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## Lecture 1

## 1 Review of Topology

### 1.1 Metric Spaces

Definition 1.1. Let $X$ be a set. Define the Cartesian product $X \times X=\{(x, y)$ : $x, y \in X\}$.

Definition 1.2. Let $d: X \times X \rightarrow \mathbb{R}$ be a mapping. The mapping $d$ is a metric on $X$ if the following four conditions hold for all $x, y, z \in X$ :
(i) $d(x, y)=d(y, x)$,
(ii) $d(x, y) \geq 0$,
(iii) $d(x, y)=0 \Longleftrightarrow x=y$, and
(iv) $d(x, z) \leq d(x, y)+d(y, z)$.

Given a metric $d$ on $X$, the pair $(X, d)$ is called a metric space.
Suppose $d$ is a metric on $X$ and that $Y \subseteq X$. Then there is an automatic metric $d_{Y}$ on $Y$ defined by restricting $d$ to the subspace $Y \times Y$,

$$
\begin{equation*}
d_{Y}=d \mid Y \times Y \tag{1.1}
\end{equation*}
$$

Together with $Y$, the metric $d_{Y}$ defines the automatic metric space $\left(Y, d_{Y}\right)$.

### 1.2 Open and Closed Sets

In this section we review some basic definitions and propositions in topology. We review open sets, closed sets, norms, continuity, and closure. Throughout this section, we let ( $X, d$ ) be a metric space unless otherwise specified.

One of the basic notions of topology is that of the open set. To define an open set, we first define the $\epsilon$-neighborhood.

Definition 1.3. Given a point $x_{o} \in X$, and a real number $\epsilon>0$, we define

$$
\begin{equation*}
U\left(x_{o}, \epsilon\right)=\left\{x \in X: d\left(x, x_{o}\right)<\epsilon\right\} . \tag{1.2}
\end{equation*}
$$

We call $U\left(x_{o}, \epsilon\right)$ the $\epsilon$-neighborhood of $x_{o}$ in $X$.
Given a subset $Y \subseteq X$, the $\epsilon$-neighborhood of $x_{o}$ in $Y$ is just $U\left(x_{o}, \epsilon\right) \cap Y$.
Definition 1.4. A subset $U$ of $X$ is open if for every $x_{o} \in U$ there exists a real number $\epsilon>0$ such that $U\left(x_{o}, \epsilon\right) \subseteq U$.

We make some propositions about the union and intersections of open sets. We omit the proofs, which are fairly straightforward.

The following Proposition states that arbitrary unions of open sets are open.
Proposition 1.5. Let $\left\{U_{\alpha}, \alpha \in I\right\}$ be a collection of open sets in $X$, where $I$ is just a labeling set that can be finite or infinite. Then, the set

$$
\bigcup_{\alpha \in I} U_{\alpha} \text { is open. }
$$

The following Corollary is an application of the above Proposition.
Corollary 1. If $Y \subset X$ and $A$ is open in $Y$ (w.r.t. $d_{Y}$ ), then there exists on open set $U$ in $X$ such that $U \cap Y=A$.

Proof. The set $A$ is open in $Y$. So, for any $p \in A$ there exists an $\epsilon_{p}>0$ such that $U\left(p, \epsilon_{p}\right) \cap Y \subseteq A$. We construct a set $U$ containing $A$ by taking the union of the sets $U\left(p, \epsilon_{p}\right)$ over all $p$ in $A$,

$$
\begin{equation*}
U=\bigcup_{p \in A} U\left(p, \epsilon_{p}\right) \tag{1.3}
\end{equation*}
$$

For every $p \in A$, we have $U\left(p, \epsilon_{p}\right) \cap Y \subseteq A$, which shows that $U \cap Y \subseteq A$. Furthermore, the union is over all $p \in A$, so $A \subseteq U$, which implies that $A \subseteq U \cap Y$. This shows that $U \cap Y=A$. To conclude the proof, we see that $U$ is open by the openness of the $U\left(p, \epsilon_{p}\right)$ and the above theorem.

The following Proposition states that finite intersections of open sets are open.
Proposition 1.6. Let $\left\{U_{i}, i=1, \ldots, N\right\}$ be a finite collection of open sets in $X$. Then the set

$$
\bigcap_{i=1}^{i=N} U_{i} \text { is open. }
$$

Definition 1.7. Define the complement of $A$ in $X$ to be $A^{c}=X-A=\{x \in X$ : $x \notin A\}$.

We use the complement to define closed sets.
Definition 1.8. The set $A$ is closed in $X$ if $A^{c}$ is open in $X$.

### 1.3 Metrics on $\mathbb{R}^{n}$

For most of this course, we will only consider the case $X=\mathbb{R}^{n}$ or $X$ equals certain subsets of $\mathbb{R}^{n}$ called manifolds, which we will define later.

There are two interesting metrics on $\mathbb{R}^{n}$. They are the Euclidean metric and the sup metric, and are defined in terms of the Euclidean norm and the sup norm, respectively.

Definition 1.9. Let $x \in \mathbb{R}^{n}$, written out in component form as $x=\left(x_{1}, x_{2}, \ldots, x_{n}\right)$. The Euclidean norm of $x$ is

$$
\|x\|=\sqrt{x_{1}^{2}+\cdots+x_{n}^{2}}
$$

and the the sup norm of $x$ is

$$
|x|=\max _{i}\left|x_{i}\right| .
$$

From these norms we obtain the Euclidean distance function

$$
\begin{equation*}
\|x-y\| \tag{1.4}
\end{equation*}
$$

and the sup distance function

$$
\begin{equation*}
|x-y| \tag{1.5}
\end{equation*}
$$

respectively.
These two distance functions are related in several ways. In particular,

$$
|x-y| \leq\|x-y\| \leq \sqrt{n}|x-y| .
$$

These distance functions are also related by the following Proposition, which will sometimes come in handy.

Proposition 1.10. A subset $U$ of $\mathbb{R}^{n}$ is open w.r.t. the $\|\|$ distance function if and only if it is open w.r.t. the $|\mid$ distance function.

So, these two distance functions give the same topologies of $\mathbb{R}^{n}$.

### 1.4 Continuity

Consider two metric spaces $\left(X, d_{X}\right)$ and $\left(Y, d_{Y}\right)$, a function $f: X \rightarrow Y$, and a point $x_{o} \in X$.

Definition 1.11. The function $f$ is continuous at $x_{o}$ if for every $\epsilon>0$ there exists a $\delta>0$ such that

$$
\begin{equation*}
d_{X}\left(x, x_{o}\right)<\delta \Longrightarrow d_{Y}\left(f(x), f\left(x_{o}\right)\right)<\epsilon . \tag{1.6}
\end{equation*}
$$

By definition, a function is continuous if it is continuous at all points in its domain.
Definition 1.12. The function $f$ is continuous if $f$ is continuous at every point $x_{o} \in X$.

There is an alternative formulation of continuity that we present here as a theorem.
Theorem 1.13. The function $f$ is continuous if and only if for every open subset $U$ of $Y$, the pre-image $f^{-1}(U)$ is open in $X$.

Continuous functions can often be combined to construct other continuous functions. For example, if $f, g: X \rightarrow \mathbb{R}$ are continuous functions, then $f+g$ and $f g$ are continuous functions.

### 1.5 Limit Points and Closure

As usual, let $(X, d)$ be a metric space.
Definition 1.14. Suppose that $A \subseteq X$. The point $x_{o} \in X$ is a limit point of $A$ if for every $\epsilon$-neighborhood $U\left(x_{o}, \epsilon\right)$ of $x_{o}$, the set $U\left(x_{o}, \epsilon\right)$ is an infinite set.

Definition 1.15. The closure of $A$, denoted by $\bar{A}$, is the union of $A$ and the set of limit points of $A$,

$$
\begin{equation*}
\bar{A}=A \cup\left\{x_{o} \in X: x_{o} \text { is a limit point of } A\right\} . \tag{1.7}
\end{equation*}
$$

Now we define the interior, exterior, and the boundary of a set in terms of open sets. In the following, we denote the complement of $A$ by $A^{c}=X-A$.

Definition 1.16. The set

$$
\begin{equation*}
\operatorname{Int} A \equiv\left(\bar{A}^{c}\right)^{c} \tag{1.8}
\end{equation*}
$$

is called the interior of $A$.
It follows that

$$
\begin{equation*}
x \in \operatorname{Int} A \Longleftrightarrow \exists \epsilon>0 \text { such that } U(x, \epsilon) \subset A . \tag{1.9}
\end{equation*}
$$

Note that the interior of $A$ is open.
We define the exterior of a set in terms of the interior of the set.
Definition 1.17. The exterior of $A$ is defined to be Ext $A \equiv \operatorname{Int} A^{c}$.
The boundary of a set is the collection of all points not in the interior or exterior.
Definition 1.18. The boundary of $A$ is defined to be $\operatorname{Bd} A \equiv X-((\operatorname{Ext} A) \cup(\operatorname{Int} A))$.
Always, we have $X=\operatorname{Int} A \cup \operatorname{Ext} A \cup \operatorname{Bd} A$.

## Lecture 2

### 1.6 Compactness

As usual, throughout this section we let $(X, d)$ be a metric space. We also remind you from last lecture we defined the open set

$$
\begin{equation*}
U\left(x_{o}, \lambda\right)=\left\{x \in X: d\left(x, x_{o}\right)<\lambda\right\} . \tag{1.10}
\end{equation*}
$$

Remark. If $U\left(x_{o}, \lambda\right) \subseteq U\left(x_{1}, \lambda_{1}\right)$, then $\lambda_{1}>d\left(x_{o}, x_{1}\right)$.
Remark. If $A_{i} \subseteq U\left(x_{o}, \lambda_{i}\right)$ for $i=1,2$, then $A_{1} \cup A_{2} \subseteq U\left(x_{o}, \lambda_{1}+\lambda_{2}\right)$.
Before we define compactness, we first define the notions of boundedness and covering.

Definition 1.19. A subset $A$ of $X$ is bounded if $A \subseteq U\left(x_{o}, \lambda\right)$ for some $\lambda$.
Definition 1.20. Let $A \subseteq X$. A collection of subsets $\left\{U_{\alpha} \subseteq X, \alpha \in I\right\}$ is a cover of $A$ if

$$
A \subset \bigcup_{\alpha \in I} U_{\alpha}
$$

Now we turn to the notion of compactness. First, we only consider compact sets as subsets of $\mathbb{R}^{n}$.

For any subset $A \subseteq \mathbb{R}^{n}$,

$$
A \text { is compact } \Longleftrightarrow A \text { is closed and bounded. }
$$

The above statement holds true for $\mathbb{R}^{n}$ but not for general metric spaces. To motivate the definition of compactness for the general case, we give the Heine-Borel Theorem.

Heine-Borel (H-B) Theorem. Let $A \subseteq \mathbb{R}^{n}$ be compact and let $\left\{U_{\alpha}, \alpha \in I\right\}$ be a cover of $A$ by open sets. Then a finite number of $U_{\alpha}$ 's already cover $A$.

The property that a finite number of the $U_{\alpha}$ 's cover $A$ is called the Heine-Borel (H-B) property. So, the H-B Theorem can be restated as follows: If $A$ is compact in $\mathbb{R}^{n}$, then $A$ has the $H$ - $B$ property.

Sketch of Proof. First, we check the H-B Theorem for some simple compact subsets of $\mathbb{R}^{n}$. Consider rectangles $Q=I_{1} \times \cdots \times I_{n} \subset \mathbb{R}^{n}$, where $I_{k}=\left[a_{k}, b_{k}\right]$ for each $k$. Starting with one dimension, it can by shown by induction that these rectangles have the H-B property.

Too prove the H-B theorem for general compact subsets, consider any closed and bounded (and therefore compact) subset $A$ of $\mathbb{R}^{n}$. Since $A$ is bounded, there exists a rectangle $Q$ such that $A \subseteq Q$. Suppose that the collection of subsets $\left\{U_{\alpha}, \alpha \in I\right\}$ is
an open cover of $A$. Then, define $U_{o}=\mathbb{R}^{n}-A$ and include $U_{o}$ in the open cover. The rectangle $Q$ has the H-B property and is covered by this new cover, so there exists a finite subcover covering $Q$. Furthermore, the rectangle $Q$ contains $A$, so the finite subcover also covers $A$, proving the H-B Theorem for general compact subsets.

The following theorem further motivates the general definition for compactness.
Theorem 1.21. If $A \subseteq \mathbb{R}^{n}$ has the $H$ - $B$ property, then $A$ is compact.
Sketch of Proof. We need to show that the H-B property implies $A$ is bounded (which we leave as an exercise) and closed (which we prove here).

To show that $A$ is closed, it is sufficient to show that $A^{c}$ is open. Take any $x_{o} \in A^{c}$, and define

$$
\begin{equation*}
C_{N}=\left\{x \in \mathbb{R}^{n}: d\left(x, x_{o}\right) \leq 1 / N\right\} \tag{1.11}
\end{equation*}
$$

and

$$
\begin{equation*}
U_{N}=C_{N}^{c} . \tag{1.12}
\end{equation*}
$$

Then,

$$
\begin{equation*}
\bigcap C_{N}=\left\{x_{o}\right\} \tag{1.13}
\end{equation*}
$$

and

$$
\begin{equation*}
\bigcup U_{N}=\mathbb{R}^{n}-\left\{x_{o}\right\} \tag{1.14}
\end{equation*}
$$

The $U_{N}$ 's cover $A$, so the H-B Theorem implies that there is a finite subcover $\left\{U_{N_{1}}, \ldots, U_{N_{k}}\right\}$ of $A$. We can take $N_{1}<N_{2}<\cdots<N_{k}$, so that $A \subseteq U_{N_{k}}$. By taking the complement, it follows that $C_{N_{k}} \subseteq A^{c}$. But $U\left(x_{o}, 1 / N_{k}\right) \subseteq C_{N_{k}}$, so $x_{o}$ is contained in an open subset of $A^{c}$. The above holds for any $x_{o} \in A^{c}$, so $A^{c}$ is open.

Let us consider the above theorem for arbitrary metric space $(X, d)$ and $A \subseteq X$.
Theorem 1.22. If $A \subseteq X$ has the $H-B$ property, then $A$ is closed and bounded.
Sketch of Proof. The proof is basically the same as for the previous theorem.
Unfortunately, the converse is not always true. Finally, we come to our general definition of compactness.

Definition 1.23. A subset $A \subseteq X$ is compact if it has the H-B property.
Compact sets have many useful properties, some of which we list here in the theorems that follow.

Theorem 1.24. Let $\left(X, d_{X}\right)$ and $\left(Y, d_{Y}\right)$ be metric spaces, and let $f: X \rightarrow Y$ be $a$ continuous map. If $A$ is a compact subset of $X$, then $f(A)$ is a compact subset of $Y$.

Proof. Let $\left\{U_{\alpha}, \alpha \in I\right\}$ be an open covering of $f(A)$. Each pre-image $f^{-1}\left(U_{\alpha}\right)$ is open in $X$, so $\left\{f^{-1}\left(U_{\alpha}\right): \alpha \in I\right\}$ is an open covering of $A$. The H-B Theorem says that there is a finite subcover $\left\{f^{-1}\left(U_{\alpha_{i}}\right): 1 \leq i \leq N\right\}$. It follows that the collection $\left\{U_{\alpha_{i}}: 1 \leq i \leq N\right\}$ covers $f(A)$, so $f(A)$ is compact.

A special case of the above theorem proves the following theorem.
Theorem 1.25. Let $A$ be a compact subset of $X$ and $f: X \rightarrow \mathbb{R}$ be a continuous map. Then $f$ has a maximum point on $A$.

Proof. By the above theorem, $f(A)$ is compact, which implies that $f(a)$ is closed and and bounded. Let $a=$ l.u.b. of $f(a)$. The point $a$ is in $f(A)$ because $f(A)$ is closed, so there exists an $x_{o} \in A$ such that $f\left(x_{o}\right)=a$.

Another useful property of compact sets involves the notion of uniform continuity.
Definition 1.26. Let $f: X \rightarrow \mathbb{R}$ be a continuous function, and let $A$ be a subset of $X$. The map $f$ is uniformly continuous on $A$ if for every $\epsilon>0$, there exists $\delta>0$ such that

$$
d(x, y)<\delta \Longrightarrow|f(x)-f(y)|<\epsilon,
$$

for all $x, y \in A$.
Theorem 1.27. If $f: X \rightarrow Y$ is continuous and $A$ is a compact subset of $X$, then $f$ is uniformly continuous on $A$.

Proof. Let $p \in A$. There exists a $\delta_{p}>0$ such that $|f(x)-f(p)|<\epsilon / 2$ for all $x \in U\left(p, \delta_{p}\right)$. Now, consider the collection of sets $\left\{U\left(p, \delta_{p} / 2\right): p \in A\right\}$, which is an open cover of $A$. The H-B Theorem says that there is a finite subcover $\left\{U\left(p_{i}, \delta_{p_{i}} / 2\right)\right.$ : $1 \leq i \leq N\}$. Choose $\delta \leq \min \delta_{p_{i}} / 2$. The following claim finishes the proof.

Claim. If $d(x, y)<\delta$, then $|f(x)-f(y)|<\epsilon$.
Proof. Given $x$, choose $p_{i}$ such that $x \in U\left(p_{i}, \delta_{p_{i}} / 2\right)$. So, $d\left(p_{i}, x\right)<\delta_{p_{i}} / 2$ and $d(x, y)<$ $\delta<\delta_{p_{i}} / 2$. By the triangle inequality we conclude that $d\left(p_{i}, y\right)<\delta_{p_{i}}$. This shows that $x, y \in U\left(p_{i}, \delta_{p_{i}}\right)$, which implies that $\left|f(x)-f\left(p_{i}\right)\right|<\epsilon / 2$ and $\left|f(y)-f\left(p_{i}\right)\right|<\epsilon / 2$. Finally, by the triangle inequality, $|f(x)-f(y)|<\epsilon$, which proves our claim.

### 1.7 Connectedness

As usual, let $(X, d)$ be a metric space.
Definition 1.28. The metric space $(X, d)$ is connected if it is impossible to write $X$ as a disjoint union $X=U_{1} \cup U_{2}$ of non-empty open sets $U_{1}$ and $U_{2}$.

Note that disjoint simply means that $U_{1} \cap U_{2}=\phi$, where $\phi$ is the empty set.
A few simple examples of connected spaces are $\mathbb{R}, \mathbb{R}^{n}$, and $I=[a, b]$. The following theorem shows that a connected space gets mapped to a connected subspace by a continuous function.

