction. Therefore the column space must have dimension at least, and hence equal to $k$.

The following corollary assembles many consequences of this theorem.

Corollary 2.3.1. (1) $r(A) \leq \min (m, n)$.
(2) $r(A B) \leq \min (r(A), r(B))$.
(3) $r(A+B) \leq r(A)+r(B)$.
(4) $r(A)=r\left(A^{T}\right)=r\left(A^{*}\right)=r(\bar{A})$.
(5) If $A \in M_{m}(F)$ and $B \in M_{m, n}(F)$, and if $A$ is invertible, then

$$
r(A B)=r(B)
$$

Similarly, if $C \in M_{n}(F)$ is invertible and $B \in M_{m, n}(F)$

$$
r(B C)=r(B)
$$

(6) $r(A)=r\left(A^{T} A\right)=r\left(A^{*} A\right)$.
(7) Let $A \in M_{m, n}(F)$, with $r(A)=k$. Then $A=X B Y$ where $X \in M_{m, k}$, $Y \in M_{k, n}$ and $B \in M_{k}$ is invertible.
(8) In particular, every rank 1 matrix has the form $A=x y^{T}$, where $x \in$ $R_{m}$ and $y \in R_{n}$. Here

$$
x y^{T}=\left[\begin{array}{cccc}
x_{1} y_{1} & x_{1} y_{2} & \ldots & x_{1} y_{n} \\
\vdots & \vdots & & \vdots \\
x_{m} y_{1} & x_{m} y_{2} & \ldots & x_{m} y_{n}
\end{array}\right] .
$$

Proof. (1) The rank of any matrix is the number of linearly independent rows, which is the same as the number of linearly independent columns. The maximum this value can be is therefore the maximum of the minimum of the dimensions of the matrix, or $r(A) \leq \min (m, n)$.
(2) The product $A B$ can be viewed in two ways. The first is as a set of linear combinations of the rows of $B$, and the other is as a set of linear combinations of the columns of $A$. In either case the number of linear independent rows (or columns as the case may be) In other words, the rank of the product $A B$ cannot be greater than the number of linearly independent columns of $A$ nor greater than the number of linearly independent rows of $B$. Another way to express this is as $r(A B) \leq \min (r(A), r(B))$
(3) Now let $S=\left\{v_{1}, \ldots v_{r(A)}\right\}$ and $T=\left\{w_{1}, \ldots, w_{r(B)}\right\}$ be basis of the column spaces of $A$ and $B$ respectively. Then, the dimension of the union $S \cup T=\left\{v_{1}, \ldots v_{r(A)}, w_{1}, \ldots, w_{r(B)}\right\}$ cannot exceed $r(A)+r(B)$. Also, every vector in the column space of $A+B$ is clearly in the span of $S \cup T$. The result follows.
(4) The rank of $A$ is the number of linearly independent rows (and columns) of $A$, which in turn is the number of linearly independent columns of $A^{T}$, which in turn is the rank of $A^{T}$. That is, $r(A)=r\left(A^{T}\right)$. Similar proofs hold for $A^{*}$ and $\bar{A}$.
(5) Now suppose that $A \in M_{m}(F)$ is invertible and $B \in M_{m, n}$. As we have emphasized many times the rows of the product $A B$ can be viewed as a set of linear combinations of the rows of $B$. Since $A$ has rank $m$ any set of linearly independent rows of $B$ remains linearly independent. To see why, let $r_{i}(A B)$ denote the $i^{\text {th }}$ row of the product $A B$. Then it is easy to see that

$$
r_{i}(A B)=\sum_{j=1}^{m} a_{i j} r_{j}(B)
$$

Suppose we can determine constants $c_{1}, \ldots, c_{m}$ not all zero so that

$$
\begin{aligned}
0 & =\sum_{j=1}^{m} c_{i} r_{i}(A B)=\sum_{i=1}^{m} c_{i} \sum_{j=1}^{m} a_{i j} r_{j}(B) \\
& =\sum_{j=1}^{m} r_{j}(B) \sum_{i=1}^{m} c_{i} a_{i j}
\end{aligned}
$$

This linear combination of the rows of $B$ has coefficient given by $A^{T} c$, where $c=\left[c_{1}, \ldots, c_{k}\right]^{T}$. Because the rank of $A$ (and $A^{T}$ ) is $m$, we can solve this system for any vector $d \in R_{m}$. Suppose that the row vectors $r_{j_{l}}(B), \quad \ell=1, \ldots, r(B)$, are linearly independent. Arrange that the components of $d$ to be zero for indices not included in the set $j_{l}, \ell=1, \ldots, r(B)$ and not all zero otherwise. Then the conclusion $0=\sum_{j=1}^{m} r_{j}(B) \sum_{i=1}^{m} c_{i} a_{i j}=\sum_{l=1}^{r(B)} r_{j_{l}}(B) d_{j_{l}}$ is impossible. Indeed, the same basis of the row space of $B$ will be a basis of the row space of $A B$. This proves the result.
(6) We postpone the proof of this result until we discuss orthogonality.
(7) Place $A$ in RREF, say $A_{\text {RREF }}$. Since $r(A)=k$ we know the top $k$ rows of $A_{\text {RREF }}$ are linearly independent and the remaining rows are zero. Define $Y$ to be the $k \times n$ matrix consisting of these top $k$ rows. Define $B=I_{k}$. Now the rows of $A$ are linear combinations of these rows. So, define the $m \times k$ matrix $X$ to have rows as follows: The first row of consists of the coefficients so that $\sum x_{1 j} r_{j}(Y)=r_{1}(A)$. In general, the $i^{\text {th }}$ row of $X$ is selected so that

$$
\sum x_{i j} r_{j}(Y)=r_{i}(A)
$$

(8) This is an application of (7) noting in this special case that $X$ is an $m \times 1$ matrix that can be interpretted as a vector $x \in R_{m}$. Similarly, $Y$ is an $1 \times n$ matrix that can be interpretted as a vector $y \in R_{n}$. Thus, with $I=[1]$, we have

$$
A=x y^{T}
$$

Example 2.3.2. Here is the decomposition of the form given in Lemma 2.3.1 (7). The $3 \times 4$ matrix $A$ has rank 2 .

$$
A=\left[\begin{array}{ccc}
1 & 2 & -1 \\
0 & 0 & 2 \\
-1 & -2 & 3 \\
2 & 4 & 0
\end{array}\right]=\left[\begin{array}{cc}
1 & -1 \\
0 & 2 \\
-1 & 3 \\
2 & 0
\end{array}\right]\left[\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right]\left[\begin{array}{lll}
1 & 2 & 0 \\
0 & 0 & 1
\end{array}\right]=X B Y
$$

The matrix $Y$ is the RREF of $A$.
Example 2.3.3. Let $x=\left[x_{1}, x_{2}, \ldots, x_{m}\right]^{T} \in R_{m}$ and $y=\left[y_{1}, y_{2}, \ldots y_{n}\right]^{T} \in$ $R_{n}$. Then the rank one $m \times n$ matrix $x y^{T}$ has the form

$$
x y^{T}=\left[\begin{array}{cccc}
x_{1} y_{1} & x_{1} y_{2} & \cdots & x_{1} y_{n} \\
x_{2} y_{1} & x_{2} y_{2} & & x_{2} y_{n} \\
\vdots & & \ddots & \vdots \\
x_{m} y_{1} & x_{m} y_{2} & \cdots & x_{m} y_{n}
\end{array}\right]
$$

In particular, with $x=[1,3,5]^{T}$, and $y=[-2,7]^{T}$, the rank one $3 \times 2$ matrix $x y^{T}$ is given by

$$
x y^{T}=\left[\begin{array}{l}
1 \\
3 \\
5
\end{array}\right][-2,7]=\left[\begin{array}{cc}
-2 & 7 \\
-6 & 21 \\
-10 & 35
\end{array}\right]
$$

## Invertible Matrices

A subclass matrices $A \in M_{n}(F)$ that have only the zero kernel is very important in applications and theoretical developments.

Definition 2.3.4. $A \in M_{n}$ is called nonsingular if $A x=0$ implies that $x=0$.

In many texts such matrices are introduced though an equivalent alternate definition involving rank.

Definition 2.3.5. $A \in M_{n}$ is nonsingular if $r(A)=n$.
We also say that nonsingular matrices have full rank. That nonsingular matrices are invertible and conversely together with many other equivalences is the content of the next theorem.

Theorem 2.3.4. [Fundamental theorem on inverses] Let $A \in M_{n}(F)$. Then the following statements are equivalent.
(a) $A$ is nonsingular.
(b) $A$ is invertible.
(c) $r(A)=n$.
(d) The rows and columns of $A$ are linearly independent.
(e) $\operatorname{dim}(\operatorname{Range}(A))=n$.
(f) $\operatorname{dim}(\operatorname{Null}(A))=0$.
(g) $A x=b$ is consistent for all $b \in R_{n}$ (or $\mathbb{C}_{n}$ ).
(h) $A x=b$ has a unique solution for every $x \in R_{n}\left(\right.$ or $\left.\mathbb{C}_{n}\right)$.
(i) $A x=0$ has only the zero solution.
(j)* 0 is not an eigenvalue of $A$.
$(k)^{*} \operatorname{det} A \neq 0$.

* The statements about eigenvalues and the determinant $(\operatorname{det} A)$ of a matrix will be clarified later after they have been properly defined. They are included now for completeness.

Definition 2.3.6. Two linear systems $A x=b$ and $B x=c$ are called equivalent if one can be converted to the other by elementary equation operations. Equivalently, the systems are equivalent if $[A \mid b]$ can be converted to $[B \mid c]$ by elementary row operations.

Alternatively, the systems are equivalent if they have the same solution set which means of course that both can be reduced to the same RREF.

Theorem 2.3.5. If $A \in M_{n}(F)$ and $B \in M_{n}(F)$ with $A B=I$, then $B$ is unique.

Proof. If $A B=I$ then for every $e_{1} \ldots e_{n}$ there is a solution to the system $A b_{i}=e_{i}$ for all $1=1,2, \ldots, n$. Thus the set $\left\{b_{i}\right\}_{i=1}^{n}$ is linearly independent (because the set $\left\{e_{i}\right\}$ is) and moreover a basis. Similarly if $A C=I$ then $A(C-B)=0$, and there are $c_{i} \in R_{n}\left(\right.$ or $\left.\mathbb{C}_{n}\right), i=1, \ldots, n$ such that $A c_{i}=e_{i}$. Suppose for example that $c_{1}-b_{1} \neq 0$. Since the $\left\{b_{i}\right\}_{i=1}^{n}$ is a basis it follows that $c_{1}-b_{1}=\Sigma \alpha_{j} b_{j}$, where not all $\alpha_{j}$ are zero. Therefore,

$$
A\left(c_{1}-b_{1}\right)=\Sigma \alpha_{j} A b_{j}=\Sigma \alpha_{j} e_{j} \neq 0
$$

and this is a contradiction.
Theorem 2.3.6. Let $A \in M_{n}(F)$. If $B$ is a right inverse, $A B=I$, then $B$ is a left inverse.

Proof. Define $C=B A-I+B$, and assume $C \neq B$ or what is the same thing that $B$ is not a left inverse. Then

$$
\begin{aligned}
A C & =A B A-A+A B=(A B)(A)-A+A B \\
& =A-A+A B=I
\end{aligned}
$$

This implies that $C$ is another right inverse of $A$, contradicting Theorem 2.3.5.

### 2.4 Orthogonality

Let V be a vector space over $\mathbb{C}$. We define an inner product $\langle\cdot, \cdot\rangle$ on $V \times V$ to be a function from $V$ to $\mathbb{C}$ that satisfies the following properties:

1. $\langle a v, w\rangle=a\langle v, w\rangle$ and $\langle v, a w\rangle=\bar{a}\langle v, w\rangle \quad(\bar{a}$ is the complex conjugate of $a$ )
2. $\langle v, w\rangle=\overline{\langle w, v\rangle}$
3. $\langle u+v, w\rangle=\langle u, w\rangle+\langle v, w\rangle$ (linearity)
4. $\langle u, v+w\rangle=\langle u, v\rangle+\langle u, w\rangle$
5. $\langle v, v\rangle \geq 0$ with $\langle v, v\rangle=0$ if and only if $v=0$.