Theorem 1.29. Given metric spaces $\left(X, d_{X}\right)$ and $\left(Y, d_{Y}\right)$, and a continuous map $f: X \rightarrow Y$, it follows that

$$
X \text { is connected } \Longrightarrow f(X) \text { is connected. }
$$

Proof. Suppose $f(X)$ can be written as a union of open sets $f(X)=U_{1} \cup U_{2}$ such that $U_{1} \cap U_{2}=\phi$. Then $X=f^{-1}\left(U_{1}\right) \cup f^{-1}\left(U_{2}\right)$ is a disjoint union of open sets. This contradicts that $X$ is connected.

The intermediate-value theorem follows as a special case of the above theorem.
Intermediate-value Theorem. Let $(X, d)$ be connected and $f: X \rightarrow \mathbb{R}$ be a continuous map. If $a, b \in f(X)$ and $a<r<b$, then $r \in f(X)$.

Proof. Suppose $r \notin f(X)$. Let $A=(-\infty, r)$ and $B=(r, \infty)$. Then $X=f^{-1}(A) \cup$ $f^{-1}(B)$ is a disjoint union of open sets, a contradiction.

## Lecture 3

## 2 Differentiation

### 2.1 Differentiation in $n$ dimensions

We are setting out to generalize to $n$ dimensions the notion of differentiation in onedimensional calculus. We begin with a review of elementary one-dimensional calculus.

Let $I \subseteq \mathbb{R}$ be an open interval, let $f: I \rightarrow \mathbb{R}$ be a map, and let $a \in I$.
Definition 2.1. The derivative of $f$ at $a$ is

$$
\begin{equation*}
f^{\prime}(a)=\lim _{t \rightarrow 0} \frac{f(a+t)-f(a)}{t}, \tag{2.1}
\end{equation*}
$$

provided that the limit exists. If the limit exists, then $f$ is differentiable at $a$.
There are half a dozen or so possible reasonable generalizations of the notion of derivative to higher dimensions. One candidate generalization which you have probably already encountered is the directional derivative.

Definition 2.2. Given an open set $U$ in $\mathbb{R}^{n}$, a map $f: U \rightarrow \mathbb{R}^{m}$, a point $a \in U$, and a point $u \in \mathbb{R}^{n}$, the directional derivative of $f$ in the direction of $u$ at $a$ is

$$
\begin{equation*}
D_{u} f(a)=\lim _{t \rightarrow 0} \frac{f(a+t u)-f(a)}{t} \tag{2.2}
\end{equation*}
$$

provided that the limit exists.
In particular, we can calculate the directional derivatives in the direction of the standard basis vectors $e_{1}, \ldots, e_{n}$ of $\mathbb{R}^{n}$, where

$$
\begin{align*}
e_{1} & =(1,0, \ldots, 0),  \tag{2.3}\\
e_{2} & =(0,1,0, \ldots, 0),  \tag{2.4}\\
& \vdots  \tag{2.5}\\
e_{n} & =(0, \ldots, 0,1) . \tag{2.6}
\end{align*}
$$

Notation. The directional derivative in the direction of a standard basis vector $e_{i}$ of $\mathbb{R}^{n}$ is denoted by

$$
\begin{equation*}
D_{i} f(a)=D_{e_{i}} f(a)=\frac{\partial}{\partial x_{i}} f(a) \tag{2.7}
\end{equation*}
$$

We now try to answer the following question: What is an adequate definition of differentiability at a point $a$ for a function $f: U \rightarrow \mathbb{R}^{m}$ ?

- Guess 1: Require that $\frac{\partial f}{\partial x_{i}}(a)$ exists.

However, this requirement is inadequate. Consider the function defined by

$$
f\left(x_{1}, x_{2}\right)= \begin{cases}0, & \text { if }\left(x_{1}, x_{2}\right) \text { lies on the } x_{1} \text {-axis or the } x_{2} \text {-axis }  \tag{2.8}\\ 1, & \text { otherwise }\end{cases}
$$

Then, both

$$
\begin{equation*}
\frac{\partial f}{\partial x_{1}}(0)=0 \text { and } \frac{\partial f}{\partial x_{2}}(0)=0 \tag{2.9}
\end{equation*}
$$

but the function $f$ is not differentiable at $(0,0)$ along any other direction.

- Guess 2: Require that all directional derivatives exist at $a$.

Unfortunately, this requirement is still inadequate. For example (from Munkres chapter 5), consider the function $f: \mathbb{R}^{2} \rightarrow \mathbb{R}$ defined by

$$
f(x, y)= \begin{cases}\frac{x y^{2}}{x^{2}+y^{4}}, & (x, y) \neq(0,0)  \tag{2.10}\\ 0, & x=y=0\end{cases}
$$

Claim. The directional derivative $D_{u} f(0)$ exists for all $u$.
Proof. Let $u=(h, k)$. Then

$$
\begin{align*}
\lim _{t \rightarrow 0} \frac{f(t u)-f(0)}{t} & =\lim _{t \rightarrow 0} \frac{f(t u)}{t} \\
& =\lim _{t \rightarrow 0}\left(\frac{t^{3} h k^{2}}{t^{2} h^{2}+t^{4} k^{4}}\right) \frac{1}{t}  \tag{2.11}\\
& = \begin{cases}0, & h=0 \\
k^{2} / h, & h \neq 0 .\end{cases}
\end{align*}
$$

So the limit exists for every $u$.
However, the function is a non-zero constant on a parabola passing through the origin: $f\left(t^{2}, t\right)=\frac{t^{4}}{2 t^{4}}=\frac{1}{2}$, except at the origin where $f(0,0)=0$. The function $f$ is discontinuous at the origin despite the existence of all directional derivatives.

- Guess 3. This guess will turn out to be correct.

Remember than in one-dimensional calculus we defined

$$
\begin{equation*}
f^{\prime}(a)=\lim _{t \rightarrow 0} \frac{f(a+t)-f(a)}{t} \tag{2.12}
\end{equation*}
$$

for a function $f: I \rightarrow \mathbb{R}$ and a point $a \in I$. Now consider the function $\lambda: \mathbb{R} \rightarrow \mathbb{R}$ defined by

$$
\begin{equation*}
\lambda(t)=f^{\prime}(a) t \tag{2.13}
\end{equation*}
$$

Then,

$$
\begin{align*}
\lim _{t \rightarrow 0} \frac{f(a+t)-f(a)-\lambda(t)}{t} & =\lim _{t \rightarrow 0} \frac{f(a+t)-f(a)}{t}-f^{\prime}(a)  \tag{2.14}\\
& =0
\end{align*}
$$

So, $\lambda(t) \approx f(a+t)-f(a)$ when $t$ is small.
Now we generalize to $n$ dimensions.
Definition 2.3. Given an open subset $U$ of $\mathbb{R}^{n}$, a map $f: U \rightarrow \mathbb{R}^{m}$, and a point $a \in U$, the function $f$ is differentiable at $a$ if there exists a linear mapping $B: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ such that for every $h \in \mathbb{R}^{n}-\{0\}$,

$$
\begin{equation*}
\frac{f(a+h)-f(a)-B h}{|h|} \rightarrow 0 \text { as } h \rightarrow 0 . \tag{2.15}
\end{equation*}
$$

That is, $f(a+h)-f(a) \approx B h$ when $h$ is small.
Theorem 2.4. If $f$ is differentiable at $a$, then for every $u$ the directional derivative of $f$ in the direction of $u$ at a exists.

Proof. The function $f$ is differentiable at $a$, so

$$
\begin{equation*}
\frac{f(a+t u)-f(a)-B(t u)}{|t u|} \rightarrow 0 \text { as } t \rightarrow 0 . \tag{2.16}
\end{equation*}
$$

Furthermore,

$$
\begin{align*}
\frac{f(a+t u)-f(a)-B(t u)}{|t u|} & =\frac{t}{|t u|} \frac{f(a+t u)-f(a)-B(t u)}{t} \\
& =\frac{t}{|t|} \frac{1}{|u|}\left(\frac{f(a+t u)-f(a)}{t}-B u\right)  \tag{2.17}\\
& \rightarrow 0,
\end{align*}
$$

as $t \rightarrow 0$, so

$$
\begin{equation*}
\frac{f(a+t u)-f(a)}{t} \rightarrow B u \text { as } t \rightarrow 0 . \tag{2.18}
\end{equation*}
$$

Furthermore, the linear map $B$ is unique, so the following definition is well-defined.
Definition 2.5. The derivative of $f$ at $a$ is $D f(a)=B$, where $B$ is the linear map defined above.

Note that $D f(a): \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ is a linear map.

Theorem 2.6. If $f$ is differentiable at $a$, then $f$ is continuous at $a$.
Sketch of Proof. Note that for $h \neq 0$ in $\mathbb{R}^{n}$,

$$
\begin{equation*}
\frac{f(a+h)-f(a)-B h}{|h|} \rightarrow 0 \text { as } h \rightarrow 0 \tag{2.19}
\end{equation*}
$$

implies that

$$
\begin{equation*}
f(a+h)-f(a)-B h \rightarrow 0 \text { as } h \rightarrow 0 . \tag{2.20}
\end{equation*}
$$

From this you can conclude that $f$ is continuous at $a$.
Remark. Let $L: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ be a linear map and $a \in \mathbb{R}^{n}$. The point $a$ can be written as a sum $a=\sum_{j=1}^{n} a_{j} e_{j}=\left(a_{1}, \ldots, a_{n}\right)$. The point $L a$ can be written as the sum $L a=\sum a_{j} L e_{j}$, and $L$ can be written out in components as $L=\left(L_{1}, \ldots, L_{m}\right)$, where each $L_{j}: \mathbb{R}^{n} \rightarrow \mathbb{R}$ is a linear map. Then $L e_{j}=\left(L_{1}, e_{j}, \ldots, L_{m} e_{j}\right)$, and $L_{i} e_{j}=\ell_{i, j}$. The numbers $\ell_{i, j}$ form an $n \times n$ matrix denoted by $\left[\ell_{i, j}\right]$.

Remark. Let $U \subseteq \mathbb{R}^{n}$, and let $f_{1}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m_{1}}$ and $f_{2}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m_{2}}$ be differentiable maps. Let $m=m_{1}+m_{2}$, so that $\mathbb{R}^{m_{1}} \times \mathbb{R}^{m_{2}}=\mathbb{R}^{m}$. Now, construct a function $f: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ defined in component form by $f=\left(f_{1}, f_{2}\right)$. The derivative of $f$ at $a$ is

$$
\begin{equation*}
D f(a)=\left(D f_{1}(a), D f_{2}(a)\right) . \tag{2.21}
\end{equation*}
$$

Remark. Let $f: U \rightarrow \mathbb{R}^{m}$ be a map. The action of $f$ on input $x$ written out in component form is $f(x)=\left(f_{1}(x), \ldots, f_{m}(x)\right)$. So, the map can be represented in component form as $f=\left(f_{1}, \ldots, f_{m}\right)$, where each $f_{i}$ as a map of the form $f_{i}: U \rightarrow \mathbb{R}$. The derivative of $f$ acting on the standard basis vector $e_{j}$ is

$$
\begin{align*}
D f(a) e_{j} & =\left(D f_{1}(a) e_{j}, \ldots, D f_{m}(a) e_{j}\right) \\
& =\left(\frac{\partial f_{1}}{\partial x_{j}}(a), \ldots, \frac{\partial f_{m}}{\partial x_{j}}(a)\right) . \tag{2.22}
\end{align*}
$$

So, the derivative $(D f)(a)$ can be represented by an $m \times n$ matrix

$$
\begin{equation*}
(D f)(a) \cong J_{f}(a)=\left[\frac{\partial f_{i}}{\partial x_{j}}(a)\right] \tag{2.23}
\end{equation*}
$$

called the Jacobian matrix of $f$ at $a$, which you probably recognize.

## Lecture 4

### 2.2 Conditions for Differentiability

In this lecture we will discuss conditions that guarantee differentiability. First, we begin with a review of important results from last lecture.

Let $U$ be an open subset of $\mathbb{R}^{n}$, let $f: U \rightarrow \mathbb{R}^{n}$ be a map, and let $a \in U$.
We defined $f$ to be differentiable at $a$ if there exists a linear map $B: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ such that for $h \in \mathbb{R}^{n}-\{0\}$,

$$
\begin{equation*}
\frac{f(a+h)-f(a)-B h}{|h|} \rightarrow 0 \text { as } h \rightarrow 0 . \tag{2.24}
\end{equation*}
$$

If such a $B$ exists, then it is unique and $B=D f(a)$. The matrix representing $B$ is the Jacobian matrix $J_{f}(a)=\left[\frac{\partial f_{i}}{\partial x_{j}}(a)\right]$, where $f=\left(f_{1}, \ldots, f_{m}\right)$.

Note that the mere existence of all of the partial derivatives in the Jacobian matrix does not guarantee differentiability.

Now we discuss conditions that guarantee differentiability.
Theorem 2.7. Suppose that all of the partial derivatives $\frac{\partial f_{i}}{\partial x_{j}}$ in the Jacobian matrix exist at all points $x \in U$, and that all of the partial derivatives are continuous at $x=a$. Then $f$ is differentiable at $a$.

Sketch of Proof. This theorem is very elegantly proved in Munkres, so we will simply give the general ideas behind the proof here.

First, we look at the case $n=2, m=1$. The main ingredient in the proof is the Mean Value Theorem from 1-D calculus, which we state here without proof.
Mean Value Theorem. Given an interval $[a, b] \subseteq \mathbb{R}$ and a map $\phi:[a, b] \rightarrow \mathbb{R}$, if $\phi$ is continuous on $[a, b]$ and differentiable on $(a, b)$, then there exists a point $c \in(a, b)$ such that $\phi(b)-\phi(a)=\phi^{\prime}(c)(b-a)$.

Now we continue with the proof. Let $f$ be a map $f: U \rightarrow \mathbb{R}$, where $U \subseteq \mathbb{R}^{2}$. So, $f$ is a function of two variables $f=f\left(x_{1}, x_{2}\right)$. Consider a point $a=\left(a_{1}, a_{2}\right) \in U$ and any point $h \in \mathbb{R}^{2}-\{0\}$ "close" to zero, where by close we mean $a+h \in U$. We want to compute $f(a+h)-f(a)$.

$$
\begin{align*}
f(a+h)-f(a)= & f\left(a_{1}+h_{1}, a_{2}+h_{2}\right)-f\left(a_{1}, a_{2}\right) \\
= & f\left(a_{1}+h_{1}, a_{2}+h_{2}\right)-f\left(a_{1}, a_{2}+h_{2}\right)  \tag{2.25}\\
& +f\left(a_{1}, a_{2}+h_{2}\right)-f\left(a_{1}, a_{2}\right) .
\end{align*}
$$

Thinking of the first two terms as functions of the first argument only, and thinking of the last two terms as functions of the second term only, and applying the Mean

Value Theorem to each pair of terms, we obtain

$$
\begin{align*}
f(a+h)-f(a)= & \frac{\partial f}{\partial x_{1}}\left(c_{1}, a_{2}+h_{2}\right) h_{1} \\
& +\frac{\partial f}{\partial x_{2}}\left(a_{1}, d_{2}\right) h_{2} \tag{2.26}
\end{align*}
$$

where $a_{1}<c_{1}<a_{1}+h_{1}$ and $a_{2}<d_{2}<a_{2}+h_{2}$. This can be rewritten as

$$
\begin{equation*}
f(a+h)-f(a)=\frac{\partial f}{\partial x_{1}}(c) h_{1}+\frac{\partial f}{\partial x_{2}}(d) h_{2} \tag{2.27}
\end{equation*}
$$

where $c=\left(c_{1}, a_{2}+h_{2}\right)$ and $d=\left(a_{1}, d_{2}\right)$.
We want to show that $(f(a+h)-f(a)-D f(a) h) /|h| \rightarrow 0$ as $h \rightarrow 0$, where $D f(a)=\left[\frac{\partial f}{\partial x_{1}}(a), \frac{\partial f}{\partial x_{2}}(a)\right]$. Using our previously derived expression for $f(a+h)-f(a)$, we find that

$$
\begin{align*}
f(a+h)-f(a)-D f(a) h & =f(a+h)-f(a)-\frac{\partial f}{\partial x_{1}}(a) h_{1}-\frac{\partial f}{\partial x_{2}}(a) h_{2} \\
& =\left(\frac{\partial f}{\partial x_{1}}(c)-\frac{\partial f}{\partial x_{1}}(a)\right) h_{1}+\left(\frac{\partial f}{\partial x_{2}}(d)-\frac{\partial f}{\partial x_{2}}(a)\right) h_{2} . \tag{2.28}
\end{align*}
$$

We can use the sup norm to show that

$$
\begin{equation*}
|f(a+h)-f(a)-D f(a) h| \leq\left|\frac{\partial f}{\partial x_{1}}(c)-\frac{\partial f}{\partial x_{1}}(a)\right|\left|h_{1}\right|+\left\lvert\,\left(\left.\frac{\partial f}{\partial x_{2}}(d)-\frac{\partial f}{\partial x_{2}}(a)| | h_{2} \right\rvert\,\right.\right. \tag{2.29}
\end{equation*}
$$

from which it follows that

$$
\begin{equation*}
\frac{|f(a+h)-f(a)-D f(a) h|}{|h|} \leq\left|\frac{\partial f}{\partial x_{1}}(c)-\frac{\partial f}{\partial x_{1}}(a)\right|+\left\lvert\,\left(\left.\frac{\partial f}{\partial x_{2}}(d)-\frac{\partial f}{\partial x_{2}}(a) \right\rvert\,,\right.\right. \tag{2.30}
\end{equation*}
$$

where we used the fact that $|h|=\max \left(\left|h_{1}\right|,\left|h_{2}\right|\right)$.
Notice that as $h \rightarrow 0$, both $c \rightarrow a$ and $d \rightarrow a$, as can be easily seen using the following diagram. This means that the r.h.s. of Equation (2.30) goes to zero as $h$

goes to zero, because the partial derivatives are continuous. It follows that the l.h.s. goes to zero, which completes our proof.

The proof in $n$ dimensions is similar to the above proof, but the details are harder to follow.

We now introduce a useful class of functions.
Definition 2.8. Given $U \subseteq \mathbb{R}^{n}$ and $f: U \rightarrow \mathbb{R}$, we define $f \in \mathcal{C}^{1}(U) \Longleftrightarrow \frac{\partial f}{\partial x_{i}}, i=1, \ldots, n$ exist and are continuous at all points $x \in U$.

Similarly, we define

$$
\begin{align*}
& f \in \mathcal{C}^{2}(U) \Longleftrightarrow \frac{\partial f}{\partial x_{i}} \in \mathcal{C}^{1}(U), i=a, \ldots, n  \tag{2.32}\\
& f \in \mathcal{C}^{k}(U) \Longleftrightarrow \frac{\partial f}{\partial x_{i}} \in \mathcal{C}^{k-1}(U), i=a, \ldots, n  \tag{2.33}\\
& f \in \mathcal{C}^{\infty}(U) \Longleftrightarrow f \in \mathcal{C}^{k}(U) \forall k \tag{2.34}
\end{align*}
$$

If $f$ is multiply differentiable, then you can perform higher order mixed partial derivatives.

One of the fundamental theorems of calculus is that the order of the partial derivatives can be taken in any order. For example,

$$
\begin{equation*}
\frac{\partial}{\partial x_{i}}\left(\frac{\partial}{\partial x_{j}}\right)=\frac{\partial}{\partial x_{j}}\left(\frac{\partial}{\partial x_{i}}\right) \equiv \frac{\partial^{2} f}{\partial x_{i} \partial x_{j}} \tag{2.35}
\end{equation*}
$$

Let's do the proof for this case. Let $U \subseteq \mathbb{R}^{2}$ and $f=f\left(x_{1}, x_{2}\right)$. We prove the following claim:

Claim.

$$
\begin{equation*}
\frac{\partial}{\partial x_{i}}\left(\frac{\partial}{\partial x_{j}}\right)=\frac{\partial}{\partial x_{j}}\left(\frac{\partial}{\partial x_{i}}\right) \tag{2.36}
\end{equation*}
$$

Proof. Take $a \in U$ written in components as $a=\left(a_{1}, a_{2}\right)$, and take $h=\left(h_{1}, h_{2}\right) \in$ $\mathbb{R}^{2}-\{0\}$ such that $a+h \in U$. That is, take $h \approx 0$.

Define

$$
\begin{equation*}
\Delta(h)=f\left(a_{1}+h_{1}, a_{2}+h_{2}\right)-f\left(a_{1}, a_{2}+h_{2}\right)-f\left(a_{1}+h_{1}, a_{2}\right)+f\left(a_{1}, a_{2}\right) \tag{2.37}
\end{equation*}
$$

and define

$$
\begin{equation*}
\phi(s)=f\left(a_{1}+h_{1}, s\right)-f\left(a_{1}, s\right) \tag{2.38}
\end{equation*}
$$

where $a_{2} \leq s \leq a_{2}+h_{2}$. We find that

$$
\begin{align*}
\Delta(h) & =\phi\left(a_{2}+h_{2}\right)-\phi\left(a_{2}\right) \\
& =\phi^{\prime}\left(c_{2}\right) h_{2}, a_{2}<c_{2}<a_{2}+h_{2} \tag{2.39}
\end{align*}
$$

by the Mean Value Theorem. Writing out $\phi^{\prime}$ using partial derivatives of $f$, and using the Mean Value Theorem again, we find

$$
\begin{align*}
\Delta(h) & =\left(\frac{\partial f}{\partial x_{2}}\left(a_{1}+h_{1}, c_{2}\right)-\frac{\partial f}{\partial x_{1}}\left(a_{1}, c_{2}\right)\right) h_{2} \\
& =\left(\frac{\partial}{\partial x_{1}}\left(\frac{\partial f}{\partial x_{2}}\left(c_{1}, c_{2}\right) h_{1}\right)\right) h_{2}, a_{1}<c_{1}<a_{1}+h_{1}  \tag{2.40}\\
& =\left(\frac{\partial}{\partial x_{1}}\left(\frac{\partial}{\partial x_{2}} f\right)\right)(c) h_{1} h_{2} \\
& =\left(\frac{\partial}{\partial x_{2}}\left(\frac{\partial}{\partial x_{1}} f\right)\right)(d) h_{1} h_{2},
\end{align*}
$$

where we obtained the last line by symmetry. This shows that

$$
\begin{equation*}
\frac{\partial}{\partial x_{1}}\left(\frac{\partial f}{\partial x_{2}}\right)(c)=\frac{\partial}{\partial x_{2}}\left(\frac{\partial f}{\partial x_{1}}\right)(d) . \tag{2.41}
\end{equation*}
$$

As $h \rightarrow 0, c \rightarrow a$ and $d \rightarrow a$, so

$$
\begin{equation*}
\frac{\partial}{\partial x_{1}}\left(\frac{\partial f}{\partial x_{2}}\right)(a)=\frac{\partial}{\partial x_{2}}\left(\frac{\partial f}{\partial x_{1}}\right)(a) \tag{2.42}
\end{equation*}
$$

for any $a \in U$.
The above argument can be iterated for $f \in \mathcal{C}^{3}(U), f \in \mathcal{C}^{3}(4)$, etc.

## Lecture 5

### 2.3 Chain Rule

Let $U$ and $v$ be open sets in $\mathbb{R}^{n}$. Consider maps $f: U \rightarrow V$ and $g: V \rightarrow \mathbb{R}^{k}$. Choose $a \in U$, and let $b=f(a)$. The composition $g \circ f: U \rightarrow \mathbb{R}^{k}$ is defined by $(g \circ f)(x)=g(f(x))$.

Theorem 2.9. If $f$ is differentiable at $a$ and $g$ is differentiable at $b$, then $g \circ f$ is differentiable at $a$, and the derivative is

$$
\begin{equation*}
(D g \circ f)(a)=(D g)(b) \circ D f(a) . \tag{2.43}
\end{equation*}
$$

Proof. This proof follows the proof in Munkres by breaking the proof into steps.

- Step 1: Let $h \in \mathbb{R}^{n}-\{0\}$ and $h \dot{=} 0$, by which we mean that $h$ is very close to zero. Consider $\Delta(h)=f(a+h)-f(a)$, which is continuous, and define

$$
\begin{equation*}
F(h)=\frac{f(a+h)-f(a)-D f(a) h}{|a|} . \tag{2.44}
\end{equation*}
$$

Then $f$ is differentiable at $a$ if and only if $F(h) \rightarrow 0$ as $h \rightarrow 0$.

$$
\begin{equation*}
F(h)=\frac{\Delta(h)-D f(a) h}{|h|}, \tag{2.45}
\end{equation*}
$$

so

$$
\begin{equation*}
\Delta(h)=D f(a) h+|h| F(h) . \tag{2.46}
\end{equation*}
$$

## Lemma 2.10.

$$
\begin{equation*}
\frac{\Delta(h)}{|h|} \text { is bounded. } \tag{2.47}
\end{equation*}
$$

Proof. Define

$$
\begin{equation*}
|D f(a)|=\sup _{i}\left|\frac{\partial f}{\partial x_{i}}(a)\right|, \tag{2.48}
\end{equation*}
$$

and note that

$$
\begin{equation*}
\frac{\partial f}{\partial x_{i}}(a)=D f(a) e_{i}, \tag{2.49}
\end{equation*}
$$

where the $e_{i}$ are the standard basis vectors of $\mathbb{R}^{n}$. If $h=\left(h_{1}, \ldots, h_{n}\right)$, then $h=\sum h_{i} e_{i}$. So, we can write

$$
\begin{equation*}
D f(a) h=\sum h_{i} D f(a) e_{i}=\sum h_{i} \frac{\partial f}{\partial x_{i}}(a) . \tag{2.50}
\end{equation*}
$$

It follows that

$$
\begin{align*}
|D f(a) h| & \leq \sum_{i=1}^{m} h_{i}\left|\frac{\partial f}{\partial x_{i}}(a)\right|  \tag{2.51}\\
& \leq m|h||D f(a)|
\end{align*}
$$

By Equation 2.46,

$$
\begin{equation*}
|\Delta(h)| \leq m|h||D f(a)|+|h| F(h), \tag{2.52}
\end{equation*}
$$

so

$$
\begin{equation*}
\frac{|\Delta(h)|}{|h|} \leq m|D f(a)|+F(h) . \tag{2.53}
\end{equation*}
$$

- Step 2: Remember that $b=f(a), g: V \rightarrow \mathbb{R}^{k}$, and $b \in V$. Let $k \dot{=} 0$. This means that $k \in \mathbb{R}^{n}-\{0\}$ and that $k$ is close to zero. Define

$$
\begin{equation*}
G(k)=\frac{g(b+k)-g(b)-(D g)(b) k}{|k|}, \tag{2.54}
\end{equation*}
$$

so that

$$
\begin{equation*}
g(b+k)-g(b)=D g(b) k+|k| G(k) . \tag{2.55}
\end{equation*}
$$

We proceed to show that $g \circ f$ is differentiable at $a$.

$$
\begin{align*}
g \circ f(a+h)-g \circ f(a) & =g(f(a+h))-g(f(a))  \tag{2.56}\\
& =g(b+\Delta(h))-g(b),
\end{align*}
$$

where $f(a)=b$ and $f(a+h)=f(a)+\Delta(h)=b+\Delta(h)$. Using Equation 2.55 we see that the above expression equals

$$
\begin{equation*}
D g(b) \Delta(h)+|\Delta(h)| G(\Delta(h)) . \tag{2.57}
\end{equation*}
$$

Substituting in from Equation 2.46, we obtain

$$
\begin{align*}
g \circ f(a+h)-g \circ f(a) & =\ldots \\
& =D g(b)(D f(a) h+|h| F(h))+\ldots \\
& =D g(b) \circ D f(a) h+|h| D g(b) F(h)+|\Delta(h)| G(\Delta(h)) \tag{2.58}
\end{align*}
$$

This shows that

$$
\begin{equation*}
\frac{g \circ f(a+h)-g \circ f(a)-D g(b) \circ D f(a) h}{|h|}=D g(b) F(h)+\frac{\Delta(h)}{|h|} G(\Delta(h)) . \tag{2.59}
\end{equation*}
$$

We see in the above equation that $g \circ f$ is differentiable at $a$ if and only if the l.h.s. goes to zero as $h \rightarrow 0$. It suffices to show that the r.h.s. goes to zero as $h \rightarrow 0$, which it does: $F(h) \rightarrow 0$ as $h \rightarrow 0$ because $f$ is differentiable at $a$; $G(\Delta(h)) \rightarrow 0$ because $g$ is differentiable at $b$; and $\Delta(h) /|h|$ is bounded.

We consider the same maps $g$ and $f$ as above, and we write out $f$ in component form as $f=\left(f_{1}, \ldots, f_{n}\right)$ where each $f_{i}: U \rightarrow \mathbb{R}$. We say that $f$ is a $\mathcal{C}^{r}$ map if each $f_{i} \in \mathcal{C}^{r}(U)$. We associate $D f(x)$ with the matrix

$$
\begin{equation*}
D f(x) \sim\left[\frac{\partial f_{i}}{\partial x_{j}}(x)\right] . \tag{2.60}
\end{equation*}
$$

By definition, $f$ is $\mathcal{C}^{r}$ (that is to say $f \in \mathcal{C}^{r}(U)$ ) if and only if $D f$ is $\mathcal{C}^{r-1}$.
Theorem 2.11. If $f: U \rightarrow V \subseteq \mathbb{R}^{n}$ is a $\mathcal{C}^{r}$ map and $g: V \rightarrow \mathbb{R}^{p}$ is a $\mathcal{C}^{r}$ map, then $g \circ f: U \rightarrow \mathbb{R}^{p}$ is a $\mathcal{C}^{r}$ map.

Proof. We only prove the case $r=1$ and leave the general case, which is inductive, to the student.

- Case $r=1$ :

$$
\begin{equation*}
D g \circ f(x)=D g(f(x)) \circ D f(x) \sim\left[\frac{\partial g_{i}}{\partial x_{j}} f(x)\right] \tag{2.61}
\end{equation*}
$$

The map $g$ is $\mathcal{C}^{1}$, which implies that $\partial g_{i} / \partial x_{j}$ is continuous. Also,

$$
\begin{equation*}
D f(x) \sim\left[\frac{\partial f_{i}}{\partial x_{j}}\right] \tag{2.62}
\end{equation*}
$$

is continuous. It follows that $D g \circ f(x)$ is continuous. Hence, $g \circ f$ is $\mathcal{C}^{1}$.

### 2.4 The Mean-value Theorem in $n$ Dimensions

Theorem 2.12. Let $U$ be an open subset of $\mathbb{R}^{n}$ and $f: U \rightarrow \mathbb{R}$ a $\mathcal{C}^{1}$ map. For $a \in U$, $h \in \mathbb{R}^{n}$, and $h \doteq 0$,

$$
\begin{equation*}
f(a+h)-f(a)=D f(c) h, \tag{2.63}
\end{equation*}
$$

where $c$ is a point on the line segment $a+t h, 0 \leq t \leq 1$, joining a to $a+h$.
Proof. Define a map $\phi:[0,1] \rightarrow \mathbb{R}$ by $\phi(t)=f(a+t h)$. The Mean Value Theorem implies that $\phi(1)-\phi(0)=\phi^{\prime}(c)=(D f)(c) h$, where $0<c<1$. In the last step we used the chain rule.

### 2.5 Inverse Function Theorem

Let $U$ and $V$ be open sets in $\mathbb{R}^{n}$, and let $f: U \rightarrow V$ be a $\mathcal{C}^{1}$ map. Suppose there exists a map $g: V \rightarrow U$ that is the inverse map of $f$ (which is also $\mathcal{C}^{1}$ ). That is, $g(f(x))=x$, or equivalently $g \circ f$ equals the identity map.

Using the chain rule, if $a \in U$ and $b=f(a)$, then

$$
\begin{equation*}
(D g)(b)=\text { the inverse of } D f(a) \tag{2.64}
\end{equation*}
$$

That is, $D g(b) \circ D f(a)$ equals the identity map. So,

$$
\begin{equation*}
D g(b)=(D f(a))^{-1} \tag{2.65}
\end{equation*}
$$

However, this is not a trivial matter, since we do not know if the inverse exists. That is what the inverse function theorem is for: if $D f(a)$ is invertible, then $g$ exists for some neighborhood of $a$ in $U$ and some neighborhood of $f(a)$ in $V$. We state this more precisely in the following lecture.

## Lecture 6

We begin with a review of some earlier definitions.
Let $\delta>0$ and $a \in \mathbb{R}^{n}$.

$$
\begin{equation*}
\text { Euclidean ball: } B_{\delta}(a)=\left\{x \in \mathbb{R}^{n}:\|x-a\|<\delta\right\} \tag{2.66}
\end{equation*}
$$

$$
\text { Supremum ball: } \begin{align*}
R_{\delta}(a) & =\left\{x \in \mathbb{R}^{n}:|x-a|<\delta\right\} \\
& =I_{1} \times \cdots \times I_{n}, I_{j}=\left(a_{j}-\delta, a_{j}+\delta\right) . \tag{2.67}
\end{align*}
$$

Note that the supremum ball is actually a rectangle. Clearly, $B_{\delta}(a) \subseteq R_{\delta}(a)$. We use the notation $B_{\delta}=B_{\delta}(0)$ and $R_{\delta}=R_{\delta}(0)$.

Continuing with our review, given $U$ open in $\mathbb{R}^{n}$, a map $f: U \rightarrow \mathbb{R}^{k}$, and a point $a \in U$, we defined the derivate $D f(a): \mathbb{R}^{n} \rightarrow \mathbb{R}^{k}$ which we associated with the matrix

$$
\begin{equation*}
D f(a) \sim\left[\frac{\partial f_{i}}{\partial x_{j}}(a)\right] \tag{2.68}
\end{equation*}
$$

and we define

$$
\begin{equation*}
|D f(a)|=\sup _{i, j}\left|\frac{\partial f_{i}}{\partial x_{j}}(a)\right| \tag{2.69}
\end{equation*}
$$

Lastly, we define $U \subseteq \mathbb{R}^{n}$ to be convex if

$$
\begin{equation*}
a, b \in U \Longrightarrow(1-t) a+t b \in U \text { for all } 0 \leq t \leq 1 \tag{2.70}
\end{equation*}
$$

Before we state and prove the Inverse Function Theorem, we give the following definition.

Definition 2.13. Let $U$ and $V$ be open sets in $\mathbb{R}^{n}$ and $f: U \rightarrow V$ a $\mathcal{C}^{r}$ map. The map $f$ is is a $\mathcal{C}^{r}$ diffeomorphism if it is bijective and $f^{-1}: V \rightarrow U$ is also $\mathcal{C}^{r}$.

Inverse Function Theorem. Let $U$ be an open set in $\mathbb{R}^{n}, f: U \rightarrow \mathbb{R}^{n}$ a $\mathcal{C}^{r}$ map, and $a \in U$. If $D f(a): \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is bijective, then there exists a neighborhood $U_{1}$ of a in $U$ and a neighborhood $V$ of $f(a)$ in $\mathbb{R}^{n}$ such that $F \mid U_{1}$ is a $\mathcal{C}^{r}$ diffeomorphism of $U_{1}$ at $V$.

Proof. To prove this we need some elementary multi-variable calculus results, which we provide with the following lemmas.

Lemma 2.14. Let $U$ be open in $\mathbb{R}^{n}$ and $F: U \rightarrow \mathbb{R}^{k}$ be a $\mathcal{C}^{1}$ mapping. Also assume that $U$ is convex. Suppose that $|D f(a)| \leq c$ for all $A \in U$. Then, for all $x, y \in U$,

$$
\begin{equation*}
|f(x)-f(y)| \leq n c|x-y| . \tag{2.71}
\end{equation*}
$$

Proof. Consider any $x, y \in U$. The Mean Value Theorem says that for every $i$ there exists a point $c$ on the line joining $x$ to $y$ such that

$$
\begin{equation*}
f_{i}(x)-f_{i}(y)=\sum_{j} \frac{\partial f_{i}}{\partial x_{j}}(d)\left(x_{j}-y_{j}\right) \tag{2.72}
\end{equation*}
$$

It follows that

$$
\begin{align*}
\left|f_{i}(x)-f_{i}(y)\right| & \leq \sum_{j}\left|\frac{\partial f_{i}}{\partial x_{j}}(d)\right|\left|x_{j}-y_{j}\right| \\
& \leq \sum_{j} c\left|x_{j}-y_{j}\right|  \tag{2.73}\\
& \leq n c|x-y|
\end{align*}
$$

This is true for each $i$, so $|f(x)-f(y)| \leq n c|x-y|$
Lemma 2.15. Let $U$ be open in $\mathbb{R}^{n}$ and $f: U \rightarrow \mathbb{R}$ a $\mathcal{C}^{1}$ map. Suppose $f$ takes $a$ minimum value at some point $b \in U$. Then

$$
\begin{equation*}
\frac{\partial f}{\partial x_{i}}(b)=0, i=1, \ldots, n . \tag{2.74}
\end{equation*}
$$

Proof. We reduce to the one-variable result. Let $b=\left(b_{1}, \ldots, b_{n}\right)$ and let $\phi(t)=$ $f\left(b_{1}, \ldots, b_{i-1}, t, b_{i+1}, \ldots, b_{n}\right)$, which is $\mathcal{C}^{1}$ near $b_{1}$ and has a minimum at $b_{i}$. We know from one-variable calculus that this implies that $\frac{\partial \phi}{\partial t}\left(b_{i}\right)=0$.