For inner products over real vector spaces, we neglect the complex conjugate operation. In addition, we want our inner products to define a norm as follows:
6. For any $v \in V,\|v\|^{2}=\langle v, v\rangle$

We assume thoughout the text that all vector spaces with inner products have norms defined exactly in this way. With the norm and vector $v$ can be normalized by dilating it to have length 1 , say $v_{n}=v \frac{1}{\|v\|}$. The simplest type of inner product on $\mathbb{C}_{n}$ is given by

$$
\langle v, w\rangle=\sum_{i=1}^{n} x_{i} \bar{y}_{i}
$$

We call this the standard inner product.
Using any inner product, we can define an angle between vectors.
Definition 2.4.1. The angle $\theta_{x y}$ between vectors $x$ and $y$ in $R_{n}$ is defined by

$$
\begin{aligned}
\cos \theta_{x y} & =\frac{\langle x, y\rangle}{\|x\|\|y\|} \\
& =\frac{\langle x, y\rangle}{(\langle x, x\rangle)^{1 / 2}(\langle y, y\rangle)^{1 / 2}} .
\end{aligned}
$$

This comes from the well known result in $R_{2}$

$$
x \cdot y=\|x\|\|y\| \cos \theta
$$

which can be proved using the law of cosines. With angle comes the notion of orthogonality.
Definition 2.4.2. Two vectors $u$ and $v$ are said to be orthogonal if the angle between them if $\frac{\pi}{2}$ or what is the same thing $\langle u, v\rangle=0$. In this case we commonly write $x \perp y$. We extend this notation to sets $U$ writing $x \perp U$ to mean that $x \perp u$ for every $u \in U$. Similarly two sets $U$ and $V$ are called orthogonal if $u \perp v$ for every $u \in U$ and $v \in V$.

Remark 2.4.1. It is important to note that the notion of orthogonality depends completely on the inner product. For example, the weighted inner product defined by $\langle v, w\rangle=\sum_{i=1}^{n} w_{i} x_{i} \bar{y}_{i}$ where the $w_{i}>0$ gives very different orthogonal vectors from the standard inner product.

Example 2.4.1. In $R_{n}$ or $\mathbb{C}_{n}$ the standard unit vectors are orthogonal with respect to the standard inner product.

Example 2.4.2. In the $R_{3}$ the vectors $u=(1,2,-1)$ and $v=(1,1,3)$ are orthogonal because

$$
\langle x, y\rangle=1(1)+2(2)-1(3)=0
$$

Note that in $R_{3}$ the complex conjugate is not written. The set of vectors $\left(x_{1}, x_{2}, x_{3}\right) \in R_{3}$ orthogonal to $u=(1,2,-1)$ satisfies the equation $x_{1}+$ $2 x_{2}-x_{3}=0$ is recognizable as the plane with normal vector $u$.

Definition 2.4.3. We define the projection $P_{u} v$ of one vector $v$ in the direction of an other vector $u$ to be

$$
P_{u} v=\frac{\langle u, v\rangle}{\|u\|^{2}} u
$$

As you can see, we have merely written an expression for the more intuitive version of the projection in question given by $\|v\| \cos \theta_{u v} \frac{u}{\|u\|}$. In the figure below, we show the fundamental diagram for the projection of one vector in the direction of another.


If the vectors $u$ and $v$ are orthogonal, it is easy to see that $P_{u} v=0$. (Why?)

Example 2.4.3. Find the projection of the vector $v=(1,2,1)$ on the vector $u=(-2,1,3)$

Solution. We have

$$
\begin{aligned}
P_{u} v & =\frac{\langle u, v\rangle}{\|u\|^{2}} u=\frac{\langle(1,2,1),(-2,1,3)\rangle}{\|(-2,1,3)\|^{2}}(-2,1,3) \\
& =\frac{1(-2)+2(1)+1(3)}{14}(-2,1,3) \\
& =\frac{5}{14}(-2,1,3)
\end{aligned}
$$

We are now ready to find orthogonal sets of vectors and orthogonal bases. First we make an important definition.

Definition 2.4.4. Let $V$ be a vector space with an inner product. A set of vectors $S=\left\{x_{1}, \ldots, x_{n}\right\}$ in $V$ is said to be orthogonal if $\left\langle x_{i}, x_{j}\right\rangle=0$ for $i \neq j$. It is called orthonormal if also $\left\langle x_{i}, x_{i}\right\rangle=1$. If, in addition, $S$ is a basis it is called an orthogonal basis or orthonomal basis.

Note: Sometimes the conditions for orthonormality are written as

$$
\left\langle x_{i}, x_{j}\right\rangle=\delta_{i j}
$$

where $\delta_{i j}$ is the "Dirac" delta: $\delta_{i j}=0, i \neq j, \delta_{i i}=1$.
Theorem 2.4.1. Suppose $U$ is a subspace of the (inner product) vector space $V$ and that $U$ has the basis $S=\left\{x_{1} \ldots x_{k}\right\}$, then $U$ has an orthogonal basis.

Proof. Define $y_{1}=\frac{x_{1}}{\|x\| ा}$. Thus $y_{1}$ is the "normalized" $x_{1}$. Now define the new orthonormal basis recursively by

$$
\begin{aligned}
y_{j+1}^{\prime} & =x_{j+1}-\sum_{i=1}^{j}\left\langle y_{i}, x_{j+1}\right\rangle y_{i} \\
y_{j+1} & =\frac{y_{j+1}^{\prime}}{\left\|y_{j+1}^{\prime}\right\|}
\end{aligned}
$$

for $j=1,2, \ldots, k-1$. Then
(1) $y_{j+1}$ is orthogonal to $y_{1}, \ldots, y_{j}$
(2) $y_{j+1} \neq 0$.

In the language above we have $\left\langle y_{i}, y_{j}\right\rangle=\delta_{i j}$.

Basically, what the proof accomplishes is to take the differences of the vector from the projections to the others. Referring to the figure above we compute $v-P_{u} v$ as noted in the figure below. The process of orthogonalization described above is called the Gram-Schmidt process.


## Representation of vectors

One of the great advantages of orthonormal bases is that they make the representation of vectors particularly easy. It is as simple as computing an inner product. Let $V$ be a vector space with inner product $\langle\cdot, \cdot\rangle$ and with subspace $U$ having basis $S=\left\{u_{1}, u_{2}, \ldots, u_{k}\right\}$. Then for every $u \in U$ we know there are constants $a_{1}, a_{2}, \ldots, a_{k}$ such that

$$
x=a_{1} u_{1}+a_{2} u_{2}+\cdots+a_{k} u_{k}
$$

Taking the inner product of both sides with $u_{j}$ and applying the orthogonality relations

$$
\begin{aligned}
\left\langle x, u_{j}\right\rangle & =\left\langle a_{1} u_{1}+a_{2} u_{2}+\cdots+a_{k} u_{k .}, u_{j}\right\rangle \\
& =\sum_{j=1}^{k} a_{i}\left\langle u_{i .}, u_{j}\right\rangle=a_{j}
\end{aligned}
$$

Thus $a_{j}=\left\langle x, u_{j}\right\rangle, j=1,2, \ldots, k$, and

$$
x=\sum_{j=1}^{k}\left\langle u_{.}, u_{j}\right\rangle u_{j}
$$

Example 2.4.4. One basis of $R_{2}$ is given by the orthonormal vectors $S=$ $\left\{u_{1}, u_{2}\right\}$, where $u_{1}=\left[\frac{1}{\sqrt{2}}, \frac{1}{\sqrt{2}}\right]^{T}$ and $u_{2}=\left[\frac{1}{\sqrt{2}},-\frac{1}{\sqrt{2}}\right]^{T}$. The representation of $x=[3,2]^{T}$ is given by

$$
x=\sum_{j=1}^{2}\left\langle u_{.}, u_{j}\right\rangle u_{j}=\frac{5}{2} \sqrt{2}\left[\frac{1}{\sqrt{2}}, \frac{1}{\sqrt{2}}\right]^{T}+\frac{1}{2} \sqrt{2}\left[\frac{1}{\sqrt{2}},-\frac{1}{\sqrt{2}}\right]^{T}
$$

## Orthogonal subspaces

Definition 2.4.5. For any set of vectors $S$ we define

$$
S^{\perp}=\{v \in V \mid v \perp S\}
$$

That is, $S^{\perp}$ is the set of vectors orthogonal to $S$. Often, $S^{\perp}$ is called the orthogonal complement or orthocomplement of $S$.

For example the orthocomplement of any vector $v=\left[v_{1}, v_{2}, v_{3}\right]^{T} \in R_{3}$ is the (unique) plane passing through the origin that is orthogonal to $v$. It is easy to see that the equation of the plane is $x_{1} v_{1}+x_{2} v_{2}+x_{3} v_{3}=0$.

For any set of vectors $S$ the orthocomplement $S^{\perp}$ has the remarkable property of being a subspace of $V$, and therefore it is must have an orthogonal basis.

Proposition 2.4.1. Suppose that $V$ is a vector space with an inner product, and $S \subset V$. Then $S^{\perp}$ is a subspace of $V$.

Proof. If $y_{1}, \ldots, y_{m} \in S^{\perp}$ then $\Sigma a_{i} y_{i} \in S^{\perp}$ for every set of coefficients $a_{1}, \ldots, a_{m}$ in $R$ (or $\mathbb{C}$ ).

Corollary 2.4.1. Suppose that $V$ is a vector space with an inner product, and $S \subset V$.
(i) If $S$ is a basis of $V, S^{\perp}=\{0\}$.
(ii) If $U=\mathfrak{S}(S)$, then $U^{\perp}=S^{\perp}$.

The proofs of these facts are elementary consequences of the proposition. An important decomposition result is based on orthogonality of subspaces. For example, suppose that $V$ is a finite dimensional inner product space and that $U$ is a subspace of $V$. Let $U^{\perp}$ be the orthocomplement of $U$, and let $S=\left\{u_{1}, u_{2}, \ldots, u_{k}\right\}$ be an orthonormal basis of $U$. Let $x \in V$. Define $x_{1}=\sum_{j=1}^{k}\left\langle x, u_{j}\right\rangle u_{j}$, and $x_{2}=x-x_{1}$. Then it follows that $x_{1} \in U$ and $x_{2} \in U^{\perp}$. Moreover, $x=x_{1}+x_{2}$. We summarize this in the following.

Proposition 2.4.2. Let $V$ is a vector space with inner product $\langle\cdot, \cdot\rangle$ and with subspace $U$. Then every vector $x \in V$ can be written as a sum of two orthogonal vectors $x=x_{1}+x_{2}$, where $x_{1} \in U$ and $x_{2} \in U^{\perp}$.

Geometrically what this results asserts is that for a given subspace of an inner product space, every vector has an orthogonal decomposition as two unique sum of a vector from the subspace and its orthocomplement. We write the vector components as the respective projections of the given vector to the orthogonal subspaces

$$
\begin{aligned}
& x_{1}=P_{U} x \\
& x_{2}=P_{U \perp x}
\end{aligned}
$$

Such decompositions are important in the analysis of vector spaces and matrices. In the case of vector spaces, of course, the representation of vectors is of great value. In the case of matrices, this type of decomposition serves to allow reductions of the matrices while preserving the information they carry.

### 2.4.1 An important equality for matrix multiplication and the inner product

Let $A \in M_{m n}(\mathbb{C})$. Then we know that both $A^{*} A$ and $A A^{*}$ (Alternatively, $A^{T} A$ and $A A^{T}$ exist) exist, and we can surely inquire about the rank of these matrices. The main result of this section is on the rank of $A^{T} A$, namely that $r(A)=r\left(A^{*} A\right)=r\left(A A^{*}\right)$. The proof is quite simple but requires an important equality. Let $A \in M_{m n}(\mathbb{C})$ and $v \in \mathbb{C}_{n}$ and $w \in \mathbb{C}_{m}$. Then

$$
\begin{aligned}
\langle A v, w\rangle & =\left\langle\sum_{i=1}^{m}(A v)_{i}, w_{i}\right\rangle \\
& =\sum_{i=1}^{m} \sum_{j=1}^{n} a_{i j} v_{j} \bar{w}_{i} \\
& =\sum_{j=1}^{n} v_{j} \sum_{i=1}^{m} a_{i j} \bar{w}_{i} \\
& =\sum_{j=1}^{n} v_{j} \overline{\sum_{i=1}^{m} \bar{a}_{i j} w_{i}} \\
& =\sum_{j=1}^{n} v_{j} \overline{\left(A^{*} w\right)_{j}} \\
& =\left\langle v, A^{*} w\right\rangle
\end{aligned}
$$

As a consequence we have $\left\langle A^{*} A v, w\right\rangle=\langle A v, A w\rangle$ if both $v, w \in \mathbb{C}_{n}$. This important equality allows the adjoint or transpose matrices to be used on either side of the inner product, as needed. Indeed we shall use this below.