In our proof of the Inverse Function Theorem, we want to show that $f$ is locally a diffeomorphism at $a$. We will make the following simplifying assumptions:

$$
\begin{equation*}
a=0, \quad f(a)=0, \quad D f(0)=I \text { (identity). } \tag{2.75}
\end{equation*}
$$

Then, we define a map $g: U \rightarrow \mathbb{R}^{n}$ by $g(x)=x-f(x)$, so that we obtain the further simplification

$$
\begin{equation*}
D g(0)=D f(0)-I=0 \tag{2.76}
\end{equation*}
$$

Lemma 2.16. Given $\epsilon>0$, there exists $\delta>0$ such that for any $x, y \in R_{\delta}$,

$$
\begin{equation*}
\mid g(x-g(y)|<\epsilon| x-y \mid . \tag{2.77}
\end{equation*}
$$

Proof. The result that $D g(0)=0$ implies that there exists $\delta>0$ such that for any $x \in R_{\delta},|D g(x)| \leq \epsilon / n$. Applying the first lemma, the proof is complete.

Now, remember that $g(x)=x-f(x)$. Take any $x, y \in R_{\delta}$. Then

$$
\begin{align*}
x-y & =x-f(x)+f(x)-f(y)+f(y)-y  \tag{2.78}\\
& =g(x)-g(y)+f(x)-f(y) .
\end{align*}
$$

Using the Triangle Inequality we obtain

$$
\begin{equation*}
|x-y| \leq|g(x)-g(y)|+|f(x)-f(y)| \tag{2.79}
\end{equation*}
$$

Using the previous lemma, we find that

$$
\begin{equation*}
(1-\epsilon)|x-y| \leq|f(x)-f(y)| . \tag{2.80}
\end{equation*}
$$

We choose $\delta$ such that $\epsilon>1 / 2$, so that

$$
\begin{equation*}
|x-y| \leq 2|f(x)-f(y)| . \tag{2.81}
\end{equation*}
$$

This proves that $f: R_{\delta} \rightarrow \mathbb{R}^{n}$ is one-to-one.
We also want to prove that $f$ is onto. We have $D f(0)=I$, so $\operatorname{det}\left(\frac{\partial f_{i}}{\partial x_{j}}(0)\right)=1$. We can choose $\delta$ such that for any $x \in R_{\delta}$,

$$
\begin{equation*}
\operatorname{det}\left(\frac{\partial f_{i}}{\partial x_{j}}(x)\right)>\frac{1}{2} \tag{2.82}
\end{equation*}
$$

Lemma 2.17. If $y \in B_{\delta / 4}$, than there exists a point $c \in R_{\delta}$ such that $f(c)=y$.
Proof. Let $h: \bar{R}_{\delta} \rightarrow \mathbb{R}$ be a map defined by $h(x)=\|f(x)-y\|^{2}$. The domain $\bar{R}_{\delta}$ is compact, so $h$ has a minimum at some point $c \in R_{\delta}$.

Claim. The point $c$ is an interior point. That is, $c \in R_{\delta}$.
Proof. For any $x \in \bar{R}_{\delta},|x|=\delta$ implies that $|f(x)-f(0)|=|f(x)| \geq \delta / 2$

$$
\begin{align*}
& \Longrightarrow\|f(x)\| \geq \frac{\delta}{2} \\
& \Longrightarrow\|f(x)-y\| \geq \frac{\delta}{4}, \text { when } x \in \operatorname{Bd} R_{\delta} .  \tag{2.83}\\
& \Longrightarrow h(x) \geq\left(\frac{\delta}{4}\right)^{2} .
\end{align*}
$$

At the origin, $h(0)=\|f(0)-y\|^{2}=\|y\|^{2}<(\delta / 4)^{2}$, since $y \in B_{\delta / 4}$. So, $h(0) \leq h$ on $\mathrm{Bd} R_{\delta}$, which means that the minimum point $c$ of $h$ is in $R_{\delta}$. This ends the proof of the claim.

Now that we know that the minimum point $c$ occurs in the interior, we can apply the second lemma to $h$ to obtain

$$
\begin{equation*}
\frac{\partial h}{\partial x_{j}}(c)=0, \quad j=1, \ldots, n \tag{2.84}
\end{equation*}
$$

From the definition of $h$,

$$
\begin{equation*}
h(x)=\sum_{i=1}^{n}\left(f_{i}(c)-y_{i}\right) \frac{\partial f_{i}}{\partial x_{j}}(c)=0, i=1, \ldots, n \tag{2.85}
\end{equation*}
$$

SO

$$
\begin{equation*}
\frac{\partial h}{\partial x_{j}}(c)=2 \sum_{i=1}^{n}\left(f_{i}(c)-y_{i}\right) \frac{\partial f_{i}}{\partial x_{j}}(c)=0, i=1, \ldots, n \tag{2.86}
\end{equation*}
$$

Note that

$$
\begin{equation*}
\operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}(c)\right] \neq 0 \tag{2.87}
\end{equation*}
$$

so, by Cramer's Rule,

$$
\begin{equation*}
f_{i}(c)-y_{i}=0, i=1, \ldots, n \tag{2.88}
\end{equation*}
$$

Let $U_{1}=R_{\delta} \sim f^{-1}\left(B_{\delta / 4}\right)$, where we have chosen $V=B_{\delta / 4}$. We have shown that $f$ is a bijective map.

Claim. The map $f^{-1}: V \rightarrow U_{1}$ is continuous.
Proof. Let $a, b \in V$, and define $x=f^{-1}(a)$ and $y=f^{-1}(b)$. Then $a=f(x)$ and $b=f(y)$.

$$
\begin{equation*}
|a-b|=|f(x)-f(y)| \geq \frac{\partial|x-y|}{\partial 2} \tag{2.89}
\end{equation*}
$$

so

$$
\begin{equation*}
|a-b| \geq \frac{1}{2}\left|f^{-1}(a)-f^{-1}(b)\right| . \tag{2.90}
\end{equation*}
$$

This shows that $f^{-1}$ is continuous on $V=B_{\delta / 4}$.
As a last item for today's lecture, we show the following:
Claim. The map $f^{-1}$ is differentiable at 0 , and $D f^{-1}(0)=I$.
Proof. Let $k \in \mathbb{R}^{n}-\{0\}$ and choose $k \doteq 0$. We are trying to show that

$$
\begin{equation*}
\frac{f^{-1}(0+k)-f^{-1}(0)-D f^{-1}(0) k}{|k|} \rightarrow 0 \text { as } k \rightarrow 0 . \tag{2.91}
\end{equation*}
$$

We simplify

$$
\begin{equation*}
\frac{f^{-1}(0+k)-f^{-1}(0)-D f^{-1}(0) k}{|k|}=\frac{f^{-1}(k)-k}{|k|} . \tag{2.92}
\end{equation*}
$$

Define $h=f^{-1}(k)$ so that $k=f(h)$ and $|k| \leq 2|h|$. To show that

$$
\begin{equation*}
\frac{f^{-1}(k)-k}{|k|} \rightarrow 0 \text { as } k \rightarrow 0, \tag{2.93}
\end{equation*}
$$

it suffices to show that

$$
\begin{equation*}
\frac{f^{-1}(k)-k}{|h|} \rightarrow 0 \text { as } h \rightarrow 0 . \tag{2.94}
\end{equation*}
$$

That is, it suffices to show that

$$
\begin{equation*}
\frac{h-f(h)}{|h|} \rightarrow 0 \text { as } h \rightarrow 0 \tag{2.95}
\end{equation*}
$$

But this is equal to

$$
\begin{equation*}
-\frac{f(h)-f(0)-D f(0) h}{|h|}, \tag{2.96}
\end{equation*}
$$

which goes to zero as $h \rightarrow 0$ because $f$ is differentiable at zero.
The proof of the Inverse Function Theorem continues in the next lecture.

## Lecture 7

We continue our proof of the Inverse Function Theorem.
As before, we let $U$ be an open set in $\mathbb{R}^{n}$, and we assume that $0 \in U$. We let $f: U \rightarrow \mathbb{R}^{n}$ be a $\mathcal{C}^{1}$ map, and we assume $f(0)=0$ and that $D f(0)=I$. We summarize what we have proved so far in the following theorem.
Theorem 2.18. There exists a neighborhood $U_{0}$ of 0 in $U$ and a neighborhood $V$ of 0 in $\mathbb{R}^{n}$ such that

1. f maps $U_{0}$ bijectively onto $V$
2. $f^{-1}: V \rightarrow U_{0}$ is continuous,
3. $f^{-1}$ is differentiable at 0 .

Now, we let $U$ be an open set in $\mathbb{R}^{n}$, and we let $f: U \rightarrow \mathbb{R}^{n}$ be a $\mathcal{C}^{2}$ map, as before, but we return to our original assumptions that $a \in U, b=f(a)$, and $\operatorname{Df}(a): \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is bijective. We prove the following theorem.
Theorem 2.19. There exists a neighborhood $U_{0}$ of a in $U$ and a neighborhood $V$ of $b$ in $\mathbb{R}^{n}$ such that

1. $f$ maps $U_{0}$ bijectively onto $V$
2. $f^{-1}: V \rightarrow U_{0}$ is continuous,
3. $f^{-1}$ is differentiable at $b$.

Proof. The map $f: U \rightarrow \mathbb{R}^{n}$ maps $a$ to $b$. Define $U^{\prime}=U-a=\{x-a: x \in U\}$. Also define $f_{1}: U^{\prime} \rightarrow \mathbb{R}^{n}$ by $f_{1}(x)=f(x+a)-b$, so that $f_{1}(0)=0$ and $D f_{1}(0)=D f(a)$ (using the Chain Rule).

Let $A=D f(a)=D f_{1}(0)$. We know that $A$ is invertible.
Now, define $f_{2}: U^{\prime} \rightarrow \mathbb{R}^{n}$ by $f_{2}=A^{-1} f_{1}$, so that $f_{2}(0)=0$ and $D f_{2}(0)=I$. The results from last lecture show that the theorem at hand is true for $f_{2}$. Because $f_{1}=A \circ f_{2}$, the theorem is also true for $f_{1}$. Finally, because $f(x)=f_{1}(x-a)+b$, the theorem is true for $f$.

So, we have a bijective map $f: U_{0} \rightarrow V$. Let us take $c \in U_{0}$ and look at the derivative

$$
\begin{equation*}
D f(c) \sim\left[\frac{\partial f_{i}}{\partial x_{j}}(c)\right]=J_{f}(c) \tag{2.97}
\end{equation*}
$$

Note that

$$
\begin{equation*}
D f(c) \text { is bijective } \Longleftrightarrow \operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}(c)\right] \neq 0 \tag{2.98}
\end{equation*}
$$

Because $f$ is $\mathcal{C}^{1}$, the functions $\frac{\partial f_{i}}{\partial x_{j}}$ are continuous on $U_{0}$. If $\operatorname{det} J_{f}(a) \neq 0$, then $\operatorname{det} J_{f}(c) \neq 0$ for $c$ close to $a$. We can shrink $U_{0}$ and $V$ such that $\operatorname{det} J_{f}(c) \neq 0$ for
all $c \in U_{0}$, so for every $c \in U_{0}$, the map $f^{-1}$ is differentiable at $f(c)$. That is, $f^{-1}$ is differentiable at all points of $V$.

We have thus improved the previous theorem. We can replace the third point with

$$
\begin{equation*}
\text { 3. } f^{-1} \text { is differentiable at all points of } V \text {. } \tag{2.99}
\end{equation*}
$$

Let $f^{-1}=g$, so that $g \circ f=$ identity map. The Chain Rule is used to show the following. Suppose $p \in U_{0}$ and $q=f(p)$. Then $D g(q)=D f(p)^{-1}$, so $J_{g}(q)=J_{f}(p)^{-1}$. That is, for all $x \in V$,

$$
\begin{equation*}
\left[\frac{\partial g_{i}}{\partial x_{j}}(x)\right]=\left[\frac{\partial f_{i}}{\partial x_{j}}(g(x))\right]^{-1} \tag{2.100}
\end{equation*}
$$

The function $f$ is $\mathcal{C}^{1}$, so $\frac{\partial f_{i}}{\partial x_{j}}$ is continuous on $U_{0}$. It also follows that $g$ is continuous, so $\frac{\partial f_{i}}{\partial x_{j}}(g(x))$ is continuous on $V$.

Using Cramer's Rule, we conclude that the entries of matrix on the r.h.s. of Equation 2.100 are continuous functions on $V$. This shows that $\frac{\partial f_{i}}{\partial x_{j}}$ is continuous on $V$, which implies that $g$ is a $\mathcal{C}^{1}$ map.

We leave as an exercise to show that $f \in \mathcal{C}^{r}$ implies that $g \in \mathcal{C}^{r}$ for all $r$. The proof is by induction.

This concludes the proof of the Inverse Function Theorem, signifying the end of this section of the course.

## 3 Integration

### 3.1 Riemann Integral of One Variable

We now begin to study the next main topic of this course: integrals. We begin our discussion of integrals with an 18.100 level review of integrals.

We begin by defining the Riemann integral (sometimes written in shorthand as the R. integral).

Let $[a, b] \subseteq \mathbb{R}$ be a closed interval in $\mathbb{R}$, and let $P$ be a finite subset of $[a, b]$. Then $P$ is a partition if $a, b \in P$ and if all of the elements $t_{i}, \ldots, t_{N}$ in $P$ can be arranged such that $t_{1}=a<t_{2}<\cdots<t_{n}=b$. We define $I_{i}=\left[t_{i}, t_{i+1}\right]$, which are called the subintervals of $[a, b]$ belonging to $P$.

Let $f:[a, b] \rightarrow \mathbb{R}$ be a bounded function, and let $I_{i}$ be a subinterval belonging to $P$. Then we define

$$
\begin{align*}
& m_{i}=\inf f: I_{i} \rightarrow \mathbb{R} \\
& M_{i}=\sup f: I_{i} \rightarrow \mathbb{R} \tag{3.1}
\end{align*}
$$

from which we define the lower and upper Riemann sums

$$
\begin{align*}
L(f, P) & =\sum_{i} m_{i} \times\left(\text { length of } I_{i}\right) \\
U(f, P) & =\sum_{i} M_{i} \times\left(\text { length of } I_{i}\right), \tag{3.2}
\end{align*}
$$

respectively.
Clearly,

$$
\begin{equation*}
L(f, P) \leq U(f, P) \tag{3.3}
\end{equation*}
$$

Now, let $P$ and $P^{\prime}$ be partitions.
Definition 3.1. The partition $P$ is a refinement of $P$ if $P^{\prime} \supset P$.
If $P^{\prime}$ is a refinement of $P$, then

$$
\begin{align*}
L\left(f, P^{\prime}\right) & \geq L(f, P), \text { and }  \tag{3.4}\\
U\left(f, P^{\prime}\right) & \leq U(f, P)
\end{align*}
$$

If $P$ and $P^{\prime}$ are any partitions, then you can take $P^{\prime \prime}=P \cup P^{\prime}$, which is a refinement of both $P$ and $P^{\prime}$. So,

$$
\begin{equation*}
L(f, P) \leq L\left(f, P^{\prime \prime}\right) \leq U\left(f, P^{\prime \prime}\right) \leq U\left(f, P^{\prime}\right) \tag{3.5}
\end{equation*}
$$

for any partitions $P, P^{\prime}$. That is, the lower Riemann sum is always less than or equal to the upper Riemann sum, regardless of the partitions used.

Now we can define the Lower and Upper Riemann integrals

$$
\begin{align*}
& \int_{[a, b]} f=\text { l.u.b. }\{L(f, P) \mid P \text { a partition of }[a, b]\} \\
& \bar{\int}_{[a, b]} f=\text { g.l.b. }\{U(f, P) \mid P \text { a partition of }[a, b]\} \tag{3.6}
\end{align*}
$$

We can see from the above that

$$
\begin{equation*}
\int f \leq \bar{\int} f \tag{3.7}
\end{equation*}
$$

Claim. If $f$ is continuous, then

$$
\begin{equation*}
\underline{\int} f=\bar{\int} f . \tag{3.8}
\end{equation*}
$$

Definition 3.2. For any bounded function $f:[a, b] \rightarrow \mathbb{R}$, the function $f$ is (Riemann) integrable if

$$
\begin{equation*}
\underline{\mathcal{S}}_{[a, b]} f=\bar{\int}_{[a, b]} f . \tag{3.9}
\end{equation*}
$$

In the next lecture we will begin to generalize these notions to multiple variables.

## Lecture 8

### 3.2 Riemann Integral of Several Variables

Last time we defined the Riemann integral for one variable, and today we generalize to many variables.

Definition 3.3. A rectangle is a subset $Q$ of $\mathbb{R}^{n}$ of the form

$$
\begin{equation*}
Q=\left[a_{1}, b_{1}\right] \times \cdots \times\left[a_{n}, b_{n}\right], \tag{3.10}
\end{equation*}
$$

where $a_{i}, b_{i} \in \mathbb{R}$.
Note that $x=\left(x_{1}, \ldots, x_{n}\right) \in Q \Longleftrightarrow a_{i} \leq x_{i} \leq b_{i}$ for all $i$. The volume of the rectangle is

$$
\begin{equation*}
v(Q)=\left(b_{1}-a_{1}\right) \cdots\left(b_{n}-a_{n}\right), \tag{3.11}
\end{equation*}
$$

and the width of the rectangle is

$$
\begin{equation*}
\operatorname{width}(Q)=\sup _{i}\left(b_{i}-a_{i}\right) \tag{3.12}
\end{equation*}
$$

Recall (stated informally) that given $[a, b] \in \mathbb{R}$, a finite subset $P$ of $[a, b]$ is a partition of $[a, b]$ if $a, b \in P$ and you can write $P=\left\{t_{i}: i=1, \ldots, N\right\}$, where $t_{1}=a<t_{2}<\ldots<t_{N}=b$. An interval $I$ belongs to $P$ if and only if $I$ is one of the intervals $\left[t_{i}, t_{i+1}\right]$.

Definition 3.4. A partition $P$ of $Q$ is an $n$-tuple $\left(P_{1}, \ldots, P_{n}\right)$, where each $P_{i}$ is a partition of $\left[a_{i}, b_{i}\right]$.

Definition 3.5. A rectangle $R=I_{1} \times \cdots \times I_{n}$ belongs to $P$ if for each $i$, the interval $I_{i}$ belongs to $P_{i}$.

Let $f: Q \rightarrow \mathbb{R}$ be a bounded function, let $P$ be a partition of $Q$, and let $R$ be a rectangle belonging to $P$.

We define

$$
\begin{align*}
& m_{R} f=\inf _{R} f=\text { g.l.b. }\{f(x): x \in \mathbb{R}\} \\
& M_{R} f=\sup _{R} f=\text { l.u.b. }\{f(x): x \in \mathbb{R}\} \tag{3.13}
\end{align*}
$$

from which we define the lower and upper Riemann sums,

$$
\begin{align*}
L(f, P) & =\sum_{R} m_{R}(f) v(R) \\
U(f, P) & =\sum_{R} M_{R}(f) v(R) . \tag{3.14}
\end{align*}
$$

It is evident that

$$
\begin{equation*}
L(f, P) \leq U(f, P) \tag{3.15}
\end{equation*}
$$

Now, we will take a sequence of partitions that get finer and finer, and we will define the integral to be the limit.

Let $P=\left(P_{1}, \ldots, P_{n}\right)$ and $P^{\prime}=\left(P_{1}^{\prime}, \ldots, P_{n}^{\prime}\right)$ be partitions of $Q$. We say that $P^{\prime}$ refines $P$ if $P_{i}^{\prime} \supset P_{i}$ for each $i$.

Claim. If $P^{\prime}$ refines $P$, then

$$
\begin{equation*}
L\left(f, P^{\prime}\right) \geq L(f . P) \tag{3.16}
\end{equation*}
$$

Proof. We let $P_{j}=P_{j}^{\prime}$ for $j \neq i$, and we let $P_{i}^{\prime}=P_{i} \cup\{a\}$, where $a \in\left[a_{i}, b_{i}\right]$. We can create any refinement by multiple applications of this basic refinement. If $R$ is a rectangle belonging to $P$, then either

1. $R$ belongs to $P^{\prime}$, or
2. $R=R^{\prime} \cup R^{\prime \prime}$, where $R^{\prime}, R^{\prime \prime}$ belong to $P^{\prime}$.

In the first case, the contribution of $R$ to $L\left(f, P^{\prime}\right)$ equals the contribution of $R$ to $L(f, P)$, so the claim holds.

In the second case,

$$
\begin{equation*}
m_{R} v(R)=m_{R}\left(v\left(R^{\prime}\right)+v\left(R^{\prime \prime}\right)\right) \tag{3.17}
\end{equation*}
$$

and

$$
\begin{equation*}
m_{r}=\inf _{R} f \leq \inf _{R^{\prime}} f, \inf _{R^{\prime \prime}} f \tag{3.18}
\end{equation*}
$$

So,

$$
\begin{equation*}
m_{R} \leq m_{R^{\prime}}, m_{R^{\prime \prime}} \tag{3.19}
\end{equation*}
$$

Taken altogether, this shows that

$$
\begin{equation*}
m_{R} v(R) \leq m_{R^{\prime}} v\left(R^{\prime}\right)+m_{R^{\prime \prime}} v\left(R^{\prime \prime}\right) \tag{3.20}
\end{equation*}
$$

Thus, $R^{\prime}$ and $R^{\prime \prime}$ belong to $P^{\prime}$.
Claim. If $P^{\prime}$ refines $P$, then

$$
\begin{equation*}
U\left(f, P^{\prime}\right) \leq U(f, P) \tag{3.21}
\end{equation*}
$$

The proof is very similar to the previous proof. Combining the above two claims, we obtain the corollary

Corollary 2. If $P$ and $P^{\prime}$ are partitions, then

$$
\begin{equation*}
U\left(f, P^{\prime}\right) \geq L(f, P) \tag{3.22}
\end{equation*}
$$

Proof. Define $P^{\prime \prime}=\left(P_{1} \cup P_{1}^{\prime}, \ldots, P_{n} \cup P_{n}^{\prime}\right)$. So, $P^{\prime \prime}$ refines $P$ and $P^{\prime}$. We have shown that

$$
\begin{align*}
U\left(f, P^{\prime \prime}\right) & \leq U(f, P) \\
L\left(f, P^{\prime}\right) & \leq L\left(f, P^{\prime \prime}\right)  \tag{3.23}\\
L\left(f, P^{\prime \prime}\right) & \leq U\left(f, P^{\prime \prime}\right)
\end{align*}
$$

Together, these show that

$$
\begin{equation*}
U(f, P) \geq L\left(f, P^{\prime}\right) \tag{3.24}
\end{equation*}
$$

With this result in mind, we define the lower and upper Riemann integrals:

$$
\begin{align*}
& {\underline{\int_{Q}}} f=\sup _{P} L(f, P)  \tag{3.25}\\
& \overline{\int_{Q}} f=\inf _{P} U(f, P)
\end{align*}
$$

Clearly, we have

$$
\begin{equation*}
\underline{\int}_{Q} f \leq \bar{\int}_{Q} f \tag{3.26}
\end{equation*}
$$

Finally, we define Riemann integrable.
Definition 3.6. A function $f$ is Riemann integrable over $Q$ if the lower and upper Riemann integrals coincide (are equal).

### 3.3 Conditions for Integrability

Our next problem is to determine under what conditions a function is (Riemann) integrable.

Let's look at a trivial case:
Claim. Let $F: Q \rightarrow \mathbb{R}$ be the constant function $f(x)=c$. Then $f$ is $R$. integrable over $Q$, and

$$
\begin{equation*}
\int_{Q} c=c v(Q) \tag{3.27}
\end{equation*}
$$

Proof. Let $P$ be a partition, and let $R$ be a rectangle belonging to $P$. We see that $m_{R}(f)=M_{R}(f)=c$, so

$$
\begin{align*}
U(f, P) & =\sum_{R} M_{R}(f) v(R)=c \sum_{R} v(R)  \tag{3.28}\\
& =c v(Q) .
\end{align*}
$$

Similarly,

$$
\begin{equation*}
L(f, P)=c v(Q) \tag{3.29}
\end{equation*}
$$

Corollary 3. Let $Q$ be a rectangle, and let $\left\{Q_{i}: i=1, \ldots, N\right\}$ be a collection of rectangles covering $Q$. Then

$$
\begin{equation*}
v(Q) \leq \sum v\left(Q_{i}\right) \tag{3.30}
\end{equation*}
$$

Theorem 3.7. If $f: Q \rightarrow \mathbb{R}$ is continuous, then $f$ is $R$. integrable over $Q$.
Proof. We begin with a definition
Definition 3.8. Given a partition $P$ of $Q$, we define

$$
\begin{equation*}
\operatorname{mesh} \operatorname{width}(P)=\sup _{R} \operatorname{width}(R) . \tag{3.31}
\end{equation*}
$$

Remember that

$$
\begin{equation*}
Q \text { compact } \Longrightarrow f: Q \rightarrow \mathbb{R} \text { is uniformly continuous. } \tag{3.32}
\end{equation*}
$$

That is, given $\epsilon>0$, there exists $\delta>0$ such that if $x, y \in Q$ and $|x-y|<\delta$, then $|f(x)-f(y)|<\epsilon$.

Choose a partition $P$ of $Q$ with mesh width less than $\delta$. Then, for every rectangle $R$ belonging to $P$ and for every $x, y \in R$, we have $|x-y|<\delta$. By uniform continuity we have, $M_{R}(f)-m_{R}(f) \leq \epsilon$, which is used to show that

$$
\begin{align*}
U(f, P)-L(f, P) & =\sum_{R}\left(M_{R}(f)-m_{R}(f)\right) v(R) \\
& \leq \epsilon \sum v(R)  \tag{3.33}\\
& \leq \epsilon v(Q) .
\end{align*}
$$

We can take $\epsilon \rightarrow 0$, so

$$
\begin{equation*}
\sup _{P} L(f, P)=\inf _{P} U(f, P), \tag{3.34}
\end{equation*}
$$

which shows that $f$ is integrable.
We have shown that continuity is sufficient for integrability. However, continuity is clearly not necessary. What is the general condition for integrability? To state the answer, we need the notion of measure zero.

Definition 3.9. Suppose $A \subseteq \mathbb{R}^{n}$. The set $A$ is of measure zero if for every $\epsilon>$ 0 , there exists a countable covering of $A$ by rectangles $Q_{1}, Q_{2}, Q_{3}, \ldots$ such that $\sum_{i} v\left(Q_{i}\right)<\epsilon$.

Theorem 3.10. Let $f: Q \rightarrow \mathbb{R}$ be a bounded function, and let $A \subseteq Q$ be the set of points where $f$ is not continuous. Then $f$ is $R$. integrable if and only if $A$ is of measure zero.

Before we prove this, we make some observations about sets of measure zero:

1. Let $A, B \subseteq \mathbb{R}^{n}$ and suppose $B \subset A$. If $A$ is of measure zero, then $B$ is also of measure zero.
2. Let $A_{i} \subseteq \mathbb{R}^{n}$ for $i=1,2,3, \ldots$, and suppose the $A_{i}$ 's are of measure zero. Then $\cup A_{i}$ is also of measure zero.
3. Rectangles are not of measure zero.

We prove the second observation:
For any $\epsilon>0$, choose coverings $Q_{i, 1}, Q_{i, 2}, \ldots$ of $A_{i}$ such that each covering has total volume less than $\epsilon / 2^{i}$. Then $\left\{Q_{i, j}\right\}$ is a countable covering of $\cup A_{i}$ of total volume

$$
\begin{equation*}
\sum_{i=1}^{\infty} \frac{\epsilon}{2^{i}}=\epsilon \tag{3.35}
\end{equation*}
$$

## Lecture 9

We quickly review the definition of measure zero.
A set $A \subseteq \mathbb{R}^{n}$ is of measure zero if for every $\epsilon>0$, there exists a covering of $A$ by rectangles $Q_{1}, Q_{2}, Q_{3}, \ldots$ such that the total volume $\sum v\left(Q_{i}\right)<\epsilon$.

Remark. In this definition we can replace "rectangles" by "open rectangles." To see this, given any $\epsilon>0$ let $Q_{1}, Q_{2}, \ldots$ be a cover of $A$ with volume less than $\epsilon / 2$. Next, choose $Q_{i}^{\prime}$ to be rectangles such that Int $Q_{i}^{\prime} \supset Q_{i}$ and $v\left(Q_{i}^{\prime}\right)<2 v\left(Q_{i}\right)$. Then Int $Q_{1}^{\prime}$, Int $Q_{2}^{\prime}, \ldots$ cover $A$ and have total volume less than $\epsilon$.

We also review the three properties of measure zero that we mentioned last time, and we prove the third.

1. Let $A, B \subseteq \mathbb{R}^{n}$ and suppose $B \subset A$. If $A$ is of measure zero, then $B$ is also of measure zero.
2. Let $A_{i} \subseteq \mathbb{R}^{n}$ for $i=1,2,3, \ldots$, and suppose the $A_{i}$ 's are of measure zero. Then $\cup A_{i}$ is also of measure zero.
3. Rectangles are not of measure zero.

We prove the third property:
Claim. If $Q$ is a rectangle, then $Q$ is not of measure zero.
Proof. Choose $\epsilon<v(Q)$. Suppose $Q_{1}, Q_{2}, \ldots$ are rectangles such that the total volume is less than $\epsilon$ and such that Int $Q_{1}$, Int $Q_{2}, \ldots$ cover $Q$.

The set $Q$ is compact, so the H-B Theorem implies that the collection of sets Int $Q_{1}, \ldots$, Int $Q_{N}$ cover $Q$ for $N$ sufficiently large. So,

$$
\begin{equation*}
Q \subseteq \bigcup_{i=1}^{N} Q_{i} \tag{3.36}
\end{equation*}
$$

which implies that

$$
\begin{equation*}
v(Q) \leq \sum_{i=1}^{N} v\left(Q_{i}\right)<\epsilon<v(Q) \tag{3.37}
\end{equation*}
$$

which is a contradiction.
We then have the following simple result.
Claim. If Int $A$ is non-empty, then $A$ is not of measure zero.
Proof. Consider any $p \in \operatorname{Int} A$. There exists a $\delta>0$ such that $U(p, \delta)=\{x:|x-p|<$ $\delta\}$ is contained in $A$. Then let $Q=\{x:|x-p| \leq \delta\}$. It follows that if $A$ is of measure zero, then $Q$ is of measure zero, by the first property. We know that $Q$ is not of measure zero by the third property.

We restate the necessary and sufficient condition for $R$. integrability from last time, and we now prove the theorem.

Theorem 3.11. Let $Q$ be a rectangle and $f: Q \rightarrow \mathbb{R}$ be a bounded function. Let $D$ be the set of points in $Q$ where $f$ is not continuous. Then $f$ is $R$. integrable if and only if $D$ is of measure zero.

Proof. First we show that

$$
\begin{equation*}
D \text { is of measure zero } \Longrightarrow f \text { is R. integrable } \tag{3.38}
\end{equation*}
$$

Lemma 3.12. Let $Q=\left[a_{1}, b_{1}\right] \times \cdots \times\left[a_{n}, b_{n}\right]$, and let $Q^{\alpha}, \alpha=1, \ldots, N$, be a covering of $Q$ by rectangles. Then there exists a partition $P$ of $Q$ such that every rectangle $R$ belonging to $P$ is contained in $Q^{\alpha}$ for some $\alpha$.

Proof. Write out $Q^{\alpha}=I_{1}^{\alpha} \times \cdots \times I_{n}^{\alpha}$, and let

$$
\begin{equation*}
P_{j}=\left(\bigcup_{\alpha} \text { Endpoints of } I_{j}^{\alpha}\right) \cap\left[a_{j}, b_{j}\right] \cup\left\{a_{j}, b_{j}\right\} . \tag{3.39}
\end{equation*}
$$

One can show that $P_{j}$ is a partition of $\left[a_{j}, b_{j}\right]$, and $P=\left(P_{1}, \ldots, P_{n}\right)$ is a partition of $Q$ with the above properties.

Let $f: Q \rightarrow \mathbb{R}$ be a bounded function, and let $D$ be the set of points at which $f$ is discontinuous. Assume that $D$ is of measure zero. We want to show that $f$ is R . integrable.

Let $\epsilon>0$, and let $Q_{i}^{\prime}, i=1,2,3, \ldots$ be a collection of rectangles of total volume less than $\epsilon$ such that Int $Q_{1}^{\prime}, Q_{2}^{\prime}, \ldots$ cover $D$.

If $p \in Q-D$, we know that $f$ is continuous at $p$. So, there exists a rectangle $Q_{p}$ with $p \in \operatorname{Int} Q_{p}$ and $|f(x)-f(p)|<\epsilon / 2$ for all $x \in Q_{p}$ (for example, $Q_{p}=\{x| | x-p \mid \leq \delta\}$ for some $\delta$ ). Given any $x, y \in Q_{p}$, we find that $|f(x)-f(y)|<\epsilon$.

The rectangles Int $Q_{p}, p \in Q-D$ along with the rectangles Int $Q_{i}^{\prime}, i=1,2, \ldots$ cover $Q$. The set $Q$ is compact, so the H-B Theorem implies that there exists a finite open subcover:

$$
\begin{equation*}
Q_{i} \equiv \operatorname{Int} Q_{p_{i}}, i=1, \ldots, \ell ; \quad \operatorname{Int} Q_{j}^{\prime}, j=1, \ldots, \ell \tag{3.40}
\end{equation*}
$$

Using the lemma, there exists a partition $P$ of $Q$ such that every rectangle belonging to $P$ is contained in a $Q_{i}$ or a $Q_{j}^{\prime}$.

We now show that $f$ is R . integrable.

$$
\begin{align*}
U(f, P)-L(f, P) & =\sum_{R}\left(M_{R}(f)-m_{R}(f)\right) v(R) \\
& +\sum_{R^{\prime}}\left(M_{R^{\prime}}(f)-m_{R^{\prime}}(f)\right) v\left(R^{\prime}\right) \tag{3.41}
\end{align*}
$$

where each $R$ in the first sum belongs to a $Q_{i}$, and each $R^{\prime}$ in the second sum belongs to a $Q_{j}^{\prime}$.

We look at the first sum. If $x, y \in R \subseteq Q_{i}$, then $|f(x)-f(y)| \leq \epsilon$. So, $M_{R}(f)-$ $m_{R}(f) \leq \epsilon$. It follows that

$$
\begin{align*}
\sum_{R}\left(M_{R}(f)-m_{R}(f)\right) v(R) & \leq \epsilon \sum_{R} v(R)  \tag{3.42}\\
& \leq \epsilon v(Q) .
\end{align*}
$$

We now look at the second sum. The function $f: Q \rightarrow \mathbb{R}$ is bounded, so there exists a number $c$ such that $-c \leq f(x) \leq c$ for all $x \in Q$. Then, $M_{R^{\prime}}(f)-m_{R^{\prime}}(f) \leq 2 c$ so

$$
\begin{align*}
\sum_{R^{\prime}}\left(M_{R^{\prime}}(f)-f_{R^{\prime}}(f)\right) v\left(R^{\prime}\right) & \leq 2 c \sum_{R^{\prime}} v\left(R^{\prime}\right) \\
& =2 c \sum_{i=1}^{\ell} \sum_{R^{\prime} \subseteq Q_{i}^{\prime}} v\left(R^{\prime}\right)  \tag{3.43}\\
& \leq 2 c \sum_{i} v\left(Q_{i}^{\prime}\right) \\
& \leq 2 c \epsilon .
\end{align*}
$$

Substituting back into Equation 3.41, we get

$$
\begin{equation*}
U(f, P)-L(f, P) \leq \epsilon(v(Q)+2 c) \tag{3.44}
\end{equation*}
$$

So,

$$
\begin{equation*}
\overline{\int_{Q}} f-\underline{\int}_{Q} f \leq \epsilon(v(Q)+2 c) \tag{3.45}
\end{equation*}
$$

because

$$
\begin{equation*}
U(f, P) \geq \int_{Q} f \text { and } L(f, P) \leq{\underline{\int_{Q}}}_{Q} f \tag{3.46}
\end{equation*}
$$

Letting $\epsilon$ go to zero, we conclude that

$$
\begin{equation*}
\bar{\int}_{Q} f={\underline{\int_{Q}}}_{Q} f \tag{3.47}
\end{equation*}
$$

which shows that $f$ is Riemann integrable.
This concludes the proof in one direction. We do not prove the other direction.
Corollary 4. Suppose $f: Q \rightarrow \mathbb{R}$ is $R$. integrable and that $f \geq 0$ everywhere. If $\int_{Q} f=0$, then $f=0$ except on a set of measure zero.