Proposition 2.4.3. Let $A \in M_{m n}(\mathbb{C})$ have rank $r(A)$. Then

$$
r(A)=r\left(A^{*} A\right)=r\left(A A^{*}\right)
$$

Proof. Assume that $r(A)=k$. Then there are $k$ standard vectors $e_{j_{1}}, \ldots, e_{j_{k}}$ such that for each $l=1,2, \ldots k$, the vectors $A e_{j_{l}}$ is one of the linearly independent columns of $A$. Moreover, it also follows that for every set of constants $a_{1}, \ldots, a_{k}$ the vector $A\left(\sum a_{l} e_{l_{j}}\right) \neq 0$. Now $A^{*} A\left(\sum a_{l} e_{l_{j}}\right) \neq 0$ follows because

$$
\begin{aligned}
\left\langle A^{*} A\left(\sum a_{l} e_{l_{j}}\right),\left(\sum a_{l} e_{l_{j}}\right)\right\rangle & =\left\langle A\left(\sum a_{l} e_{l_{j}}\right), A\left(\sum a_{l} e_{l_{j}}\right)\right\rangle \\
& =\left\|A\left(\sum a_{l} e_{l_{j}}\right)\right\|^{2} \neq 0
\end{aligned}
$$

This in turn establishes that $A^{*} A$ cannot be zero on a linear space of dimension $k$ except for the zero element of course, and since the rank of $A^{*} A$ cannot be larger than $k$ the result is proved.

Remark 2.4.2. This establishes (6) of the Corollary 2.3 .1 above. Also, it is easy to see that the result is also true for real matrices.

### 2.4.2 The Legendre Polynomials

When a vector space has an inner product, it is possible to construct an orthogonal basis from any given basis. We do this now for the polynomial space $P_{n}(-1,1)$ and a particular basis.

Consider the space the polynomials of degree $n$ defined on the interval $[-1,1]$ over the reals. Recall that this is a vector space and has as a basis the monomials $\left\{1, x, x^{2}, \ldots, x^{n}\right\}$. We can define an assortment of inner products on this space, but the most common inner product is given by

$$
\langle p, q\rangle=\int_{-1}^{1} p(x) q(x) d x
$$

Verifying the inner product properties is fairly straight forward and we leave it as an exercise. This inner product also defines a norm

$$
\|p\|^{2}=\int_{-1}^{1}|p(x)|^{2} d x
$$

This norm satisfies the triangle inequality requires an integral version of the Cauchy-Schwartz inequality.

Now that we have an inner product and norm, we could proceed to find an othogonal basis of $P_{n}(-1,1)$ by applying the Gram-Schmidt procedure to the basis $\left\{1, x, x^{2}, \ldots, x^{n}\right\}$. This procedure can be clumsy and tedious. It is easier to build an orthogonal basis from scratch. Following tradition we will use capital letters $P_{0}, P_{1}, \ldots$ to denote our orthogonal polynomials. Toward this end take $P_{0}=1$. Note we are numbering from 0 onwards so that the polynomial degree will agree with the index. Now let $P_{1}=a x+b$. For orthogonality, we need

$$
\int_{-1}^{1} P_{0}(x) P_{1}(x) d x=\int_{-1}^{1} 1 \cdot(a x+b) d x=2 b=0
$$

Thus $b=0$ and $a$ can be arbitrary. We take $a=1$. This gives $y_{1}=x$. Now we assume the model for the next orthogonal function to be $y_{2}=a x^{2}+b x+c$. This time there are two orthogonality conditions to satisfy.

$$
\begin{aligned}
& \int_{-1}^{1} P_{0}(x) P_{2}(x) d x=\int_{-1}^{1} 1 \cdot\left(a x^{2}+b x+c\right) d x=\frac{2}{3} a+2 c=0 \\
& \int_{-1}^{1} P_{1}(x) P_{2}(x) d x=\int_{-1}^{1} x \cdot\left(a x^{2}+b x+c\right) d x=\frac{2}{3} b=0
\end{aligned}
$$

We conclude that $b=0$. From the equation $\frac{2}{3} a+2 c=0$, we can assign one of the variables and solve for the other one. Following tradition we take $c=-\frac{1}{2}$ and solve for $a$ to get $a=\frac{3}{2}$.

The next polynomial will be modeled as $P_{3}(x)=a x^{3}+b x^{2}+c x+d$. Three orthogonality relations need to be satisfied.

$$
\begin{aligned}
\int_{-1}^{1} P_{0}(x) P_{3}(x) d x & =\int_{-1}^{1} 1 \cdot\left(a x^{3}+b x^{2}+c x+d\right) d x=\frac{2}{3} b+2 d=0 \\
\int_{-1}^{1} P_{1}(x) P_{3}(x) d x & =\int_{-1}^{1} x \cdot\left(a x^{3}+b x^{2}+c x+d\right) d x=\frac{2}{5} a+\frac{2}{3} c=0 \\
\int_{-1}^{1} P_{2}(x) P_{3}(x) d x & =\int_{-1}^{1} \frac{1}{2}(3 x-1)\left(a x^{3}+b x^{2}+c x+d\right) d x \\
& =\frac{3}{5} a-\frac{1}{3} b+c-d=0
\end{aligned}
$$

It is easy to see that $b=d=0$ (why?) and from $\frac{2}{5} a+\frac{2}{3} c=0$, we select
$c=-\frac{3}{2}$ and $a=\frac{5}{2}$. Our table of orthogonal polynomials so far is

| $k$ | $P_{k}(x)$ |
| :---: | :---: |
| 0 | 1 |
| 1 | $x$ |
| 2 | $\frac{1}{2}(3 x-1)$ |
| 3 | $\frac{1}{2}\left(5 x^{3}-3 x\right)$ |

Continue in this fashion, generating polynomials of increasing order each orthogonal to all of the lower order ones.

$$
\begin{aligned}
P_{0}(x) & =1 \\
P_{1}(x) & =x \\
P_{2}(x) & =3 / 2 x^{2}-1 / 2 \\
P_{3}(x) & =5 / 2 x^{3}-3 / 2 x \\
P_{4}(x) & =\frac{35}{8} x^{4}-\frac{15}{4} x^{2}+3 / 8 \\
P_{5}(x) & =\frac{63}{8} x^{5}-\frac{35}{4} x^{3}+\frac{15}{8} x \\
P_{6}(x) & =\frac{231}{16} x^{6}-\frac{315}{16} x^{4}+\frac{105}{16} x^{2}-\frac{5}{16} \\
P_{7}(x) & =\frac{429}{16} x^{7}-\frac{693}{16} x^{5}+\frac{315}{16} x^{3}-\frac{35}{16} x \\
P_{8}(x) & =\frac{6435}{128} x^{8}-\frac{3003}{32} x^{6}+\frac{3465}{64} x^{4}-\frac{315}{32} x^{2}+\frac{35}{128} \\
P_{9}(x) & =\frac{12155}{128} x^{9}-\frac{6435}{32} x^{7}+\frac{9009}{64} x^{5}-\frac{1155}{32} x^{3}+\frac{315}{128} x \\
P_{10}(x) & =\frac{46189}{256} x^{10}-\frac{109395}{256} x^{8}+\frac{45045}{128} x^{6}-\frac{15015}{128} x^{4}+\frac{3465}{256} x^{2}-\frac{63}{256}
\end{aligned}
$$

### 2.4.3 Orthogonal matrices

Besides sets of vectors being orthogonal, there is also a definition of orthogonal matrices. The two notions are closely linked.

Definition 2.4.6. We say a matrix $A \in M_{n}(\mathbb{C})$ is orthogonal if $A^{*} A=I$. The same definition applies to matrices $A \in M_{n}(R)$ with $A^{*}$ replaced by $A^{T}$.

For example, the rotation matrices (Exercise ??) $B_{\theta}=\left[\begin{array}{cc}\cos \theta & -\sin \theta \\ \sin \theta & \cos \theta\end{array}\right]$ are all orthogonal.

A simple consequence of this definition is that the rows and the columns of $A$ are orthonormal. We see for example that when $A$ is orthogonal then $\left(A^{*}\right)^{2}=\left(A^{-1}\right)^{2}=A^{-1} A^{-1}=\left(A^{2}\right)^{-1}$. Such a definition applies, as well to higher powers. For instance, if $A$ is orthogonal then $A^{m}$ is orthogonal for every positive integer $m$.

One way to generate orthogonal matrices in $\mathbb{C}_{n}$ (or $R_{n}$ ) is to begin with an orthonormal basis and arrange it into an $n \times n$ matrix either as its columns or rows.

Theorem 2.4.2. (i) Let $\left\{x_{i}\right\}, i=1, \ldots, n$ be an orthonormal basis of $\mathbb{C}_{n}$ or $\left(R_{n}\right)$. Then the matrices

$$
U=\left[\begin{array}{ccc}
x_{1} & \cdots & x_{n} \\
\downarrow & \cdots & \downarrow \\
\cdot & & \cdot
\end{array}\right] \quad \text { and } \quad V=\left[\begin{array}{ccc}
x_{1} & \longrightarrow & \cdot \\
\vdots & \vdots & \\
x_{n} & \longrightarrow & \cdot
\end{array}\right]
$$

formed by arranging the vectors $x_{i}$ as its respective columns or rows are orthogonal.
(ii) Conversely, $U$ is an orthogonal matrix, the sets of its rows and columns are each orthonormal, and moreover each forms a basis of $\mathbb{C}_{n}$ or $\left(R_{n}\right)$.

The proofs are entirely trivial. We shall consider these types of results in more detail later in Chapter 4. In the meantime there are a few more interesting results that are direct consequences of the definition and facts about the transpose (adjoint).

Theorem 2.4.3. Let $A, B \in M_{n}(C)$ (or $\left.M_{n}(R)\right)$ be orthogonal matrices. Then
(a) $A$ is invertible and $A^{-1}=A^{*}$.
(b) For each integer $k=0, \pm 1, \pm 2, \ldots$, both $A^{k}$ and $-A^{k}$ are orthogonal. (c) $A B$ is orthogonal.

### 2.5 Determinants

This section is about determinants that can be regarded as a measure of singularity of a matrix. More generally, in many applied situations that deal with complex objects, a single number is sought that will in some way
classify an aspect of those objects. The determinant is such a measure for singularity of the matrix. The determinant is difficult to calculate and of not much practical use. However, it has considerable theoretical value and certainly has a place of historical interest.

Definition 2.5.1. Let $A \in M_{n}(F)$. Define the determinant of $A$ to be the value in $F$

$$
\operatorname{det} A=\sum_{\sigma}\left(\prod_{i=1}^{n} a_{i \sigma(i)}\right) \cdot \operatorname{sgn} \sigma
$$

where $\sigma$ is a permutation of the integers $\{1,2, \ldots, n\}$ and
(1) $\sum_{\sigma}$ denotes the sum over all permutations
(2) $\operatorname{sgn} \sigma=\operatorname{sign}$ of $\sigma= \pm 1$

A transposition is the exchange of two elements of an ordered list with all others staying the same. With respect to permutations, a transposition of one permutation is another permutation formed by the exchange of two values. For example a transposition of $\{1,4,3,2\}$ is $\{1,3,4,2\}$. The sign of a given permutation $\sigma$ is
(a) +1 , if the number of transpositions required to bring $\sigma$ to $\{1,2, \ldots, n\}$ is even.
(b) -1 , if the number of transpositions required to bring $\sigma$ to $\{1,2, \ldots, n\}$ is odd.

Alternatively, and what is the same thing, we may count the number $m$ of transpositions required to bring $\sigma$ to $\{1,2, \ldots, n\}$ and to compute the sign is $(-1)^{m}$.

## Example 2.5.1.

$$
\begin{array}{rlll}
\sigma_{1}=\{2,1,3\} & \xrightarrow{1 \leftrightarrow 2}\{1,2,3\} & & \text { odd } \\
\sigma_{2}=\{2,3,1\} & \xrightarrow{3 \leftrightarrow 1}\{2,1,3\} & \xrightarrow{1 \leftrightarrow 2}\{1,2,3\} & \text { even } \\
\operatorname{sgn} \sigma_{1}=-1 & \operatorname{sgn} \sigma_{2}=+1 &
\end{array}
$$

Proposition 2.5.1. Let $A \in M_{n}$.
(i) If two rows of $A$ are interchanged to obtain $B$, then

$$
\operatorname{det} B=-\operatorname{det} A
$$

(ii) Given $A \in M_{n}(F)$. If any row is multiplied by a scalar $c$, the resulting matrix $B$ has determinant

$$
\operatorname{det} B=c \operatorname{det} A \text {. }
$$

(iii) If any two rows of $A \in M_{n}(F)$ are equal,

$$
\operatorname{det} A=0 .
$$

Proof. (i) Suppose rows $i_{1}$ and $i_{2}$ are interchanged. Now for the given permutations $\sigma$ apply the transposition $i_{1} \leftrightarrow i_{2}$ to get $\sigma_{1}$. Then

$$
\prod_{i=1}^{n} a_{i_{1} \sigma(i)}=\prod_{i=1}^{n} b_{i_{2} \sigma_{1}(i)}
$$

because

$$
a_{i_{1} \sigma\left(i_{1}\right)}=b_{i_{2} \sigma_{1}\left(i_{2}\right)}
$$

as

$$
b_{i_{2} j}=a_{i_{1} j} \quad \text { and } \quad \sigma_{1}\left(i_{2}\right)=\sigma_{2}\left(i_{1}\right)
$$

and similarly $a_{i_{2} \sigma\left(i_{2}\right)}=b_{i_{1} \sigma_{1}\left(i_{1}\right)}$. All other terms are equal. In the computation of the full determinant with signs of the permutations, we see that the change is caused only by the fact $\operatorname{sgn}\left(\sigma_{1}\right)=-\operatorname{sgn}(\sigma)$. Thus,

$$
\operatorname{det} B=-\operatorname{det} A .
$$

(ii) Is trivial.
(iii) If two rows are equal then by part (i)

$$
\operatorname{det} A=-\operatorname{det} A
$$

and this implies $\operatorname{det} A=0$.