Proof. Let $D$ be the set of points where $f$ is discontinuous. The function $f$ is R . integrable, so $D$ is of measure zero.

If $p \in Q-D$, then $f(p)=0$. To see this, suppose that $f(p)=\delta>0$. The function $f$ is continuous at $p$, so there exists a rectangle $R_{0}$ centered at $p$ such that $f \geq n \delta / 2$ on $R_{0}$. Choose a partition $P$ such that $R_{0}$ is a rectangle belonging to $P$. On any rectangle $R$ belonging to $P, f \geq 0$, so $m_{R}(f) \geq 0$. This shows that

$$
\begin{align*}
L(f, P) & =m_{R_{0}}(f) v\left(R_{0}\right)+\sum_{R \neq R_{0}} m_{R}(f) v(R)  \tag{3.48}\\
& \geq \frac{\delta}{2} v\left(R_{0}\right)+0 .
\end{align*}
$$

But we assumed that $\int_{Q} f=0$, so we have reached a contradiction. So $f=0$ at all points $p \in Q-D$.

## Lecture 10

We begin today's lecture with a simple claim.
Claim. Let $Q \subseteq \mathbb{R}^{n}$ be a rectangle and $f, g: Q \rightarrow \mathbb{R}$ be bounded functions such that $f \leq g$. Then

$$
\begin{equation*}
\int_{Q} f \leq{\underline{\int_{Q}}} g . \tag{3.49}
\end{equation*}
$$

Proof. Let $P$ be a partition of $Q$, and let $R$ be a rectangle belonging to $P$. Clearly, $m_{R}(f) \leq m_{R}(g)$, so

$$
\begin{align*}
L(f, P) & =\sum_{R} m_{R}(f) v(R)  \tag{3.50}\\
L(g, P) & =\sum_{R} m_{R}(g) v(R)  \tag{3.51}\\
\Longrightarrow L(f, P) & \leq L(g, P) \leq{\underline{\int_{Q}}} g \tag{3.52}
\end{align*}
$$

for all partitions $P$. The lower integral

$$
\begin{equation*}
\underline{\int}_{Q} f \tag{3.53}
\end{equation*}
$$

is the l.u.b. of $L(f, P)$, so

$$
\begin{equation*}
\underline{\int}_{Q} f \leq \underline{\int}_{Q} g . \tag{3.54}
\end{equation*}
$$

Similarly,

$$
\begin{equation*}
\bar{\int}_{Q} f \leq \bar{\int}_{Q} g \tag{3.55}
\end{equation*}
$$

It follows that if $f \leq g$, then

$$
\begin{equation*}
\int_{Q} f \leq \int_{Q} g \tag{3.56}
\end{equation*}
$$

This is the monotonicity property of the R. integral.

### 3.4 Fubini Theorem

In one-dimensional calculus, when we have a continuous function $f:[a, b] \rightarrow \mathbb{R}$, then we can calculate the R. integral

$$
\begin{equation*}
\int_{a}^{b} f(x) d x=F(b)-F(a) \tag{3.57}
\end{equation*}
$$

where $F$ is the anti-derivative of $f$.
When we integrate a continuous function $f: Q \rightarrow \mathbb{R}$ over a two-dimensional region, say $Q=\left[a_{1}, b_{1}\right] \times\left[a_{2}, b_{2}\right]$, we can calculate the R. integral

$$
\begin{equation*}
\int_{Q} f=\int_{a_{1}}^{b_{1}} \int_{a_{2}}^{b_{2}} f(x, y) d x d y=\int_{a_{1}}^{b_{1}}\left(\int_{a_{2}}^{b_{2}} f(x, y) d x d y\right) \tag{3.58}
\end{equation*}
$$

That is, we can break up Q into components and integrate separately over those components. We make this more precise in the following Fubini Theorem.

First, we define some notation that will be used.
Let $n=k+\ell$ so that $\mathbb{R}^{n}=\mathbb{R}^{l} \times \mathbb{R}^{\ell}$. Let $c=\left(c_{1}, \ldots, c_{n}\right) \in \mathbb{R}^{n}$. We can write $c=(a, b)$, where $a=\left(c_{1}, \ldots, c\right) \in \mathbb{R}^{k}$ and $b=\left(c_{k+1}, \ldots, c_{k+\ell}\right) \in \mathbb{R}^{\ell}$. Similarly, let $Q=I_{1} \times \cdots I_{n}$ be a rectangle in $\mathbb{R}^{n}$. Then we can write $Q=A \times B$, where $A=I_{1} \times \cdots \times I_{k} \in \mathbb{R}^{k}$ and $B=I_{k+1} \times \cdots \times I_{k+\ell} \in \mathbb{R}^{\ell}$. Along the same lines, we can write a partition $P=\left(P_{1}, \ldots, P_{n}\right)$ as $P=\left(P_{A}, P_{B}\right)$, where $P_{A}=\left(P_{1}, \ldots, P_{k}\right)$ and $P_{B}=\left(P_{k+1}, \ldots, P_{k+\ell}\right)$.

Fubini Theorem. Let $f: Q \rightarrow \mathbb{R}$ be a bounded function and $Q=A \times B$ a rectangle as defined above. We write $f=f(x, y)$, where $x \in A$, and $y \in B$. Fixing $x \in A$, we can define a function $f_{x}: B \rightarrow \mathbb{R}$ by $f_{x}(y)=f(x, y)$. Since this function is bounded, we can define new functions $g, h: A \rightarrow \mathbb{R}$ by

$$
\begin{align*}
& g(x)={\underline{\int_{B}}}_{B} f_{x},  \tag{3.59}\\
& h(x)=\int_{B} f_{x} . \tag{3.60}
\end{align*}
$$

Note that $g \leq h$. The Fubini Theorem concludes the following: If $f$ is integrable over $Q$, then $g$ and $h$ are integrable over $A$ and

$$
\begin{equation*}
\int_{A} g=\int_{A} h=\int_{Q} f \tag{3.61}
\end{equation*}
$$

Proof. Let $P=\left(P_{A}, P_{B}\right)$ be a partition of $Q$, and let $R=R_{A} \times R_{B}$ be a rectangle belonging to $P$ (so $R_{A}$ belongs to $P_{A}$ and $R_{B}$ belongs to $P_{B}$ ). Fix $x_{0} \in A$.

First, we claim that

$$
\begin{equation*}
m_{R_{A} \times R_{B}}(f) \leq m_{R_{b}}\left(f_{x_{0}}\right), \tag{3.62}
\end{equation*}
$$

the proof of which is straightforward.
Next,

$$
\begin{align*}
\sum_{R_{B}} m_{R_{A} \times R_{B}}(f) v\left(R_{B}\right) & \leq \sum_{R_{B}} m_{R_{B}}\left(f_{x_{0}}\right) v\left(R_{B}\right) \\
& =L\left(f_{x_{0}}, P_{B}\right)  \tag{3.63}\\
& \leq \underline{\int}_{B} f_{x_{0}}=g\left(x_{0}\right)
\end{align*}
$$

So,

$$
\begin{equation*}
\sum_{R_{B}} m_{R_{A} \times R_{B}}(f) v\left(R_{B}\right) \leq g\left(x_{0}\right) \tag{3.64}
\end{equation*}
$$

for all $x_{0} \in R_{A}$. The above equation must hold for the infimum of the r.h.s, so

$$
\begin{equation*}
\sum_{R_{B}} m_{R_{A} \times R_{B}}(f) v\left(R_{B}\right) \leq m_{R_{A}}(g) \tag{3.65}
\end{equation*}
$$

Observe that $v\left(R_{A} \times R_{B}\right)=v\left(R_{A}\right) v\left(R_{B}\right)$, so

$$
\begin{align*}
L(f, P) & =\sum_{R_{A} \times R_{B}} m_{R_{A} \times R_{B}}(f) v\left(R_{A} \times R_{B}\right) \\
& \leq \sum_{R_{A}} m_{R_{A}}(g) v\left(R_{A}\right)  \tag{3.66}\\
& \leq \int_{A} g .
\end{align*}
$$

We have just shown that for any partition $P=\left(P_{A}, P_{B}\right)$,

$$
\begin{equation*}
L(f, P) \leq L\left(g, P_{A}\right) \leq \underline{\int}_{A} g \tag{3.67}
\end{equation*}
$$

so

$$
\begin{equation*}
{\underline{\int_{Q}}} f \leq{\underline{\int_{A}}}_{A} g \tag{3.68}
\end{equation*}
$$

By a similar argument, we can show that

$$
\begin{equation*}
\bar{\int}_{A} h \leq \bar{\int}_{Q} f \tag{3.69}
\end{equation*}
$$

Summarizing, we have shown that

$$
\begin{equation*}
{\underline{\int_{Q}}}_{Q} f \leq{\underline{\int_{A}}}_{A} g \leq \bar{\int}_{A} h \leq \bar{\int}_{Q} f \tag{3.70}
\end{equation*}
$$

where we used monotonicity for the middle inequality. Since $f$ is R. integrable,

$$
\begin{equation*}
{\underline{\int_{Q}}}_{Q} f=\bar{\int}_{Q} f \tag{3.71}
\end{equation*}
$$

so all of the inequalities are in fact equalities.

Remark. Suppose that for every $x \in A$, that $f_{x}: B \rightarrow \mathbb{R}$ is R . integrable. That's the same as saying $g(x)=h(x)$. Then

$$
\begin{align*}
\int_{A}\left(\int_{B} f_{x}\right) & =\int_{A} d x\left(\int_{B} f(x, y) d y\right)  \tag{3.72}\\
& =\int_{A \times B} f(x, y) d x d y
\end{align*}
$$

using standard notation from calculus.
Remark. In particular, if $f$ is continuous, then $f_{x}$ is continuous. Hence, the above remark holds for all continuous functions.

### 3.5 Properties of Riemann Integrals

We now prove some standard calculus results.
Theorem 3.13. Let $Q \subseteq \mathbb{R}^{n}$ be a rectangle, and let $f, g: Q \rightarrow \mathbb{R}$ be $R$. integrable functions. Then, for all $a, b \in \mathbb{R}$, the function $a f+b g$ is $R$. integrable and

$$
\begin{equation*}
\int_{Q} a f+b g=a \int_{Q} f+b \int_{Q} g \tag{3.73}
\end{equation*}
$$

Proof. Let's first assume that $a, b \leq 0$. Let $P$ be a partition of $Q$ and $R$ a rectangle belonging to $P$. Then

$$
\begin{equation*}
a m_{R}(f)+b m_{R}(g) \leq m_{R}(a f+b g), \tag{3.74}
\end{equation*}
$$

so

$$
\begin{align*}
a L(f, P)+b L(g, P) \leq L(a f+b g, P) & \\
& \leq \underline{\int}_{Q} a f+b g . \tag{3.75}
\end{align*}
$$

Claim. For any pair of partitions $P^{\prime}$ and $P^{\prime \prime}$,

$$
\begin{equation*}
a L\left(f, P^{\prime}\right)+b L\left(g, P^{\prime \prime}\right) \leq{\underline{\int_{Q}}} a f+b g \tag{3.76}
\end{equation*}
$$

To see that the claim is true, take $P$ to be a refinement of $P^{\prime}$ and $P^{\prime \prime}$, and apply Equation 3.75. Thus,

$$
\begin{equation*}
a{\underline{\int_{Q}}}_{Q} f+b{\underline{\int_{Q}}}_{Q} g \leq{\underline{\int_{Q}}} a f+b g . \tag{3.77}
\end{equation*}
$$

Similarly, we can show that

$$
\begin{equation*}
\bar{\int}_{Q} a f+b g \leq a \bar{\int}_{Q} f+b \bar{\int}_{Q} g \tag{3.78}
\end{equation*}
$$

Since $f$ and $g$ are R. integrable, we know that

$$
\begin{equation*}
\bar{\int}_{Q} f={\underline{\int_{Q}}}_{Q} f, \bar{\int}_{Q} g={\underline{\int_{Q}}} g \tag{3.79}
\end{equation*}
$$

These equalities show that the previous inequalities were in fact equalities, so

$$
\begin{equation*}
\int_{Q} a f+b g=a \int_{Q} f+b \int_{Q} g \tag{3.80}
\end{equation*}
$$

However, remember that we assumed that $a, b \geq 0$. To deal with the case of arbitrary $a, b$, it suffices to check what happens when we change the sign of $a$ or $b$.
Claim.

$$
\begin{equation*}
\int_{Q}-f=-\int_{Q} f \tag{3.81}
\end{equation*}
$$

Proof Hint. Let $P$ be any partition of $Q$. Then $L(f, P)=-U(-f, P)$.
You should check this claim, and then use it to complete the proof.

## Lecture 11

We review some basic properties of the Riemann integral.
Let $Q \subseteq \mathbb{R}^{n}$ be a rectangle, and let $f, g: q \rightarrow \mathbb{R}$ be bounded functions. Assume that $f, g$ are R . integrable. We have the following properties of R . integrals:

- Linearity: $a, b \in \mathbb{R} \Longrightarrow a f+b g$ is R . integrable and

$$
\begin{equation*}
\int_{Q} a f+b g=a \int_{Q} f+b \int_{Q} g . \tag{3.82}
\end{equation*}
$$

- Monotonicity: If $f \leq g$, then

$$
\begin{equation*}
\int_{Q} f \leq \int_{Q} g \tag{3.83}
\end{equation*}
$$

- Maximality Property: Let $h: Q \rightarrow \mathbb{R}$ be a function defined by $h(x)=$ $\max (f(x), g(x))$.

Theorem 3.14. The function $h$ is $R$. integrable and

$$
\begin{equation*}
\int_{Q} h \geq \max \left(\int_{Q} f, \int_{Q} g\right) . \tag{3.84}
\end{equation*}
$$

Proof. We need the following lemma.
Lemma 3.15. If $f$ and $g$ are continuous at some point $x_{0} \in Q$, then $h$ is continuous at $x_{0}$.

Proof. We consider the case $f\left(x_{0}\right)=g\left(x_{0}\right)=h\left(x_{0}\right)=r$. The functions $f$ and $g$ are continuous at $x_{0}$ if and only if for every $\epsilon>0$, there exists a $\delta>0$ such that $\left|f(x)-f\left(x_{0}\right)\right|<\epsilon$ and $\left|g(x)-g\left(x_{0}\right)\right|<\epsilon$ whenever $\left|x-x_{0}\right|<\delta$.
Substitute in $f\left(x_{0}\right)=g\left(x_{0}\right)=r$. The value of $h(x)$ is either $f(x)$ or $g(x)$, so $|h(x)-r|<\epsilon$ for $\left|x-x_{0}\right|<\delta$. That is $\left|h(x)-h\left(x_{0}\right)\right|<\epsilon$ for $\left|x-x_{0}\right|<\delta$, so $h$ is continuous at $x_{0}$.

The proofs of the other cases are left to the student.

We defined $h=\max (f, g)$. The lemma tells is that $h$ is integrable.
Define $E, F$, and $G$ as follows:

$$
\begin{align*}
& E=\text { Set of points in } Q \text { where } f \text { is discontinuous, }  \tag{3.85}\\
& F=\text { Set of points in } Q \text { where } g \text { is discontinuous, }  \tag{3.86}\\
& G=\text { Set of points in } Q \text { where } h \text { is discontinuous. } \tag{3.87}
\end{align*}
$$

The functions $f, g$ are integrable over $Q$ if and only if $E, F$ are of measure zero. The lemma shows that $G \subseteq E \cup F$, so $h$ is integrable over $Q$. To finish the proof, we notice that

$$
\begin{equation*}
h=\max (f, g) \geq f, g . \tag{3.88}
\end{equation*}
$$

Then, by monotonicity,

$$
\begin{equation*}
\int_{Q} h \geq \max \left(\int_{Q} f, \int_{Q} g\right) . \tag{3.89}
\end{equation*}
$$

Remark. Let $k=\min (f, g)$. Then $k=-\max (-f,-g)$. So, the maximality property also implies that $k$ is integrable and

$$
\begin{equation*}
\int_{Q} k \leq \min \left(\int_{Q} f, \int_{Q} g\right) . \tag{3.90}
\end{equation*}
$$

A useful trick for when dealing with functions is to change the sign. The preceding remark and the following are examples where such a trick is useful.

Let $f: Q \rightarrow \mathbb{R}$ be a R . integrable function. Define

$$
\begin{equation*}
f_{+}=\max (f, 0), \quad f_{-}=\max (-f, 0) \tag{3.91}
\end{equation*}
$$

Both of these functions are R . integrable and non-negative: $f_{+}, f_{-} \geq 0$. Also note that $f=f_{+}-f_{-}$. This decomposition is a trick we will use over and over again.

Also note that $|f|=f_{+}+f_{-}$, so $|f|$ is R . integrable. By monotonicity,

$$
\begin{align*}
\int_{Q}|f| & =\int_{Q} f_{+}+\int_{Q} f_{-} \\
& \geq \int_{Q} f_{+}-\int_{Q} f_{-}  \tag{3.92}\\
& =\int_{Q} f
\end{align*}
$$

By replacing $f$ by $-f$, we obtain

$$
\begin{align*}
\int_{Q}|f| & \geq \int_{Q}-f  \tag{3.93}\\
& =-\int_{Q} f
\end{align*}
$$

Combining these results, we arrive at the following claim

## Claim.

$$
\begin{equation*}
\int_{Q}|f| \geq\left|\int_{Q} f\right| \tag{3.94}
\end{equation*}
$$

Proof. The proof is above.

### 3.6 Integration Over More General Regions

So far we've been defining integrals over rectangles. Let us now generalize to other sets.