Corollary 2.5.1. Let $A \in M_{n}$. If $A$ has two rows equal up to a multiplicative constant, it has has determinant zero.

What happens to the determinant when two matrices are added. The result is too complicated to write down is not very important. However, when a single vector is added to a row or a column of a matrix, then the result can be simply stated.

Proposition 2.5.2. Suppose $A \in M_{n}(F)$. Suppose $B$ is obtained from $A$ by adding a vector $v$ to a given row (resp. column) and $C$ is obtained from $A$ by replacing the given row (resp. column) by the vector $v$. Then

$$
\operatorname{det} B=\operatorname{det} A+\operatorname{det} C .
$$

Proof. Assume the $j^{\text {th }}$ row is altered. Using the definition of the determinant,

$$
\begin{aligned}
\operatorname{det} A= & \sum_{\sigma} \operatorname{sgn}(\sigma) \prod_{i} b_{i \sigma(i)}=\sum_{\sigma} \operatorname{sgn}(\sigma)\left(\prod_{i \neq j} b_{i \sigma(i)}\right) b_{j \sigma(j)} \\
= & \sum_{\sigma} \operatorname{sgn}(\sigma)\left(\prod_{i \neq j} a_{i \sigma(i)}\right)(a+v)_{j \sigma(j)} \\
= & \sum_{\sigma} \operatorname{sgn}(\sigma)\left(\prod_{i \neq j} a_{i \sigma(i)}\right) a_{j \sigma(j)}+\sum_{\sigma} \operatorname{sgn}(\sigma)\left(\prod_{i \neq j} a_{i \sigma(i)}\right) v_{j \sigma(j)} \\
& \operatorname{det} A+\operatorname{det} C
\end{aligned}
$$

For column replacement the proof is similar, particularly using the alternate representation of the determinant given in Exercise 15.

Corollary 2.5.2. Suppose $A \in M_{n}(F)$ and $B$ is obtained by multiplying a given row (resp. column) of $A$ by a scalar and adding it to another row (resp. column), then

$$
\operatorname{det} B=\operatorname{det} A \text {. }
$$

Proof. First note that in applying Proposition 2.5.2 $C$ has two rows equal up to a multiplicative constant. Thus $\operatorname{det} C=0$.

Computing determinants is usually difficult and many techniques have been devised to compute them out. As is evident from counting, computing
the determinant of an $n \times n$ matrix using the definition above would require the expression of all $n$ ! permutations of the integers $\{1,2, \ldots, n\}$ and the determination of their signs together with all the concommitant products and summation. This method is prohibitively costly. Using elementary row operations and Gaussian elimination, the evaluation of the determinant becomes more manageable. First we need the result below.

Theorem 2.5.1. For the elementary matrices the following results hold.
(a) for Type 1 (row interchange) $E_{1}$

$$
\operatorname{det} E_{1}=-1
$$

(b) for Type 2 (multiply a row by a constant c) $E_{2}$

$$
\operatorname{det} E_{2}=c
$$

(c) for Type 3 (add a multiple of one row to another row) $E_{3}$

$$
\operatorname{det} E_{3}=1
$$

Note that (c) is a consequence of Corollary 2.5.2. Proof of parts (a) and (b) are left as exercises. Thus for any matrix $A \in M_{n}(F)$ we have

$$
\begin{aligned}
\operatorname{det}\left(E_{1} A\right)=-\operatorname{det} A & =\operatorname{det} E_{1} \operatorname{det} A \\
\operatorname{det}\left(E_{2} A\right)=c \operatorname{det} A & =\operatorname{det} E_{2} \operatorname{det} A \\
\operatorname{det}\left(E_{3} A\right)=\operatorname{det} A & =\operatorname{det} E_{3} \operatorname{det} A .
\end{aligned}
$$

Suppose $F_{1} \ldots F_{k}$ is a sequence of row operations to reduce $A$ to its RREF. Then

$$
F_{k} F_{k-1} \ldots F_{1} A=B
$$

Now we see that

$$
\begin{aligned}
\operatorname{det} B & =\operatorname{det}\left(F_{k} F_{k-1} \ldots F_{1} A\right) \\
& =\operatorname{det}\left(F_{k}\right) \operatorname{det}\left(F_{k-1} \ldots F_{1} A\right) \\
& =\vdots \\
& =\operatorname{det}\left(F_{k}\right) \operatorname{det}\left(F_{k-1}\right) \ldots \operatorname{det}\left(F_{1}\right) \operatorname{det} A .
\end{aligned}
$$

For $B$ in RREF and $B \in M_{n}(F)$, we have that $B$ is upper triangular. The next result establishes Theorem 2.3.4(k) about the determinant of singular and non singular matrices. Moreover, the determinant of triangular matrices is computed simply as the product of its diagonal elements.

Proposition 2.5.3. Let $A \in M_{n}$. Then
(i) If $r(A)<n$, then $\operatorname{det} A=0$.
(ii) If $A$ is triangular then

$$
\operatorname{det} A=\prod a_{i i}
$$

(iii) If $r(A)=n$, then $\operatorname{det} A \neq 0$.

Proof. (i) If $r(A)<n$, then its RREF has a row of zeros, and $\operatorname{det} A=0$ by Theorem 2.5.1. (ii) If $A$ is triangular the only product without possible zero entries is $\prod a_{i i}$. Hence $\operatorname{det} A=\prod a_{i i}$. (iii) If If $r(A)=n$, then its RREF has no nonzero rows. Since it is square and has a leading one in each column, it follows that the RREF is the identity matrix. Therefore $\operatorname{det} A \neq 0$.

Now let $A, B \in M_{n}$. If $A$ is singular the RREF must have a zero row. It follows that $\operatorname{det} A=0$. If $A$ is singular it follows that $A B$ is singular. Therefore

$$
0=\operatorname{det} A B=\operatorname{det} A \operatorname{det} B .
$$

The same reasoning applies if $B$ is singular. If $A$ and $B$ are not singular both $A$ and $B$ can be row reduced to the identity. Let $F_{1} \ldots F_{k_{1}}$ be the row operations that reduce $A$ to $I$, and $G_{1} \ldots G_{k_{B}}$ be the row operations that reduce $B$ to $I$. Then

$$
\begin{aligned}
\operatorname{det} A & =\left[\operatorname{det}\left(F_{1}\right) \ldots \operatorname{det}\left(F_{k_{A}}\right)\right]^{-1} \\
\operatorname{det} B & =\left[\operatorname{det}\left(G_{1}\right) \ldots \operatorname{det}\left(G_{k_{B}}\right)\right]^{-1} .
\end{aligned}
$$

Also

$$
I=\left(G_{k_{B}} \ldots G_{1}\right)\left(F_{k_{B}} \ldots F_{1}\right) A B
$$

and we have

$$
\operatorname{det} I=(\operatorname{det} A)^{-1}(\operatorname{det} B)^{-1} \operatorname{det} A B
$$

This proves the
Theorem 2.5.2. If $A, B \in M_{n}(F)$, $\operatorname{det} A B=\operatorname{det} A \operatorname{det} B$.

### 2.5.1 Minors and Determinants

The method of row reduction is one of the simplest methods to compute the determinant of a matrix. Indeed, it is not necessary to use Type 2 elementary transformation. This results in the computing the determinant as the product of the diagonal elements of the resulting triangular matrix possibly multiplied by a minus sign. An alternate approach to computing determinants using minors is both interesting and useful. However, unless the matrix has some special form, it does not provide a computational alternative to row reduction.

Definition 2.5.2. Let $A \in M_{n}(C)$. For any row $i$ and column $j$ define the (ij)-minor of $A$ by

$$
M_{i j}=\left.\operatorname{det} A\right|_{\begin{array}{l}
i^{\text {th }} \text { row removed } \\
j^{\text {th }} \text { column removed }
\end{array}}
$$

The notation

$$
\left.\right|_{\mid} ^{i^{\text {th }} \text { row removed }} j^{\text {th } \text { column removed }}
$$

denotes the $(n-1) \times(n-1)$ matrix formed from $A$ by removing the $i^{\text {th }}$ row and $j^{\text {th }}$ column. With minors an alternative formulation of the determinant can be given. This method, while not of great value computationally, has some theoretical importance. For example, the inverse of a matrix can be expressed using minors. We begin by consideration the determinant.

Theorem 2.5.3. Let $A \in M_{n}(C)$. (i) Fix any row, say row $k$. The determinant of $A$ is given by

$$
\operatorname{det} A=\sum_{j=1}^{n} a_{k j}(-1)^{k+j} M_{k j}
$$

(ii) Fix any column, say column $m$. The determinant of $A$ is given by

$$
\operatorname{det} A=\sum_{j=1}^{n} a_{j m}(-1)^{m+j} M_{j m}
$$

Proof. (i) Suppose that $k=1$. Consider the quantity

$$
a_{11} M_{11}
$$

We observe that this is equivalent to all the products of the form

$$
\operatorname{sgn}(\sigma) a_{11} \cdot a_{2 \sigma(2)} \cdots \cdots a_{n \sigma(n)}
$$

where only permutations that fix the integer (i.e. position) 1 are taken. Thus $\sigma(1)=1$. Since this position is fixed the signs taken in the determinant $M_{11}$ for permutations of $n-1$ integers are respectively the same as the signs for the new permutation of $n$ integers.

Now consider all permutations that fix the integer 2 in the sense that $\sigma(1)=2$. The quantity $a_{12}(-1)^{1+2} M_{12}$ consists of all the products of the form

$$
\operatorname{sgn}(\sigma) a_{12} \cdot a_{2 \sigma(1)} a_{3 \sigma(3)} \cdots \cdots a_{n \sigma(n)}
$$

We need here the extra sign change because if the part of the permutation $\sigma$ of the integers $\{1,3,4, \ldots, n\}$ is of one sign, which is the sign used in the computation of $\operatorname{det} M_{i j}$, then the permutation of $\sigma$ of the integers $\{1,2,3,4, \ldots, n\}$ is of the other sign, and that sign is $\operatorname{sgn}(\sigma)$.

When we proceed to the $k^{\text {th }}$ component, we consider permutations that fix the integer $k$. That is, $\sigma(1)=k$. In this case the quantity $a_{1 k}(-1)^{1+k} M_{1 k}$ consists of all products of the form

$$
\operatorname{sgn}(\sigma) a_{1 k} a_{2 \sigma(1)} \cdots a_{k-1 \sigma(k-1)} a_{k+1 \sigma(k+1)} \cdots \cdots a_{n \sigma(n)}
$$

Continuing in this way we exhaust all possible products $a_{1 \sigma(1)} \cdot a_{2 \sigma(2)} \ldots$. $a_{n \sigma(n)}$ over all possible permutations of the integers $\{1,2, \ldots, n\}$. This proves the assertion. The proof for expanding from any row is similar, with only a possible change of sign needed, which is a prescribed.
(ii) The proof is similar.

Example 2.5.2. Find the determinant of

$$
A=\left[\begin{array}{ccc}
3 & 2 & -1 \\
0 & 1 & 3 \\
1 & 2 & -1
\end{array}\right]
$$

expanding across the first row and then expanding down the second column.
Solution. Expanding across the first row gives

$$
\begin{aligned}
\operatorname{det} A & =a_{11} M_{11}-a_{12} M_{12}+a_{13} M_{12} \\
& =3 \operatorname{det}\left[\begin{array}{cc}
1 & 3 \\
2 & -1
\end{array}\right]-2 \operatorname{det}\left[\begin{array}{cc}
0 & 3 \\
1 & -1
\end{array}\right]-1 \operatorname{det}\left[\begin{array}{ll}
0 & 1 \\
1 & 2
\end{array}\right] \\
& =3(-7)-2(-3)-(-1) \\
& =-14
\end{aligned}
$$

Expanding across the second column gives

$$
\begin{aligned}
\operatorname{det} A & =-2 \operatorname{det}\left[\begin{array}{cc}
0 & 3 \\
1 & -1
\end{array}\right]+1 \operatorname{det}\left[\begin{array}{ll}
3 & -1 \\
1 & -1
\end{array}\right]-2 \operatorname{det}\left[\begin{array}{cc}
3 & -1 \\
0 & 3
\end{array}\right] \\
& =-2(-3)+(-2)-2(9)=-14
\end{aligned}
$$

The inverse of the matrix can be formulated in terms of minors, which is formulated below.

Definition 2.5.3. Let $A \in M_{n}(\mathbb{C})$ (or $M_{n}(R)$ ). Define the adjugate (or adjoint) matrix $\hat{A}$ by

$$
\hat{A}_{i j}=(-1)^{i+j} M_{j i}
$$

where $M_{j i}$ is the $j i$ minor.
The adjugate has traditionally been called the "adjoint", but that terminology is somewhat ambiguous in light of the previous definition as complex conjugate transpose. Note that it is defined for all square matrices; when restricted to invertible matrices the inverse appears.

Theorem 2.5.4. Let $A \in M_{n}(\mathbb{C})$ (or $M_{n}(R)$ ) be invertible. Then $A^{-1}=$ $\frac{1}{\operatorname{det} A} \hat{A}$
Proof. A quick examination of the $i j$-entry of the product $A \hat{A}$ yields the following sum

$$
\sum_{j=1}^{n} a_{i j} \hat{A}_{j k}=\frac{1}{\operatorname{det}(A)} \sum_{j=1}^{n} a_{i j}(-1)^{k+j} M_{k j}
$$

There are two possibilities. (1) If $i=k$, then the summation above is the summation to form the determinant as described in Theorem 2.5.3. (2) If $i \neq k$, the summation is the computation of the determinant of the matrix $A$ with the $k^{\text {th }}$ row replaced by the $i^{\text {th }}$ row. Thus the determinant of a matrix with two identical rows is represented above and this must be zero. We conclude that $(A \hat{A})_{i j}=\delta_{i j}$, the usual Kronecker 'delta,' and the result is proved.

Example 2.5.1. Find the adjugate and inverse of

$$
A=\left[\begin{array}{ll}
2 & 4 \\
2 & 1
\end{array}\right]
$$

It is easy to see that

$$
\hat{A}=\left[\begin{array}{cc}
1 & -4 \\
-2 & 2
\end{array}\right]
$$

Also $\operatorname{det} A=-6$. Therefore, the inverse

$$
A^{-1}=-\frac{1}{6}\left[\begin{array}{cc}
1 & -4 \\
-2 & 2
\end{array}\right]
$$

Remark 2.5.1. The notation for cofactors of a square matrix $A$ is often used

$$
\hat{a}_{i j}=(-1)^{i+j} M_{j i}
$$

Note the reversed order of the subscripts $i j$ and then $j i$ above.

## Cramer's Rule

We know now that the solution to the system $A x=b$ is given by $x=A^{-1} b$. Moreover, the inverse $A^{-1}$ is given by $A^{-1}=\frac{\hat{A}}{\operatorname{det} A}$, where $\hat{A}$ is the adjugate matrix. The the $i^{\text {th }}$ component of the solution vector is therefore

$$
\begin{aligned}
x_{i} & =\frac{1}{\operatorname{det} A} \sum_{j=1}^{n} \hat{a}_{i j} b_{j} \\
& =\frac{1}{\operatorname{det} A} \sum_{j=1}^{n}(-1)^{i+j} M_{j i} b_{j} \\
& =\frac{\operatorname{det} A_{i}}{\operatorname{det} A}
\end{aligned}
$$

where we define the matrix $A_{i}$ to be the modification to $A$ by replacing its $i^{\text {th }}$ column by the vector $b$. In this way we obtain a very compact formula for the solution of a linear system. Called Cramer's rule we state this conclusion as

Theorem 2.5.1. (Cramer's Rule.) Let $A \in M_{n}(\mathbb{C})$ be invertible and $b \in$ $C_{n}$. For each $i=1,, n$, define the matrix $A_{i}$ to be the modification of $A$ by replacing its $i^{\text {th }}$ column by the vector $b$. Then the solution to the linear system $A x=b$ is given by components $x_{i}=\frac{\operatorname{det} A_{i}}{\operatorname{det} A}, i=1, n$.

Example 2.5.2. Given the matrix $A=\left[\begin{array}{ll}2 & 4 \\ 2 & 1\end{array}\right]$, and the vector $b=$ $\left[\begin{array}{c}2 \\ -1\end{array}\right]$. Solve the system $A x=b$ by Cramer's rule.

We have

$$
A_{1}=\left[\begin{array}{cc}
2 & 4 \\
-1 & 1
\end{array}\right] \quad \text { and } \quad A_{2}=\left[\begin{array}{cc}
2 & 2 \\
2 & -1
\end{array}\right]
$$

and $\operatorname{det} A_{1}=6, \operatorname{det} A_{2}=-6, \operatorname{det} A=-6$. Therefore

$$
x_{1}=-1 \quad \text { and } \quad x_{2}=1
$$

## A curious formula

The useful formula using cofactors given below will have some consequence when we study positive definite operators in Chapter ??.

Proposition 2.5.1. Consider the matrix

$$
B=\left[\begin{array}{ccccc}
0 & x_{1} & x_{2} & \cdots & x_{n} \\
x_{1} & a_{11} & a_{12} & \cdots & a_{1 n} \\
x_{2} & a_{21} & a_{22} & \cdots & a_{2 n} \\
\vdots & \vdots & \vdots & \ddots & \vdots \\
x_{n} & a_{n 1} & a_{n 2} & & a_{n n}
\end{array}\right]
$$

Then

$$
\operatorname{det} B=-\sum \hat{a}_{i j} x_{i} x_{j}
$$

where $\hat{a}_{i j}$ is the ij-cofactor of $A$.
Proof. Expand by minors along the top row to get

$$
\operatorname{det} B=\sum(-1)^{j} x_{j} M_{1 j}(B)
$$

Now expand the matrix of $M_{1 j}(B)$ down the first column. This gives

$$
M_{1 j}(B)=\sum(-1)^{i-1} x_{i} M_{i j}(A)
$$

Combining we obtain

$$
\begin{aligned}
\operatorname{det} B & =\sum(-1)^{j} x_{j} M_{1 j}(B) \\
& =\sum(-1)^{j} x_{j}\left[\sum(-1)^{i-1} x_{i} M_{i j}(A)\right] \\
& =\sum \sum(-1)^{i+j-1} x_{j} x_{i} M_{i j}(A) \\
& =-\sum \sum \hat{a}_{i j} x_{j} x_{i}
\end{aligned}
$$

The reader may note that in the last line of the equation above, we should have used $\hat{a}_{i j}$. However, the formulation given is correct, as well. (Why?)

### 2.6 Partitioned Matrices

It is convenient to study partitioned or "blocked" matrices, or more graphically said, matrices whose entries are themselves matrices. For example, with $I_{2}$ denoting the $2 \times 2$ identity matrix we can create the $4 \times 4$ matrix written in partitioned form and expanded form.

$$
A=\left[\begin{array}{ll}
a I_{2} & c I_{2} \\
c I_{2} & d I_{2}
\end{array}\right]=\left[\begin{array}{llll}
a & 0 & b & 0 \\
0 & a & 0 & b \\
c & 0 & d & 0 \\
0 & c & 0 & d
\end{array}\right]
$$

Partitioning matrices allows our attention to focus on certain structural properties. In many applications partititioned matrices appear in a natural way, with the particular blocks having some system context. Many similar subclasses and processes apply to partitioned matrices. In specific situations they can be added, multiplied, and inverted, just like regular matrices. It is even possible to perform "blocked" version of Gaussian elimination. In the few results here, we touch on some of these possibilities.

Definition 2.6.1. For each $1 \leq i \leq m$ and $1 \leq j \leq n$, let $A_{i j}$ be an $m_{i} \times n_{j}$ matrices where . Then the matrix

$$
A=\left[\begin{array}{cccc}
A_{11} & A_{12} & \cdots & A_{1 n} \\
A_{21} & A_{2 n} & \cdots & A_{2 n} \\
\vdots & \vdots & \ddots & \vdots \\
A_{m 1} & A_{m 2} & \cdots & A_{m n}
\end{array}\right]
$$

is a partitioned matrix of order $\left(\sum m\right)_{i} \times\left(\sum n_{j}\right)$.
The usual operations of addition and multiplication of partitioned matrices can be performed provided each of the operations makes sense. For addition of two partitioned matrices $A$ and $B$ it is necessary to have the same numbers of blocks of the respective same sizes. Then

$$
\begin{aligned}
A+B & =\left[\begin{array}{cccc}
A_{11} & A_{12} & \cdots & A_{1 n} \\
A_{21} & A_{2 n} & \cdots & A_{2 n} \\
\vdots & \vdots & \ddots & \vdots \\
A_{m 1} & A_{m 2} & \cdots & A_{m n}
\end{array}\right]+\left[\begin{array}{cccc}
B_{11} & B_{12} & \cdots & B_{1 n} \\
B_{21} & B_{2 n} & \cdots & B_{2 n} \\
\vdots & \vdots & \ddots & \vdots \\
B_{m 1} & B_{m 2} & \cdots & B_{m n}
\end{array}\right] \\
& =\left[\begin{array}{cccc}
A_{11}+B_{11} & A_{12}+B_{12} & \cdots & A_{1 n}+B_{1 n} \\
A_{21}+B_{21} & A_{2 n}+B_{2 n} & \cdots & A_{2 n}+B_{2 n} \\
\vdots & \vdots & \ddots & \vdots \\
A_{m 1}+B_{m 1} & A_{m 2}+B_{m 2} & \cdots & A_{m n}+B_{m n}
\end{array}\right]
\end{aligned}
$$

For multiplication, the situation is a bit more complicated. For definiteness, suppose that $B$ is a partitioned matrix with block sizes $s_{i} \times t_{j}$, where $1 \leq$ $i \leq p$ and $1 \leq j \leq q$ The usual operations to construct $C=A B$,

$$
\sum_{j=1}^{n} A_{i j} B_{j k}
$$

then make sense provided $p=n$ and $n_{j}=s_{j}, 1 \leq j \leq n$.
A special category of partitioned matrices are the so-called quasi-triangular matrices, wherein $A_{i j}=0$ if $i>j$ for the "lower" triangular version. The special subclass of quasi-triangular matrices wherein $A_{i j}=0$ if $i \neq j$ are called quasi-diagonal. In the case of the multiplication of partitioned matrices $(C=A B)$ with the left multiplicand $A$ a quasi-diagonal matrix, we have $C_{i k}=A_{i i} B_{i k}$. Thus the multiplication is similar in form to the usual multiplication of matrices where the left multiplicand is a diagonal matrix. In the case of the multiplication of partitioned matrices $(C=A B)$ with the right multiplicand $B$ a quasi-diagonal matrix, we have $C_{i k}=A_{i k} B_{k k}$. For quasi-triangular matrices with square diagonal blocks, there is an interesting result about the determinant.

Theorem 2.6.1. Let $A$ be a quasi-triangular matrix, where the diagonal blocks $A_{i i}$ are square. Then

$$
\operatorname{det} A=\prod_{i} \operatorname{det} A_{i i}
$$

Proof. Apply row operations on each vertical block without row interchanges between blocks, without any Type 2 operations. The resulting matrix in each diagonal block position $(i, i)$ is triangular. Be sure to multiply one of the diagonal entries by $\pm 1$, reflecting the number of row interchanges within a block. The resulting matrix can still be regarded as partitioned, though the diagonal blocks are now actually upper triangular. Now apply Proposition 2.5.3, noting that the product of each of the diagonal entries pertaining to the $i^{\text {th }}$ block is in fact $\operatorname{det} A_{i i}$.

A simple consequence of this result, proved alá Gaussian elimination, is contained in the following corollary.