Let $S$ be a bounded set in $\mathbb{R}^{n}$, and let $f: S \rightarrow \mathbb{R}$ be a bounded function. Let $f_{S}: \mathbb{R}^{n} \rightarrow \mathbb{R}$ be the map defined by

$$
f_{S}(x)= \begin{cases}f(x) & \text { if } x \in S  \tag{3.95}\\ 0 & \text { if } x \notin S\end{cases}
$$

Let $Q$ be a rectangle such that $\operatorname{Int} Q \supset \bar{S}$.
Definition 3.16. The map $f$ is Riemann integrable over $S$ if $f_{S}$ is Riemann integrable over $Q$. Additionally,

$$
\begin{equation*}
\int_{S} f=\int_{Q} f_{S} \tag{3.96}
\end{equation*}
$$

One has to check that this definition does not depend on the choice of $Q$, but we do not check this here.
Claim. Let $S$ be a bounded set in $\mathbb{R}^{n}$, and let $f, g: S \rightarrow \mathbb{R}$ be bounded functions. Assume that $f, g$ are $R$. integrable over $S$. Then the following properties hold:

- Linearity: If $a, b \in \mathbb{R}$, then $a f+b g$ is $R$. integrable over $S$, and

$$
\begin{equation*}
\int_{S} a f+b g=a \int_{S} f+b \int_{S} g \tag{3.97}
\end{equation*}
$$

- Monotonicity: If $f \leq g$, then

$$
\begin{equation*}
\int_{S} f \leq \int_{S} g \tag{3.98}
\end{equation*}
$$

- Maximality: If $h=\max (f, g)$ (over the domain $S$ ), then $h$ is $R$. integrable over S, and

$$
\begin{equation*}
\int_{S} h \geq \max \left(\int_{S} f, \int_{S} g\right) . \tag{3.99}
\end{equation*}
$$

Proof. The proofs are easy, and we outline them here.

- Linearity: Note that $a f_{S}+b g_{S}=(a f+b g)_{S}$, so

$$
\begin{align*}
\int_{S} a f+b g & =\int_{Q}(a f+b g)_{S} \\
& =a \int_{Q} f_{S}+b \int_{Q} g_{S}  \tag{3.100}\\
& =a \int_{S} f+b \int_{S} g
\end{align*}
$$

- Monotonicity: Use $f \leq g \Longrightarrow f_{S} \leq g_{S}$.
- Maximality: Use $h=\max (f, g) \Longrightarrow h_{S}=\max \left(f_{S}, g_{S}\right)$.

Let's look at some nice set theoretic properties of the Riemann integral.
Claim. Let $S, T$ be bounded subsets of $\mathbb{R}^{n}$ with $T \subseteq S$. Let $f: S \rightarrow \mathbb{R}$ be bounded and non-negative. Suppose that $f$ is $R$. integrable over both $S$ and $T$. Then

$$
\begin{equation*}
\int_{T} f \leq \int_{S} f \tag{3.101}
\end{equation*}
$$

Proof. Clearly, $f_{T} \leq f_{S}$. Let $Q$ be a rectangle with $\bar{S} \supseteq \operatorname{Int} Q$. Then

$$
\begin{equation*}
\int_{Q} f_{T} \leq \int_{Q} f_{S} \tag{3.102}
\end{equation*}
$$

Claim. Let $S_{1}, S_{2}$ be bounded subsets of $\mathbb{R}^{n}$, and let $f: S_{1} \cup S_{2} \rightarrow \mathbb{R}$ be a bounded function. Suppose that $f$ is $R$. integrable over both $S_{1}$ and $S_{2}$. Then $f$ is $R$. integrable over $S_{1} \cap S_{2}$ and over $S_{1} \cup S_{2}$, and

$$
\begin{equation*}
\int_{S_{1} \cup S_{2}} f=\int_{S_{1}} f+\int_{S_{2}} f-\int_{S_{1} \cap S_{2}} f \tag{3.103}
\end{equation*}
$$

Proof. Use the following trick. Notice that

$$
\begin{align*}
& f_{S_{1} \cup S_{2}}=\max \left(f_{S_{1}}, f_{S_{2}}\right)  \tag{3.104}\\
& f_{S_{1} \cap S_{2}}=\min \left(f_{S_{1}}, f_{S_{2}}\right) \tag{3.105}
\end{align*}
$$

Now, choose $Q$ such that

$$
\begin{equation*}
\operatorname{Int} Q \supset \overline{S_{1} \cup S_{2}} \tag{3.106}
\end{equation*}
$$

so $f_{S_{1} \cup S_{2}}$ and $f_{S_{1} \cap S_{2}}$ are integrable over $Q$.
Note the identity

$$
\begin{equation*}
f_{S_{1} \cup S_{2}}=f_{S_{1}}+f_{S_{2}}-f_{S_{1} \cap S_{2}} . \tag{3.107}
\end{equation*}
$$

So,

$$
\begin{equation*}
\int_{Q} f_{S_{1} \cup S_{2}}=\int_{Q} f_{S_{1}}+\int_{Q} f_{S_{2}}-\int_{Q} f_{S_{1} \cap S_{2}} \tag{3.108}
\end{equation*}
$$

from which it follows that

$$
\begin{equation*}
\int_{S_{1} \cup S_{2}} f=\int_{S_{1}} f+\int_{S_{2}} f-\int_{S_{1} \cap S_{2}} f \tag{3.109}
\end{equation*}
$$

## Lecture 12

So far, we have been studying only the Riemann integral. However, there is also the Lebesgue integral. These are the two basic integral theories. The Riemann integral is very intuitive and is usually adequate for problems that usually come up. The Lebesgue integral is not as intuitive, but it can handle more general problems. We do not encounter these problems in geometry or physics, but we would in probability and statistics. You can learn more about Lebesgue integrals by taking Fourier Analysis (18.103) or Measure and Integration (18.125). We do not study the Lebesgue integral.

Let $S$ be a bounded subset of $\mathbb{R}^{n}$.
Theorem 3.17. If the boundary of $S$ is of measure zero, then the constant function 1 is $R$. integrable over $S$. The converse is also true.

Proof. Let $Q$ be a rectangle such that Int $Q \supset \bar{S}$. Define

$$
1_{S}(x)= \begin{cases}1 & \text { if } x \in S  \tag{3.110}\\ 0 & \text { if } x \notin S\end{cases}
$$

The constant function 1 is integrable over $S$ if and only if the function $1_{S}$ is integrable over $Q$. The function $1_{S}$ is integrable over $Q$ if the set of points $D$ in $Q$ where $1_{S}$ is discontinuous is of measure zero. If so, then

$$
\begin{equation*}
\int_{Q} 1_{S}=\int_{S} 1 \tag{3.111}
\end{equation*}
$$

Let $x \in Q$.

1. If $x \in \operatorname{Int} S$, then $1_{S}=1$ in a neighborhood of $x$, so $1_{S}$ is continuous at $x$.
2. If $x \in \operatorname{Ext} S$, then $1_{S}=0$ in a neighborhood of $x$, so $1_{S}$ is continuous at $x$.
3. If $x \in \operatorname{Bd} X$, then in every neighborhood $U$ of $x$ there exists points in Ext $S$ where $1_{S}=0$ and points in $\operatorname{Int} S$ where $1_{S}=1$. So, $1_{S}$ is discontinuous at $x$.

Thus, $D$ is the boundary of $S, D=\operatorname{Bd} S$. Therefore, the function $1_{S}$ is integrable if and only if $\operatorname{Bd} S$ is of measure zero.

### 3.7 Improper Integrals

Definition 3.18. The set $S$ is rectifiable if the boundary of $S$ is of measure zero. If $S$ is rectifiable, then

$$
\begin{equation*}
v(S)=\int_{S} 1 \tag{3.112}
\end{equation*}
$$

Let us look at the properties of $v(S)$ :

1. Monotonicity: If $S_{1}$ and $S_{2}$ are rectifiable and $S_{1} \subseteq S_{2}$, then $v\left(S_{1}\right) \leq v\left(S_{2}\right)$.
2. Linearity: If $S_{1}, S_{2}$ are rectifiable, then $S_{1} \cup S_{2}$ and $S_{1} \cap S_{2}$ are rectifiable, and

$$
\begin{equation*}
v\left(S_{1} \cup S_{2}\right)=v\left(S_{1}\right)+v\left(S_{2}\right)-v\left(S_{1} \cap S_{2}\right) \tag{3.113}
\end{equation*}
$$

3. If $S$ is rectifiable, then $v(S)=0$ if and only if $S$ is of measure zero.
4. Let $A=\operatorname{Int} S$. If $S$ is rectifiable, then $A$ is rectifiable, and $v(S)=v(A)$.

The first two properties above are special cases of the theorems that we proved last lecture:
1.

$$
\begin{equation*}
\int_{S_{1}} 1 \leq \int_{S_{2}} \text { if } S_{1} \subseteq S_{2} \tag{3.114}
\end{equation*}
$$

2. 

$$
\begin{equation*}
\int_{S_{1} \cup S_{2}} 1=\int_{S_{1}} 1+\int_{S_{2}} 1-\int_{S_{1} \cap S_{2}} 1 . \tag{3.115}
\end{equation*}
$$

To see the the third and fourth properties are true, we use some previous results. Let $Q$ be a rectangle, and let $f: Q \rightarrow \mathbb{R}$ be R . integrable. We proved the following two theorems:

Theorem A. If $f \geq 0$ and $\int_{Q} f=0$, then $f=0$ except on a set of measure zero.
Theorem B. If $f=0$ except on a set of measure zero, then $\int_{S} f=0$.
Property 3. above is a consequence of Theorem A with $f=1_{S}$.
Property 4. above is a consequence of Theorem B with $f=1_{S}-1_{A}$.
We are still lacking some simple criteria for a bounded set to be integrable. Let us now work on that.

Let $S$ be a bounded set, and let $f: S \rightarrow \mathbb{R}$ be a bounded function. We want simple criteria on $S$ and $f$ such that $f$ to be integrable over $S$.

Theorem 3.19. If $S$ is rectifiable and $f: S \rightarrow \mathbb{R}$ is bounded and continuous, then $f$ is $R$. integrable over $S$.

Proof. Let $Q$ be a rectangle such that Int $Q \supset \bar{S}$. Define $f_{S}: Q \rightarrow \mathbb{R}$ by

$$
f_{S}(x)= \begin{cases}f(x) & \text { if } x \in S  \tag{3.116}\\ 0 & \text { if } x \notin S\end{cases}
$$

By definition, $f$ is integrable over $S$ if and only if $f_{S}$ is integrable over $Q$. If so then $\int_{S} f=\int_{Q} f_{S}$.

Let $D$ be the set of points in $Q$ where $f_{S}$ is discontinuous. Then $f_{S}$ is integrable over $Q$ if and only if $D$ is of measure zero. What is $D$ ?

1. If $x \in \operatorname{Int} S$, then $f_{S}=f$ in a neighborhood of $x$, so $f_{S}$ is continuous at $x$.
2. If $x \in \operatorname{Ext} S$, then $f_{S}=0$ in a neighborhood of $x$, so $f_{S}$ is continuous at $x$.

So, we know that $D \subseteq \operatorname{Bd} S$. Because $S$ is rectifiable, the boundary of $S$ has measure zero, so $D$ has measure zero. Thus, $f_{S}$ is R. integrable, and therefore so is $f$.

Theorem 3.20. Let $A$ be an open subset of $\mathbb{R}^{n}$. There exists a sequence of compact rectifiable sets $C_{N}, N=1,2,3, \ldots$ such that

$$
\begin{equation*}
C_{N} \subseteq \operatorname{Int} C_{N+1} \tag{3.117}
\end{equation*}
$$

and

$$
\begin{equation*}
\bigcup C_{N}=A \tag{3.118}
\end{equation*}
$$

Definition 3.21. The set $\left\{C_{N}\right\}$ is called an exhaustion of $A$.
Proof. Take the complement of $A$, namely $B=\mathbb{R}^{n}-A$. Define $d(x, B)=\inf _{y \in B}\{\mid x-$ $y \mid\}$. The function $d(x, B)$ is a continuous function of $x$ (the theorem for this is in section 4 of Munkres). Let

$$
\begin{equation*}
D_{N}=\{x \in A: d(x, B) \geq 1 / N \text { and }|x| \leq N\} \tag{3.119}
\end{equation*}
$$

The set $D_{N}$ is compact. It is easy to check that $D_{N} \subseteq \operatorname{Int} D_{N+1}$.

## Claim.

$$
\begin{equation*}
\bigcup D_{N}=A \tag{3.120}
\end{equation*}
$$

Proof. Let $x \in A$. The set $A$ is open, so there exists $\epsilon>0$ such that the set $\left\{y \in \mathbb{R}^{n}:|y-x| \leq \epsilon\right\}$ is contained in $A$. So, $d(x, B) \geq \epsilon$.

Now, choose $N$ such that $1 / N<\epsilon$ and such that $|x|<N$. Then, by definition, $x \in D_{N}$. Therefore $\cup D_{N}=A$.

So, the $D_{N}$ 's satisfy the right properties, except they are not necessarily rectifiable. We can make them rectifiable as follows.

For every $p \in D_{N}$, let $Q_{p}$ be a rectangle with $p \in \operatorname{Int} Q_{p}$ and $Q_{p} \subseteq \operatorname{Int} D_{N+1}$. Then the collection of sets $\left\{\operatorname{Int} Q_{p}: p \in D_{N}\right\}$ is an open cover of $D_{N}$. By the H-B Theorem, there exists a finite subcover Int $Q_{p_{1}}, \ldots$, Int $Q_{p_{r}}$. Now, let

$$
\begin{equation*}
C_{N}=Q_{p_{1}} \cup \cdots \cup Q_{p_{r}} . \tag{3.121}
\end{equation*}
$$

Then $C_{N} \subseteq \operatorname{Int} D_{N} \subseteq \operatorname{Int} C_{N+1}$.

## Lecture 13

Let $A$ be an open set in $\mathbb{R}^{n}$, and let $f: A \rightarrow \mathbb{R}$ be a continuous function. For the moment, we assume that $f \geq 0$. Let $D \subseteq A$ be a compact and rectifiable set. Then $f \mid D$ is bounded, so $\int_{D} f$ is well-defined. Consider the set of all such integrals:

$$
\begin{equation*}
\#=\left\{\int_{D} f: D \subseteq A, D \text { compact and rectifiable }\right\} \tag{3.122}
\end{equation*}
$$

Definition 3.22. The improper integral of $f$ over $A$ exists if $*$ is bounded, and we define the improper integral of $f$ over $A$ to be its l.u.b.

$$
\begin{equation*}
\int_{A}^{\#} f \equiv \text { l.u.b. } \int_{D} f=\text { improper integral of } f \text { over } A . \tag{3.123}
\end{equation*}
$$

Claim. If $A$ is rectifiable and $f: A \rightarrow \mathbb{R}$ is bounded, then

$$
\begin{equation*}
\int_{A}^{\#} f=\int_{A} f . \tag{3.124}
\end{equation*}
$$

Proof. Let $D \subseteq A$ be a compact and rectifiable set. So,

$$
\begin{align*}
& \int_{D} f \leq \int_{A} f  \tag{3.125}\\
\Longrightarrow & \sup _{D} \int_{D} f \leq \int_{A} f  \tag{3.126}\\
\Longrightarrow & \int_{A}^{\#} f \leq \int_{A} f \tag{3.127}
\end{align*}
$$

The proof of the inequality in the other direction is a bit more complicated.
Choose a rectangle $Q$ such that $\bar{A} \subseteq \operatorname{Int} Q$. Define $f_{A}: Q \rightarrow \mathbb{R}$ by

$$
f_{A}(x)= \begin{cases}f(x) & \text { if } x \in A  \tag{3.128}\\ 0 & \text { if } x \notin A\end{cases}
$$

By definition,

$$
\begin{equation*}
\int_{A} f=\int_{Q} f_{A} \tag{3.129}
\end{equation*}
$$

Now, let $P$ be a partition of $Q$, and let $R_{1}, \ldots, R_{k}$ be rectangles belonging to a partition of $A$. If $R$ is a rectangle belonging to $P$ not contained in $A$, then $R-A \neq \phi$. In such a case, $m_{R}\left(f_{A}\right)=0$. So

$$
\begin{equation*}
L\left(f_{A}, P\right)=\sum_{i=1}^{k} m_{R_{i}}\left(f_{A}\right) v\left(R_{i}\right) \tag{3.130}
\end{equation*}
$$

On the rectangle $R_{i}$,

$$
\begin{equation*}
f_{A}=f \geq m_{R_{i}}\left(f_{A}\right) \tag{3.131}
\end{equation*}
$$

So,

$$
\begin{align*}
\sum_{i=1}^{k} m_{R_{i}}\left(f_{A}\right) v\left(R_{i}\right) & \leq \sum \int_{R_{i}} f \\
& =\int_{D} f  \tag{3.132}\\
& \leq \int_{A}^{\#}
\end{align*}
$$

where $D=\bigcup R_{i}$, which is compact and rectifiable.
The above was true for all partitions, so

$$
\begin{equation*}
\int_{Q} f_{A} \leq \int_{Z}^{\#} f \tag{3.133}
\end{equation*}
$$

We proved the inequality in the other direction, so

$$
\begin{equation*}
\int_{A} f=\int_{A}^{\#} f \tag{3.134}
\end{equation*}
$$

### 3.8 Exhaustions

Definition 3.23. A sequence of compact sets $C_{i}, i=1,2,3 \ldots$ is an exhaustion of $A$ if $C_{i} \subseteq \operatorname{Int} C_{i_{1}}$ for every $i$, and $\bigcup C_{i}=A$.

It is easy to see that

$$
\begin{equation*}
\bigcup \operatorname{Int} C_{i}=A \tag{3.135}
\end{equation*}
$$

Let $C_{i}, i=1,2,3, \ldots$ be an exhaustion of $A$ by compact rectifiable sets. Let $f: A \rightarrow \mathbb{R}$ be continuous and assume that $f \geq 0$. Note that

$$
\begin{equation*}
\int_{C_{i}} f \leq \int_{C_{i=1}} f \tag{3.136}
\end{equation*}
$$

since $C_{i=1} \supset C_{i}$. So

$$
\begin{equation*}
\int_{C_{i}} f, i=1,2,3 \ldots \tag{3.137}
\end{equation*}
$$

is an increasing (actually, non-decreasing) sequence. Hence, either $\int_{C_{i}} f \rightarrow \infty$ as $i \rightarrow \infty$, or it has a finite limit (by which we mean $\lim _{i \rightarrow \infty} \int_{C_{i}} f$ exists).

Theorem 3.24. The following two properties are equivalent:

1. $\int_{A}^{\#} f$ exists,
2. $\lim _{i \rightarrow \infty} \int_{C_{i}} f$ exists.