Corollary 2.6.1. Consider the partitioned matrix

$$
A=\left[\begin{array}{ll}
A_{11} & A_{12} \\
A_{21} & A_{22}
\end{array}\right]
$$

with square diagonal blocks and with $A_{11}$ invertible. Then the rank of $A$ is the same as the rank of $A_{11}$ if and only if $A_{22}=A_{21} A_{11}^{-1} A_{12}$.
Proof. Multiplication of $A$ by the elementary partitioned matrix

$$
E=\left[\begin{array}{cc}
I & 0 \\
-A_{21} A_{11}^{-1} & I
\end{array}\right]
$$

yields

$$
\begin{aligned}
E A & =\left[\begin{array}{cc}
I & 0 \\
-A_{21} A_{11}^{-1} & I
\end{array}\right]\left[\begin{array}{ll}
A_{11} & A_{12} \\
A_{21} & A_{22}
\end{array}\right] \\
& =\left[\begin{array}{cc}
A_{11} & A_{12} \\
0 & A_{22}-A_{21} A_{11}^{-1} A_{12}
\end{array}\right]
\end{aligned}
$$

Since $E$ has full rank, it follows that $\operatorname{rank}(E A)=\operatorname{rank} A$. Since $E A$ is quasi-triangular, it follows that the rank of $A$ is the same as the rank of $A_{11}$ if and only if $A_{22}-A_{21} A_{11}^{-1} A_{12}=0$.

### 2.7 Linear Transformations

Definition 2.7.1. A mapping $T$ from $R_{n}$ to $R_{m}$ is called a linear transformation if

$$
\begin{aligned}
T(x+y) & =T x+T y & & \forall x, y \in R_{n} \\
T(a x) & =a T x & & \forall a \in F .
\end{aligned}
$$

Note: We normally write $T x$ instead of $T(x)$.
Example 2.7.1. $T: R_{n} \rightarrow R_{m}$. Let $a \in R_{m}$ and $y \in R_{n}$. Then for each $x \in R_{n}, T x=\langle x, y\rangle a$ is a linear transformation. Let $S=\left\{v_{1} \ldots v_{n}\right\}$ be a basis of $R_{n}$, and define the $m \times n$ matrix with columns given by the coordinates of $T v_{1}, T v_{2}, \ldots, T v_{n}$. Then this matrix

$$
A=\left[\begin{array}{cccc}
T v_{1} & T v_{2} & & T v_{n} \\
\downarrow & \downarrow & \therefore & \downarrow
\end{array}\right]
$$

is the matrix representation of $T$ with respect to the basis $S$. Thus, if $x=\Sigma a_{i} v_{i}$, whence $[x]_{S}=\left(a_{1} \ldots a_{n}\right)$, we have

$$
[T x]_{S}=A[x]_{S}
$$

* There is a duality between all linear transformations from $R_{n}$ to $R_{m}$ and the set $M_{m, n}(F)$.

Note that $M_{m, n}(F)$ is itself a vector space over $F$. Hence $\mathcal{L}\left(F_{n}, F_{m}\right)$, the set of linear transformations from $F_{n}$ to $F_{m}$ is likewise. As such it has subspaces.

Example 2.7.2. (1) Let $\bar{x} \in F_{n}$. Definite $\mathcal{J}=\{T \in \mathcal{L} \mid T \bar{x}=0\}$. Then $\mathcal{J}$ is a subspace of $\mathcal{L}\left(F_{n}, F_{m}\right)$.
(2) Let $\mathcal{U}=\left\{T \in \mathcal{L}\left(R_{n}, R_{n}\right) \mid T x \geq 0\right.$ if $\left.x \geq 0\right\}$, where $\{x \geq 0\}$ means the positive orthant of $R_{n} . \mathcal{U}$ is not a linear subspace of $\mathcal{L}\left(R_{n}, R_{n}\right)$, though it is a convex set.
(3) Define $T: P_{n} \rightarrow P_{n}$ by $T p=\frac{d}{d x} p . T$ is a linear transformation.

Example 2.7.3. Express the linear transformation $D: P_{3} \rightarrow P_{3}$ given by $D p=\frac{d}{d x} p(x)$ as a matrix with respect to the basis. $S=\left\{1, x, x^{2}, x^{3}\right\}$. We have $D 1=0=0+0 x+0 x^{2}+0 x^{3}$. Also

$$
[D 1]_{S}=[0,0,0,0]^{T}
$$

similarly

$$
\begin{aligned}
{[D x]_{S} } & =[1,0,0,0]^{T} \\
{\left[D x^{2}\right]_{S} } & =[0,2,0,0]^{T} \\
{\left[D x^{3}\right]_{S} } & =[0,0,3,0]^{T} .
\end{aligned}
$$

Hence

$$
[D]_{S}=\left[\begin{array}{llll}
0 & 1 & 0 & 0 \\
0 & 0 & 2 & 0 \\
0 & 0 & 0 & 3 \\
0 & 0 & 0 & 0
\end{array}\right]
$$

In this context the differentiation operator is rather simple.
Example 2.7.4. Consider the linear transformation $T$ defined by $T q=$ $3 x \frac{d}{d x} q+x^{2} q$ for $q \in P_{2}$. Find the matrix representation of $T$.

Solution. First off we notice that this transformation has range in $P_{4}$. Let's use the standard bases for this problem. We then determine the coordinates of $T$ for vectors in the $P_{2}$ basis $\left\{1, x, x^{2}\right\}$ in the $P_{4}$ basis $\left\{1, x, x^{2}, x^{3}, x^{4}\right\}$. Compute

$$
\begin{aligned}
T(1) & =x^{2} \\
T(x) & =3 x+x^{3} \\
T\left(x^{2}\right) & =6 x^{2}+x^{4}
\end{aligned}
$$

The coordinates of the input vectors we know are $[1,0,0]^{T},[0,1,0]^{T}$, and $[0,0,1]^{T}$. For the output vectors the coordinates are $[0,0,1,0,0]^{T},[0,3,0,1,0]^{T}$, and $[0,0,6,0,1]^{T}$. So, with respect to these two bases, the matrix of the transformation is

$$
A=\left[\begin{array}{lll}
0 & 0 & 0 \\
0 & 3 & 0 \\
1 & 0 & 6 \\
0 & 1 & 0 \\
0 & 0 & 1
\end{array}\right]
$$

Observe that the dimensionality corresponds with the dimentionality of the respective spaces.

Example 2.7.5. Let $V=R_{2}$, with $S_{0}=\left\{v_{1}, v_{2}\right\}=\left\{\left[\begin{array}{l}1 \\ 0\end{array}\right],\left[\begin{array}{l}1 \\ 1\end{array}\right]\right\}, S_{1}=$ $\left\{w_{1}, w_{2}\right\}=\left\{\left[\begin{array}{c}1 \\ 2\end{array}\right],\left[\begin{array}{c}-2 \\ 1\end{array}\right]\right\}$, and $T=I$ the identity. The vectors above are expressed in the standard $E=\left\{e_{1}, e_{2}\right\}, T v_{j}=I v_{j}=v_{j}$. To find $\left[v_{j}\right]_{S_{1}}$ we solve

$$
v_{j}=a_{j} w_{1}+\beta_{j} w_{2}
$$

$$
\begin{aligned}
& v_{1}:\left[\begin{array}{cc}
1 & -2 \\
2 & 1
\end{array}\right]\left[\begin{array}{l}
\alpha_{1} \\
\beta_{1}
\end{array}\right]=\left[\begin{array}{l}
1 \\
0
\end{array}\right] \rightarrow\left[\begin{array}{l}
\alpha_{1} \\
\beta_{1}
\end{array}\right]=\left[\begin{array}{c}
\frac{1}{5} \\
-\frac{2}{5}
\end{array}\right] \quad \nwarrow \text { solve linear } \\
& v_{2}:\left[\begin{array}{cc}
1 & -2 \\
2 & 1
\end{array}\right]\left[\begin{array}{l}
\alpha_{2} \\
\beta_{2}
\end{array}\right]=\left[\begin{array}{l}
1 \\
1
\end{array}\right] \rightarrow\left[\begin{array}{l}
\alpha_{2} \\
\beta_{2}
\end{array}\right]=\left[\begin{array}{c}
\frac{3}{5} \\
-\frac{1}{5}
\end{array}\right]
\end{aligned}
$$

Therefore

$$
S_{1}[I]_{S_{0}}=\left[\begin{array}{cc}
\frac{1}{5} & \frac{3}{5} \\
-\frac{2}{5} & -\frac{1}{5}
\end{array}\right] \leftarrow \begin{aligned}
& \text { change of } \\
& \text { basis } \\
& \text { matrix }
\end{aligned}
$$

If

$$
\begin{aligned}
{[x]_{S_{0}}=\left[\begin{array}{c}
-1 \\
2
\end{array}\right] \quad[x]_{S_{1}} } & =S_{1}[I]_{S_{0}}\left[\begin{array}{c}
-1 \\
2
\end{array}\right] \\
& =\frac{1}{5}\left[\begin{array}{cc}
1 & 3 \\
-2 & -1
\end{array}\right]\left[\begin{array}{c}
-1 \\
2
\end{array}\right]=\frac{1}{5}\left[\begin{array}{l}
5 \\
0
\end{array}\right]=\left[\begin{array}{l}
1 \\
0
\end{array}\right]
\end{aligned}
$$

Note the necessity of using the standard basis to express the vectors in both bases $S_{0}$ and $S_{1}$.

### 2.8 Change of Basis

Let $V$ be a vector space with bases $S_{0}=\left\{v_{1} \ldots v_{n}\right\}$ and $S_{1}=\left\{w_{1} \ldots w_{n}\right\}$, and suppose $T: V \rightarrow V$ is a linear transformation. We want to find the representation of $T$ as a matrix that takes a vector $x$ given in terms of its $S_{0}$ coordinates and produces the vector $T x$ given in terms of its $S_{1}$ coordinates.

We know that $x \rightarrow[x]_{S_{0}}$ is well defined. The action of $T$ is known if the $n$ vector $[x]_{S_{0}}=\left[\begin{array}{c}c_{1} \\ \vdots \\ c_{n}\end{array}\right]$ and the vectors $T v_{1}, T v_{2}, \ldots, T v_{n}$ are known, for if $x=\Sigma c_{j} v_{j}$, then $T x=\Sigma c_{j} T v_{j}$, by linearity.

To determine $[T x]_{S_{1}}$ we need to convert the $T v_{j}, j=1, \ldots, n$ to coordinates in the other $S_{1}$ basis, This is done as follows. Find

$$
\left[T v_{j}\right]_{S_{1}}=\left[\begin{array}{c}
t_{1 j} \\
t_{2 j} \\
\vdots \\
t_{n j}
\end{array}\right] \quad j=1,2, \ldots, n
$$

Then if $x \in V$

$$
\begin{aligned}
{[T x]_{S_{1}} } & =\left[\Sigma c_{j} T v_{j}\right]_{S_{1}}=\Sigma c_{j}\left[T v_{j}\right]_{S_{1}} \\
& =\left[\sum_{j} t_{i j} c_{j}\right] \\
& =\left[\begin{array}{ccc}
t_{11} & \cdots & t_{1 n} \\
t_{n 1} & & t_{n n}
\end{array}\right]\left[\begin{array}{c}
c_{1} \\
\vdots \\
c_{n}
\end{array}\right] .
\end{aligned}
$$

This $n \times n$ array $\left[t_{i j}\right]$ depends on $T, S_{0}$ and $S_{1}$ but not on $x$. We define the $S_{0} \rightarrow S_{1}$ basis representation of $T$ to be $\left[t_{i j}\right]$, and we write this as

$$
{ }_{S_{1}}[T]_{S_{0}}=\left[\begin{array}{ccc}
t_{11} & \ldots & t_{1 n} \\
\vdots & \ddots & \vdots \\
t_{n 1} & \cdots & t_{n n}
\end{array}\right]
$$

In the special case that $T$ is the identity operator the matrix ${ }_{S_{1}}[I]_{S_{0}}$ converts the coordinates of a vector in the basis $S_{0}$ to coordinates in the basis $S_{1}$. It is easy to see that ${ }_{S_{0}}[I]_{S_{1}}$ must be the inverse of $S_{S_{1}}[I]_{S_{0}}$ and thus

$$
S_{0}[I]_{S_{1}} \cdot S_{1}[I]_{S_{0}}=I .
$$

We can also establish the equality

$$
S_{1}[T]_{S_{1}}={ }_{S_{1}}[I]_{S_{0} S_{0}}[T]_{S_{0} S_{0}}[I]_{S_{1}} .
$$

In this way we see that the matrix representation of $T$ depends on the bases involved. If $X$ is any invertible matrix in $M_{n}(F)$ we can write

$$
B=X^{-1} A X
$$

The interpretation in this context is clear
$X$ : change of coordinate from one basis to another $S_{0} \rightarrow S_{1}$
$X^{-1}$ : change of coordinate $S_{1} \rightarrow S_{0}$
$A$ : matrix of the linear transformation in the basis $S_{0}$
$B$ : matrix of the same linear transformation in the basis $S_{1}$.
With this in mind it seems prudent to study linear transformations in the basis that makes their matrix representation as simple as possible.

Example 2.8.1. Let $A=\left[\begin{array}{cc}1 & 3 \\ -1 & 1\end{array}\right]$ be the matrix representation of a linear transformation given with respect to the standard basis $S_{0}=\left\{e_{1}, e_{2}\right\}=$ $\{(1,0),(0,1)\}$ Find the matrix representation of this transformation with resepect to the basis $S_{1}=\left\{v_{1}, v_{2}\right\}=\{(2,1),(1,1)\}$.
Solution. According to the analysis above we need to determine ${ }_{S_{0}}[I]_{S_{1}}$ and $S_{1}[I]_{S_{0}}$. Of course $S_{1}[I]_{S_{0}}=S_{0}[I]_{S_{1}}^{-1}$. Since the coordinates of the vectors in $S_{1}$ are expressed in terms of the basis vectors $S_{0}$ we obtain directly

$$
S_{0}[I]_{S_{1}}=\left[\begin{array}{ll}
2 & 1 \\
1 & 1
\end{array}\right]
$$

Its inverse is given by

$$
S_{0}[I]_{S_{1}}^{-1}=\left[\begin{array}{cc}
1 & -1 \\
-1 & 2
\end{array}\right]
$$

Assembling these matrices we have the final matrix converted to the new basis.

$$
\begin{aligned}
S_{1}[A]_{S_{1}} & =S_{1}[I]_{S_{0}} A_{S_{0}}[I]_{S_{1}} \\
& =\left[\begin{array}{cc}
1 & -1 \\
-1 & 2
\end{array}\right]\left[\begin{array}{cc}
1 & 3 \\
-1 & 1
\end{array}\right]\left[\begin{array}{ll}
2 & 1 \\
1 & 1
\end{array}\right] \\
& =\left[\begin{array}{cc}
6 & 4 \\
-7 & -4
\end{array}\right]
\end{aligned}
$$

Example 2.8.2. Consider the same problem as above except that the matrix $A$ is given in the basis $S_{1}$. Find matrix representation of this transformation with resepect to the basis $S_{0}$.
Solution. To solve this problem we need to determine ${ }_{S_{0}}[A]_{S_{0}}={ }_{S_{0}}[I]_{S_{1}} A_{S_{1}}[I]_{S_{0}}$. As we already have these matrices, we determine that

$$
\begin{aligned}
S_{0}[A]_{S_{0}} & ={ }_{S_{0}}[I]_{S_{1}} A_{S_{1}}[I]_{S_{0}} . \\
& =\left[\begin{array}{ll}
2 & 1 \\
1 & 1
\end{array}\right]\left[\begin{array}{cc}
1 & 3 \\
-1 & 1
\end{array}\right]\left[\begin{array}{cc}
1 & -1 \\
-1 & 2
\end{array}\right] \\
& =\left[\begin{array}{cc}
-6 & 13 \\
-4 & 8
\end{array}\right]
\end{aligned}
$$

### 2.9 Appendix A - Solving linear systems

The key to solving linear systems is to reduce the augmented system to RREF and solve the resulting equations. While this may be so, there is an intermediate step that occurs about half way through the computation of the RREF where the reduced matrix achieves an upper triangular form. At this point the solution can be determined directly by back substitution. To clarify the rules on back substitution, suppose that we have the triangular form

$$
\left[\begin{array}{cccc|c}
a_{11} & a_{12} & \cdots & a_{1 n} & b_{1} \\
0 & a_{22} & \cdots & a_{2 n} & b_{2} \\
\vdots & & \ddots & \vdots & \vdots \\
0 & \cdots & 0 & a_{n n} & b_{n}
\end{array}\right]
$$

Assuming that the diagonal part consists of all nonzero terms, we can solve this system by back substitution. First solve for $x_{n}=\frac{b_{n}}{a_{n n}}$. Now inductively solve for the remaining solution coordinates using the formula

$$
x_{n-j}=\frac{1}{b_{n-j, n-j}}\left[\sum_{k=0}^{j-1} a_{n-j, n-k} x_{n-k}\right], \quad j=1,2, \ldots, n-1
$$

This inconvenient looking formula can be replaced by

$$
x_{j}=\frac{1}{b_{j j}}\left[\sum_{k=j+1}^{n} a_{j k} x_{k}\right], \quad j=n-1, n-2, \ldots, 1
$$

where the index runs from $j=n-1$ up to $j=1$. The upshot is that the row reduction process can be halted when a triangular-like form has been attained. The applies as well to nonsingular and non square systems, where the the process is stopped when all the leading ones have been identified, entries below them have been zeroed out, and all the zero rows are present. The principle reason for using back substitution is to reduce the number of computations required, an important consideration in numerical linear algebra. In the example below we solve a $3 \times 3$ nonsingular system.

Example 2.9.1. Solve $A x=b$ where

$$
A=\left[\begin{array}{ccc}
1 & 2 & 0 \\
2 & 2 & -1 \\
-1 & 3 & 2
\end{array}\right] \quad b=\left[\begin{array}{c}
3 \\
6 \\
-2
\end{array}\right]
$$

Solution. Find the RREF of $[A \mid b]$. Then solve $A x=b$.

$$
\left.\left.\begin{array}{ccc|c}
{\left[\begin{array}{ccc|c}
1 & 2 & 0 & 3 \\
2 & 2 & -1 & 6 \\
-1 & 3 & 2 & -2
\end{array}\right]} & \begin{array}{c}
-2 R_{1}+R_{2} \\
R_{1}+R_{3}
\end{array} & {\left[\begin{array}{ccc|c}
1 & 2 & 0 & 3 \\
0 & -2 & -1 & 0 \\
0 & 5 & 2 & 1
\end{array}\right]} \\
& -\frac{1}{2} R_{2} \\
& \rightarrow & {\left[\begin{array}{ccc|c}
1 & 2 & 0 & 3 \\
0 & 1 & \frac{1}{2} & 0 \\
0 & 5 & 2 & 1
\end{array}\right]}  \tag{*}\\
& \rightarrow \\
& \rightarrow-\frac{1}{2} R_{3}+R_{2}
\end{array}\right] \begin{array}{ccc|c}
1 & 2 & 0 & 3 \\
0 & 1 & 0 & 1 \\
0 & 0 & 1 & -2
\end{array}\right]
$$

Hence solving we obtain $x_{3}=-2, \quad x_{2}=1$, and $x_{1}=1$. This is fine, but there is a faster way to solve this system. Stop the reduction when the system attains a triangular form at $(*)$. From this point solve to obtain $x_{3}=-2$. Now back substitute $x_{3}=2$ into the second row (equation) to solve for $x_{2}$. Thus $x_{2}=-\frac{1}{2}(-2)=1$. Finally, back substitute $x_{3}=2$ and $x_{2}=1$ into the first row (equation) to solve for $x_{1}$. Thus $x_{1}=3-2(1)=1$. Sometimes the form (*) is called the row reduced form.

Example 2.9.2. Given the augmented system for $A x=b$ is in RREF.

$$
\left[\begin{array}{cccccc|c}
1 & 2 & 0 & 0 & 0 & 0 & 4 \\
0 & 0 & 1 & 2 & 0 & -1 & 1 \\
0 & 0 & 0 & 0 & 1 & 3 & 1 \\
0 & 0 & 0 & 0 & 0 & 0 & 0
\end{array}\right]
$$

Find the solution.
Solution. The leading ones occur in columns 1, 3, and 5. The values in columns 2,4 , and 6 can be taken as free parameters. So, take $x_{2}=r, x_{4}=s$, and $x_{6}=t$. Now solving for the other varables we have

$$
\begin{aligned}
& x_{1}=4-2 r \\
& x_{3}=1-2 s+t \\
& x_{5}=1-3 t
\end{aligned}
$$

The solution set is comprised of the vector

$$
\begin{aligned}
x & =[4-2 r, r, 1-2 s+t, s, 1-t, t]^{T} \\
& =[4,0,1,0,1,0]^{T}+r[-2,1,0,0,0,0]^{T}+s[0,0,-2,1,0,0]^{T}+t[0,0,1,0-3,1]^{T}
\end{aligned}
$$

for all $r, s$, and $t$. We can rewrite this as the set

$$
S=\left\{\left.\left[\begin{array}{l}
4 \\
0 \\
1 \\
0 \\
1 \\
0
\end{array}\right]+r\left[\begin{array}{c}
-2 \\
1 \\
0 \\
0 \\
0 \\
0
\end{array}\right]+s\left[\begin{array}{c}
0 \\
0 \\
-2 \\
1 \\
0 \\
0
\end{array}\right]+t\left[\begin{array}{c}
0 \\
0 \\
1 \\
0 \\
-3 \\
1
\end{array}\right] \right\rvert\, r, s, t \in R \text { or } \mathbb{C}\right\}
$$

This representation shows better the connection between the free constants and the component vectors that make up the solution. Note this expression also reveals the solution of the homogeneous solution $A x=0$ as the set

$$
\left\{r[-2,1,0,0,0,0]^{T}+s[0,0,-2,1,0,0]^{T}+t[0,0,1,0-3,1]^{T} \mid r, s, t \in R \text { or } \mathbb{C}\right\}
$$

Indeed, this is a full subspace.
Example 2.9.3. The RREF can be used to determine the inverse, as well. Given the matrix $A \in M_{n}$, the inverse is given by the matrix $X$ for which $A X=I$. In turn with $x_{1}, \ldots, x_{n}$ representing the columns of $X$ and $e_{1}, \ldots, e_{n}$ representing the standard vectors we see that $A x_{j}=$ $e_{j}, j=1,2, \ldots, n$. To solve for these vectors, form the augmented matrix $\left[A \mid e_{j}\right], j=1,2, \ldots, n$ and row reduce as above. A massive short cut to this process is to augment all the standard vectors at one and row reduce the
resulting $n \times 2 n$ matrix $[A \mid I]$. If $A$ is invertible, its RREF is the identity. Therefore,

$$
[A \mid I] \stackrel{\text { row }}{\rightarrow} \underset{\text { operations }}{ }[I \mid X]
$$

and, of course, $A^{-1}=X$. Thus, for

$$
A=\left[\begin{array}{ccc}
-2 & 1 & 0 \\
1 & 1 & 2 \\
3 & -2 & -1
\end{array}\right]
$$

we row reduce $[A \mid I]$ as follows

$$
\begin{aligned}
{[A \mid I]=} & {\left[\begin{array}{ccc|ccc}
-2 & 1 & 0 & 1 & 0 & 0 \\
1 & 1 & 2 & 0 & 1 & 0 \\
3 & -2 & -1 & 0 & 0 & 1
\end{array}\right] } \\
& \stackrel{\text { row }}{\rightarrow} \\
& \text { operations }
\end{aligned}\left[\begin{array}{ccc|ccc}
1 & 0 & 0 & 3 & 1 & 2 \\
0 & 1 & 0 & 7 & 2 & 4 \\
0 & 0 & 1 & -5 & -1 & -3
\end{array}\right]
$$

### 2.10 Exercises

1. Consider the differential operator $T=2 x \frac{d}{d x}(\cdot)-4$ acting on the vector space of cubic polynomials, $P_{3}$. Show that $T$ is a linear transformation and find a matrix representation of it. Assume the basis is given by $\left\{1, x, x^{2}, x^{3}\right\}$.
2. (i) Find matrices $A$ and $B$, each with positive rank, for which $r(A+$ $B)=r(A)+r(B)$. (ii) Find matrices $A$ and $B$, each with positive rank, for which $r(A+B)=0$. (iii) Give a method to find two nonzero matrices $A$ and $B$ for which the sum has any preassigned rank. Of course, the matrix sizes may depend on this value.
3. Find square matrices $A$ and $B$ for which $r(A)=r(B)=2$ and for which $r(A B)=0$.
4. Suppose that A is an $m \times n$ matrix and that $x$ is a solution of $A x=b$ over the prescribed field. Show that every solution of $A x=b$ have the form $x+x_{0}$, where $x_{0}$ is a solution of $A x_{0}=0$.
5. Find a matrix $A \in M_{n}$ of $\operatorname{rank} n-1$ for which $r\left(A^{k}\right)=n-k$ for $k \leq n$. Is it possible to begin this process with a matrix $A \in M_{n}$ of rank $n$ and for which $r\left(A^{k}\right)=n-k+1$ for $k \leq n$ ?
6. Show that $A x=b$ has a solution if and only if $y^{T} b=0$ if and only if $y^{T} A=0$ for some column vector.
7. In $R_{2}$ the linear transformation that rotates any vector by $\theta$ radians counter clockwise $T$ has matrix representation with respect to the standard basis given by

$$
A=\left[\begin{array}{cc}
\cos \theta & -\sin \theta \\
\sin \theta & \cos \theta
\end{array}\right]
$$

What is the matrix representation with respect to the standard basis of the transformation that rotates any vector by $\theta$ radians clockwise? What is the relation between the matrices?
8. Show that if $B, C \in M_{n}(F)$, where $B$ is symmetric and $C$ is skewsymmetric, then $B=C$ implies that $B=C=0$.
9. Prove the general formula for the inverse of the $2 \times 2$ matrix $A=$ $\left[\begin{array}{ll}a & b \\ c & d\end{array}\right]$ is $A^{-1}=\frac{1}{\operatorname{det} A}\left[\begin{array}{cc}d & -b \\ -c & a\end{array}\right]$.
10. Prove Theorem 2.5.1(a).
11. Prove Theorem 2.5.1(b).
12. Prove that every permutation $\sigma$ must have an inverse $\sigma^{-1}$ (i.e. $\sigma^{-1}(\sigma(j))=$ $j$ ), and the signs of $\sigma^{-1}$ and $\sigma$ are the same.
13. Show that the sign of every transposition is -1 .
14. Prove that $\operatorname{det} A=\sum_{\sigma}(-1)^{\operatorname{sgn}(\sigma)}\left(\prod_{i} a_{\sigma(i) i}\right)$
15. Prove Proposition 2.5.2 using minors.
16. Suppose that the $n \times n$ matrix $A$ is singular. Show that each column of the adjugate matrix $\hat{A}$ is a solution of $A x=0$. (McDuffee, Chapter 3, Theorem 29.)
17. Suppose that $A$ is an $(n-1) \times n$ matrix, and consider the homogeneous system $A x=0$ for $x \in R_{n}$. Define $h_{i}$ to be the determinant of the $(n-1) \times(n-1)$ matrix formed by removing the $i^{\text {th }}$ column of $A$. Show that the vector $h=\left(h_{1}, \ldots, h_{n}\right)^{T}$ is a solution to $A x=0$. (McDuffee, Chapter 3, Corollary 29.)
18. Show that $A=\left[\begin{array}{cc}1 & -1 \\ -1 & 1\end{array}\right]$ has no inverse by trying to solve $A B=I$ That is, assume the form

$$
B=\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]
$$

multiply the matrices ( $A$ and $B$ ) together, and then solve for the unknowns $a, b, c$, and $d$. (This is not a very efficient way to determine inverses of matrices. Try the same thing for any $3 \times 3$ matrix.)
19. Prove that the elementary equation operations do not change the solution set of a linear system.
20. Find the inverses of $E_{1}, E_{2}$, and $E_{3}$.
21. Find the matrix representation of linear transformation $T$ that rotates any vector by $\theta$ radians counter clockwise (ccw) with respect to the basis $S=\{(2,1),(1,1)\}$.
22. Consider $R_{3}$. Suppose that we have angles $\left\{\theta_{i}\right\}_{i=1}^{3}$ and pairs of coordinate vectors $\left\{\left(e_{1}, e_{2}\right),\left(e_{1}, e_{3}\right),\left(e_{2}, e_{3}\right)\right\}$. Let $T$ be the linear transformation that successively rotates a vector in the respective planes $\left\{\left(e_{i_{1}}, e_{i_{2}}\right)\right\}_{i=1}^{k}$ through the respective angles $\left\{\theta_{i}\right\}_{i=1}^{k}$. Find the matrix representation of $T$ with respect to the standard basis. Prove that it is invertible.

## 23. Prove Theorem 2.2.4.

24. Prove or disprove the equivalence of the linear systems.

$$
\begin{aligned}
2 x-3 y & =-1 & & -x+4 y=3 \\
x+4 y & =5 & & x+2 y=3
\end{aligned}
$$

25. Find basis for the orthocomplement of the subspace of $R_{3}$ spanned by the vectors $\left\{[2,1,1]^{T},[1,1,2]^{T}\right\}$.
26. Find basis for the orthocomplement of the subspace of $R_{3}$ spanned by the vector $[1,1,1]^{T}$.
27. Consider planar rotations in $R_{n}$ with respect to the standard bases elements. Prove that there must be $\frac{n(n-1)}{2}$ of them - discounting the particular angle. Display the general representation of any of them. Prove or disprove that any two of them are commutative. That is for two angles $\left\{\theta_{i}\right\}_{i=1}^{2}$ and pairs of coordinate vectors $\left\{\left(e_{i_{1}}, e_{i_{2}}\right)\right\}_{i=1}^{k}$ the respective counter clockwise rotations are commutative.
28. Suppose that $A \in M_{m k}, B \in M_{k n}$ and both have rank $k$. Show that the rank of $A B$ is $k$.
29. Suppose that $A \in M_{m k}$ has rank $k$. Prove that $A_{\text {RREF }}=\left[\begin{array}{c}I_{k} \\ 0\end{array}\right]$ where $I_{k}$ is the identity matrix of size $k$ and 0 is the $m-k \times k$ zero matrix.
30. Suppose that $B \in M_{k n}$ has rank $k$. Prove that $B_{\text {RREF }}=\left[\begin{array}{ll}I_{k} & 0\end{array}\right]$ where $I_{k}$ is the identity matrix of size $k$ and 0 is the $\times n-k$ zero matrix.
31. Determine and prove a version of Corollary 2.6 .1 for $3 \times 3$ blocked matrices, where we assume the diagonal blocks $A_{11}$ is invertible and wish to conclude the result that the rank of $A$ is the equal to the rank of $A_{11}$.
32. Suppose that we have angles $\left\{\theta_{i}\right\}_{i=1}^{k}$ and pairs of coordinate vectors $\left\{\left(e_{i_{1}}, e_{i_{2}}\right)\right\}_{i=1}^{k}$. Let $T$ be the linear transformation that successively rotates a vector in the respective planes $\left\{\left(e_{i_{1}}, e_{i_{2}}\right)\right\}_{i=1}^{k}$ through the respective angles $\left\{\theta_{i}\right\}_{i=1}^{k}$. Prove that the matrix representation of the linear transformation with respect to any basis must be invertible.
33. The super-diagonal of a matrix is the set of elements $a_{i, i+1}$. The subdiagonal of a matrix is the set of elements $a_{i-1, i}$. A tri-banded matrix is one for which the entries are zero above the super-diagonal and below the subdiagonal. Suppose that for an $n \times n$ tri-banded matrix $T$, we have $a_{i-1, i}=a, \quad a_{i i}=0$, and $a_{i, i+1}=c$. Prove the following facts:
(a) If $n$ is odd $\operatorname{det} A=0$.
(b) If $n=2 m$ is even $\operatorname{det} T=(-1)^{m} a^{m} c^{m}$.
34. For the banded matrix of the previous example, prove the following for the powers $T^{p}$ of $T$.
(a) If $p$ is odd, prove that $\left(T^{p}\right)_{i j}=0$ if $i+j$ is even.
(b) If $p$ is even, prove that $\left(T^{p}\right)_{i j}=0$ if $i+j$ is odd.
35. Consider the vector space $P_{2}(1,2)$ with inner product defined by $\langle p, q\rangle=$ $\int_{1}^{2} p(x) q(x) d x$. Find an orthogonal basis of $P_{2}(1,2)$. (Hint. Begin with the standard basis $\left\{1, x, x^{2}\right\}$. Apply the Gram-Schmidt procedure.)
36. For what values of $a$ and $b$ is the matrix below singular

$$
A=\left[\begin{array}{ccc}
a & 2 & 1 \\
2 & 1 & b \\
1 & a & -2
\end{array}\right]
$$

37. The Vandermonde matrix, defined for a sequence of numbers $\left\{x_{1}, \ldots x_{n}\right\}$, is given by the $n \times n$ matrix

$$
V_{n}=\left[\begin{array}{ccccc}
1 & x_{1} & x_{1}^{2} & \cdots & x_{1}^{n-1} \\
1 & x_{2} & x_{2}^{2} & \cdots & x_{2}^{n-1} \\
\vdots & \vdots & \vdots & \ddots & \vdots \\
1 & x_{n} & x_{n}^{2} & \cdots & x_{n}^{n-1}
\end{array}\right]
$$

Prove that the determinant is given by

$$
\operatorname{det} V_{n}=\prod_{i>j=1}^{n}\left(x_{i}-x_{j}\right)
$$

38. In the case the $x$-values are the integers $\{1, \ldots, n\}$, prove that $\operatorname{det} V_{n}$ is divisible by $\prod_{i=1}^{n}(i-1)!$. (These numbers are called superfactorials.)
39. Prove that for the weighted functional defined in Remark 2.4.1, it is necessary and sufficient that the weights be strictly positive for it to be an inner product.
40. For what values of $a$ and $b$ is the matrix below singular

$$
A=\left[\begin{array}{cccc}
b & 0 & a & 0 \\
0 & 0 & b & a \\
0 & a & 0 & b \\
b & a & 0 & 0
\end{array}\right]
$$

41. Find an orthogonal basis for $R_{2}$ from the vectors $\{(1,2),(2,1)\}$.
42. Find an orthogonal basis of the subspace of $R_{3}$ spanned by $\{(1,0,1),(0,1,-1)\}$
43. Suppose that $V$ is a vector space with an inner product, and $S \subset V$. Show that if $S$ is a basis of $V, S^{\perp}=\{0\}$.
44. Suppose that $V$ is a vector space with an inner product, and $S \subset V$. Show that if $U=\mathfrak{S}(S)$, then $U^{\perp}=S^{\perp}$.
45. Let $A \in M_{m n}(F)$. Show it may not be true that $r(A)=r\left(A^{T} A\right)=$ $r\left(A A^{T}\right)$ unless $F=R$, in which case it is true.
46. If $A \in M_{n}(C)$ is orthogonal, show that the rows and columns of $A$ are orthogonal.
47. If $A$ is orthogonal then $A^{m}$ is orthogonal for every positive integer $m$. (This is a part of Theorem 2.4.3(b).)
48. Consider the polynomial space $P_{n}[-1,1]$ with the inner product $\langle p, q\rangle=$ $\int_{-1}^{1} p(t) q(t) d t$. Show that every polynomial $p \in P_{n}$ for which $p(1)=$ $p(-1)=0$ is orthogonal to its derivative.
49. Consider the polynomial space $P_{n}[-1,1]$ with the inner product $\langle p, q\rangle=$ $\int_{-1}^{1} p(t) q(t) d t$. Show that the subspace of polynomials in even powers (e.g. $p(t)=t^{2}-5 t^{6}$ ) is orthogonal to the subspace of polynomials in odd powers.
50. Let $A=\left[\begin{array}{cc}1 & 3 \\ -1 & 1\end{array}\right]$ be the matrix representation of a linear transformation given with respect to the standard basis $S_{0}=\left\{e_{1}, e_{2}\right\}=$ $\{(1,0),(0,1)\}$ Find the matrix representation of this transformation with resepect to the basis $S_{1}=\left\{v_{1}, v_{2}\right\}=\{(2,-3),(1,-2)\}$.
51. Show that the sign of every transposition from the set $\{1,2, \ldots, n\}$ is -1 .
52. What are the signs of the permutations $\{7,6,5,4,3,2,1\}$ and $\{7,1,6,4,3,5,2\}$ of the integers $\{1,2,3,4,5,6,7\}$ ?
53. Prove that the sign of the permutation $\{m, m-1, \ldots, 2,1\}$ is $(-1)^{m}$.
54. Suppose $A, B \in M_{n}(F)$. If $A$ is singular, use a row space argument to show that $\operatorname{det} A B=0$.
55. Show that if $A \in M_{m, n}$ and $B \in M_{m, n}$ and both $r(A)=r(B)=m$. If $r(A B)=m-k$, what can be said about $n$ ?
56. Prove that

$$
\operatorname{det}\left|\begin{array}{cccc}
x_{1} & x_{2} & x_{3} & x_{4} \\
-x_{2} & x_{1} & -x_{4} & x_{3} \\
-x_{3} & x_{4} & x_{1} & -x_{2} \\
-x_{4} & -x_{3} & -x_{2} & x_{1}
\end{array}\right|=\left(x_{1}^{2}+x_{2}^{2}+x_{3}^{2}+x_{4}^{2}\right)^{2}
$$

57. Prove that there is no invertible $3 \times 3$ matrix that has all the same cofactors. What similar statement can be made for $n \times n$ matrices?
58. Show by example that there are matrices $A$ and $B$ for which $\lim _{n \rightarrow \infty} A^{n}$ and $\lim _{n \rightarrow \infty} B^{n}$ both exist, but for which $\lim _{n \rightarrow \infty}(A B)^{n}$ does not exist.
59. Let $A \in M_{2}(\mathbb{C})$. Show that there is no matrix solution $B \in M_{2}(\mathbb{C})$ to $A B-B A=I$. What can you say about the same problem with $A, B \in M_{n}(\mathbb{C})$ ?
60. Show by example that if $A C=B C$ then it does not follow that $A=B$. However, show that if $C$ is inveritble the conclusion $A=B$ is valid.