Moreover, if either (and hence both) property holds, then

$$
\begin{equation*}
\int_{A}^{\#} f=\lim _{i \rightarrow \infty} \int_{C_{i}} f . \tag{3.138}
\end{equation*}
$$

Proof. The set $C_{i}$ is a compact and rectifiable set contained in $A$. So, if

$$
\begin{align*}
& \int_{A}^{\#} f \text { exists, then }  \tag{3.139}\\
& \int_{C_{i}} f \leq \int_{A}^{\#} f \tag{3.140}
\end{align*}
$$

That shows that the sets

$$
\begin{equation*}
\int_{C_{i}} f, i=1,2,3 \ldots \tag{3.141}
\end{equation*}
$$

are bounded, and

$$
\begin{equation*}
\lim _{i \rightarrow \infty} \int_{C_{i}} f \leq \int_{A}^{\#} f \tag{3.142}
\end{equation*}
$$

Now, let us prove the inequality in the other direction.
The collection of sets $\left\{\operatorname{Int} C_{i}: i=1,2,3 \ldots\right\}$ is an open cover of $A$. Let $D \subseteq A$ be a compact rectifiable set contained in $A$. By the H-B Theorem,

$$
\begin{equation*}
D \subseteq \bigcup_{i=1}^{N} \operatorname{Int} C_{i} \tag{3.143}
\end{equation*}
$$

for some $N$. So, $D \subseteq \operatorname{Int} C_{N} \subseteq C_{N}$. For all such $D$,

$$
\begin{equation*}
\int_{D} f \leq \int_{C_{i}} f \leq \lim _{i \rightarrow \infty} \int_{C_{i}} f \tag{3.144}
\end{equation*}
$$

Taking the infimum over all $D$, we get

$$
\begin{equation*}
\int_{A}^{\#} f \leq \lim _{i \rightarrow \infty} \int_{C_{i}} f \tag{3.145}
\end{equation*}
$$

We have proved the inequality in both directions, so

$$
\begin{equation*}
\int_{A}^{\#} f=\lim _{i \rightarrow \infty} \int_{C_{i}} f \tag{3.146}
\end{equation*}
$$

A typical illustration of this theorem is the following example.
Consider the integral

$$
\begin{equation*}
\int_{0}^{1} \frac{d x}{\sqrt{x}} \tag{3.147}
\end{equation*}
$$

which we wrote in the normal integral notation from elementary calculus. In our notation, we would write this as

$$
\begin{equation*}
\int_{(0,1)} \frac{1}{\sqrt{x}} \tag{3.148}
\end{equation*}
$$

Let $C_{N}=\left[\frac{1}{N}, 1-\frac{1}{N}\right]$. Then

$$
\begin{align*}
\int_{(0,1)}^{\#} \frac{1}{\sqrt{x}} & =\lim _{N \rightarrow \infty} \int_{C_{N}} \frac{A}{\sqrt{x}}  \tag{3.149}\\
& =\left.2 \sqrt{x}\right|_{1 / N} ^{1-1 / N} \rightarrow 2 \text { as } N \rightarrow \infty
\end{align*}
$$

So,

$$
\begin{equation*}
\int_{(0,1)}^{\#} \frac{1}{\sqrt{x}}=2 \tag{3.150}
\end{equation*}
$$

Let us now remove the assumption that $f \geq 0$. Let $f: A \rightarrow \mathbb{R}$ be any continuous function on $A$. As before, we define

$$
\begin{align*}
f_{+}(x) & =\max \{f(x), 0\},  \tag{3.151}\\
f_{-}(x) & =\max \{-f(x), 0\} \tag{3.152}
\end{align*}
$$

We can see that $f_{+}$and $f_{-}$are continuous.
Definition 3.25. The improper R. integral of $f$ over $A$ exists if and only if the improper R. integral of $f_{+}$and $f_{-}$over $A$ exist. Moreover, we define

$$
\begin{equation*}
\int_{A}^{\#} f=\int_{A}^{\#} f_{+}-\int_{A}^{\#} f_{-} . \tag{3.153}
\end{equation*}
$$

We compute the integral using an exhaustion of $A$.

$$
\begin{align*}
\int_{A}^{\#} f & =\lim _{N \rightarrow \infty}\left(\int_{C_{N}} f_{+}-\int_{C_{N}} f_{-}\right)  \tag{3.154}\\
& =\lim _{N \rightarrow \infty} \int_{C_{N}} f .
\end{align*}
$$

Note that $|f|=f_{+}+f_{-}$, so

$$
\begin{equation*}
\lim _{N \rightarrow \infty}\left(\int_{C_{N}} f_{+}+\int_{C_{N}} f_{-}\right)=\lim _{N \rightarrow \infty} \int_{C_{N}}|f| . \tag{3.155}
\end{equation*}
$$

Therefore, the improper integral of $f$ exists if and only if the improper integral of $|f|$ exists.

Define a function $f: \mathbb{R} \rightarrow \mathbb{R}$ by

$$
f(x)= \begin{cases}0 & \text { if } x \leq 0  \tag{3.156}\\ e^{-1 / x} & \text { if } x>0\end{cases}
$$

This is a $\mathcal{C}^{\infty}(\mathbb{R})$ function. Clearly, $f^{\prime}(x)=f^{\prime \prime}(x)=\ldots=0$ when $x=0$, so in the Taylor series expansion of $f$ at zero,

$$
\begin{equation*}
\sum a_{n} x^{n}=0 \tag{3.157}
\end{equation*}
$$

all of the coefficients $a_{n}$ are zero. However, $f$ has a non-zero value in every neighborhood of zero.

Take $a \in \mathbb{R}$ and $\epsilon>0$. Define a new function $g_{a, a+\epsilon}: \mathbb{R} \rightarrow \mathbb{R}$ by

$$
\begin{equation*}
g_{a, a+\epsilon}(x)=\frac{f(x-a)}{f(x-a)+f(a+\epsilon-x)} . \tag{3.158}
\end{equation*}
$$

The function $g_{a, a+\epsilon}$ is a $\mathcal{C}^{\infty}(\mathbb{R})$ function. Notice that

$$
g_{a, a+\epsilon}= \begin{cases}0 & \text { if } x \leq a  \tag{3.159}\\ 1 & \text { if } x \geq a+\epsilon\end{cases}
$$

Take $b$ such that $a<a+\epsilon<b-\epsilon<b$. Define a new function $h_{a, b} \in \mathcal{C}^{\infty}(\mathbb{R})$ by

$$
\begin{equation*}
h_{a, b}(x)=g_{a, a+\epsilon}(x)\left(1-g_{a-\epsilon, b}(x)\right) . \tag{3.160}
\end{equation*}
$$

Notice that

$$
h_{a, b}= \begin{cases}0 & \text { if } x \leq a  \tag{3.161}\\ 1 & \text { if } a+\epsilon \leq x \leq b-\epsilon, \\ 0 & \text { if } b \leq x\end{cases}
$$

## Lecture 14

As before, let $f: \mathbb{R} \rightarrow \mathbb{R}$ be the map defined by

$$
f(x)= \begin{cases}0 & \text { if } x \leq 0  \tag{3.162}\\ e^{-1 / x} & \text { if } x>0\end{cases}
$$

This is a $\operatorname{Cinf}(\mathbb{R})$ function. Take the interval $[a, b] \in \mathbb{R}$ and define the function $f_{a, b}: \mathbb{R} \rightarrow \mathbb{R}$ by $f_{a, b}(x)=f(x-a) f(b-x)$. Note that $f_{a, b}>0$ on $(a, b)$, and $f_{a, b}=0$ on $\mathbb{R}-(a, b)$.

We generalize the definition of $f$ to higher dimensions. Let $Q \subseteq \mathbb{R}^{n}$ be a rectangle, where $Q=\left[a_{1}, b_{1}\right] \times \cdots \times\left[a_{n}, b_{n}\right]$. Define a new map $f_{Q}: \mathbb{R}^{n} \rightarrow \mathbb{R}$ bye

$$
\begin{equation*}
f_{Q}\left(x_{1}, \ldots, x_{n}\right)=f_{a_{1}, b_{1}}\left(x_{1}\right) \ldots f_{a_{n}, b_{n}}\left(x_{n}\right) . \tag{3.163}
\end{equation*}
$$

Note that $f_{Q}>0$ on Int $Q$, and that $f_{Q}=0$ on $\mathbb{R}^{n}-\operatorname{Int} Q$.

### 3.9 Support and Compact Support

Now for some terminology. Let $U$ be an open set in $\mathbb{R}^{n}$, and let $f: U \rightarrow \mathbb{R}$ be a continuous function.

Definition 3.26. The support of $f$ is

$$
\begin{equation*}
\operatorname{supp} f=\overline{\{x \in U: f(x) \neq 0\}} . \tag{3.164}
\end{equation*}
$$

For example, supp $f_{Q}=Q$.
Definition 3.27. Let $f: U \rightarrow \mathbb{R}$ be a continuous function. The function $f$ is compactly supported if supp $f$ is compact.

## Notation.

$$
\begin{equation*}
\mathcal{C}_{0}^{k}(U)=\text { The set of compactly supported } \mathcal{C}^{k} \text { functions on } U . \tag{3.165}
\end{equation*}
$$

Suppose that $f \in \mathcal{C}_{0}^{k}(U)$. Define a new set $U_{1}=\left(\mathbb{R}^{n}-\operatorname{supp} f\right)$. Then $U \cup U_{1}=\mathbb{R}^{n}$, because supp $f \subseteq U$.

Define a new map $\tilde{f}: \mathbb{R}^{n} \rightarrow \mathbb{R}$ by

$$
\tilde{f}= \begin{cases}f & \text { on } U  \tag{3.166}\\ 0 & \text { on } U_{1}\end{cases}
$$

The function $\tilde{f}$ is $\mathcal{C}^{k}$ on $U$ and $\mathcal{C}^{k}$ on $U_{1}$, so $\tilde{f}$ is in $\mathcal{C}_{0}^{k}\left(\mathbb{R}^{n}\right)$.
So, whenever we have a function $f \in \mathcal{C}^{k}$ is compactly supported on $U$, we can drop the tilde and think of $f$ as in $\mathcal{C}_{0}^{k}\left(\mathbb{R}^{n}\right)$.

### 3.10 Partitions of Unity

Let $\left\{U_{\alpha}: \alpha \in I\right\}$ be a collection of of open subsets of $\mathbb{R}^{n}$ such that $U=\cup_{\alpha} U_{\alpha}$.
Theorem 3.28. There exists a sequence of rectangles $Q_{i}, i=1,2,3, \ldots$ such that

1. Int $Q_{i}, i=1,2,3 \ldots$ is a cover of $U$,
2. Each $Q_{i} \subset I_{\alpha}$ for some $\alpha$,
3. For every point $p \in U$, there exists a neighborhood $U_{p}$ of $p$ such that $U_{p} \cap Q_{i}=\phi$ for all $i>N_{p}$.

Proof. Take an exhaustion $A_{1}, A_{2}, A_{3}, \ldots$ of $U$. By definition, the exhaustion satisfies

$$
\left\{\begin{array}{l}
A_{i} \subseteq \operatorname{Int} A_{i+1} \\
A_{i} \text { is compact } \\
\cup A_{i}=U .
\end{array}\right.
$$

We previously showed that you can always find an exhaustion.
Let $B_{i}=A_{i}-\operatorname{Int} A_{i-1}$. For each $x \in B_{i}$, let $Q_{x}$ be a rectangle with $x \in \operatorname{Int} Q_{x}$ such that $Q_{x} \subseteq U_{\alpha}$, for some alpha, and $Q_{x} \subset \operatorname{Int} A_{i+1}-A_{i-2}$. Then, the collection of sets $\left\{\right.$ Int $\left.Q_{x}: x \in B_{i}\right\}$ covers $B_{i}$. Each set $B_{i}$ is compact, so, by the H-B Theorem, there exists a finite subcover Int $Q_{x_{r}} \equiv \operatorname{Int} Q_{i, r}, r=1, \ldots, N_{i}$.

The rectangles $Q_{i, r}, 1 \leq r \leq N_{i}, i=1,2,3 \ldots$ satisfy the hypotheses of the theorem, after relabeling the rectangles in linear sequence $Q_{1}, Q_{2}, Q_{3}$, etc. (you should check this).

The following theorem is called the Partition of Unity Theorem.
Theorem 3.29. There exist functions $f_{i} \subseteq \mathcal{C}_{0}^{\infty}(U)$ such that

1. $f_{1} \geq 0$,
2. $\operatorname{supp} f_{i} \subseteq U_{\alpha}$, for some $\alpha$,
3. For every $p \in U$, there exists a neighborhood $U_{p}$ of $p$ such that $U_{p} \cup \operatorname{supp} f_{i}=\phi$ for all $i>N_{p}$,
4. $\sum f_{i}=1$.

Proof. Let $Q_{i}, i=1,2,3, \ldots$ be a collection of rectangles with the properties of the previous theorem. Then the functions $f_{Q_{i}}, i=1,2,3, \ldots$ have all the properties presented in the theorem, except for property 4 . We now prove the fourth property. We now that $f_{Q_{i}}>0$ on Int $Q_{i}$, and $\left\{\right.$ Int $\left.Q_{i}: i=1,2,3, \ldots\right\}$ is a cover of $U$. So, for every $p \in U, f_{Q_{i}}(p)>0$ for some $i$. So

$$
\begin{equation*}
\sum f_{Q_{i}}>0 \tag{3.167}
\end{equation*}
$$

We can divide by a nonzero number, so we can define

$$
\begin{equation*}
f_{i}=\frac{f_{Q_{i}}}{\sum_{i=1}^{\infty} f_{Q_{i}}} . \tag{3.168}
\end{equation*}
$$

This new function satisfies property 4 . Note that the infinite sum converges because the sum has only a finite number of nonzero terms.

## Lecture 15

We restate the partition of unity theorem from last time. Let $\left\{U_{\alpha}: \alpha \in I\right\}$ be a collection of open subsets of $\mathbb{R}^{n}$ such that

$$
\begin{equation*}
U=\bigcup_{\alpha \in I} U_{\alpha} \tag{3.169}
\end{equation*}
$$

Theorem 3.30. There exist functions $f_{i} \subseteq \mathcal{C}_{0}^{\infty}(U)$ such that

1. $f_{1} \geq 0$,
2. $\operatorname{supp} f_{i} \subseteq U_{\alpha}$, for some $\alpha$,
3. For every $p \in U$, there exists a neighborhood $U_{p}$ of $p$ such that $U_{p} \cap \operatorname{supp} f_{i}=\phi$ for all $i>N_{p}$,
4. $\sum f_{i}=1$.

Remark. Property (4) makes sense because of property (3), because at each point it is a finite sum.
Remark. A set of functions satisfying properties (1), (3), and (4) is called a partition of unity.
Remark. Property (2) can be restated as "the partition of unity is subordinate to the cover $\left\{U_{\alpha}: \alpha \in I\right\}$."

Let us look at some typical applications of partitions of unity.
The first application is to improper integrals. Let $\phi: U \rightarrow \mathbb{R}$ be a continuous map, and suppose

$$
\begin{equation*}
\int_{U} \phi \tag{3.170}
\end{equation*}
$$

is well-defined. Take a partition of unity $\sum f_{i}=1$. The function $f_{i} \phi$ is continuous and compactly supported, so it bounded. Let supp $f_{i} \subseteq Q_{i}$ for some rectangle $Q_{i}$. Then,

$$
\begin{equation*}
\int_{Q_{i}} f_{i} \phi \tag{3.171}
\end{equation*}
$$

is a well-defined R. integral. It follows that

$$
\begin{equation*}
\int_{U} f_{i} \phi=\int_{Q_{i}} f_{i} \phi \tag{3.172}
\end{equation*}
$$

It follows that

$$
\begin{equation*}
\int_{U} \phi=\sum_{i=1}^{\infty} \int_{Q_{i}} f_{i} \phi \tag{3.173}
\end{equation*}
$$

This is proved in Munkres.
The second application of partitions of unity involves cut-off functions.
Let $f_{i} \in \mathcal{C}_{0}^{\infty}(U), i=1,2,3, \ldots$ be a partition of unity, and let $A \subseteq U$ be compact.

Lemma 3.31. There exists a neighborhood $U^{\prime}$ of $A$ in $U$ and a number $N>0$ such that $A \cup \operatorname{supp} f_{i}=\phi$ for all $i>N$.

Proof. For any $p \in A$, there exists a neighborhood $U_{p}$ of $p$ and a number $N_{p}$ such that $U^{\prime} \cup \operatorname{supp} f_{i}=\phi$ for all $i>N_{p}$. The collection of all these $U_{p}$ is a cover of $A$. By the H-B Theorem, there exists a finite subcover $U_{p_{i}}, i=1,2,3, \ldots$ of $A$. Take $U_{p}=\cup U_{p_{i}}$ and take $N=\max \left\{N_{p_{i}}\right\}$.

We use this lemma to prove the following theorem.
Theorem 3.32. Let $A \subseteq \mathbb{R}^{n}$ be compact, and let $U$ be an open set containing $A$. There exists a function $f \in \mathcal{C}_{0}^{\infty}(U)$ such that $f \equiv 1$ (identically equal to 1) on a neighborhood $U^{\prime} \subset U$ of $A$.

Proof. Choose $U^{\prime}$ and $N$ as in the lemma, and let

$$
\begin{equation*}
f=\sum_{i=1}^{N} f_{i} . \tag{3.174}
\end{equation*}
$$

Then supp $f_{i} \cap U^{\prime}=\phi$ for all $i>N$. So, on $U^{\prime}$,

$$
\begin{equation*}
f=\sum_{i=1}^{\infty} f_{i}=1 \tag{3.175}
\end{equation*}
$$

Such an $f$ can be used to create cut-off functions. We look at an application.
Let $\phi: U \rightarrow \mathbb{R}$ be a continuous function. Define $\psi=f \phi$. The new function $\psi$ is called a cut-off function, and it is compactly supported with $\operatorname{supp} \phi \subseteq U$. We can extend the domain of $\psi$ by defining $\psi=0$ outside of $U$. The extended function $\psi: \mathbb{R}^{n} \rightarrow \mathbb{R}$ is still continuous, and it equals $\phi$ on a neighborhood of $A$.

We look at another application, this time to exhaustion functions.
Definition 3.33. Given an open set $U$, and a collection of compact subsets $A_{i} i=$ $1,2,3, \ldots$ of $U$, the sets $A_{i}$ form an exhaustion of $U$ if $A_{i} \subseteq$ Int $A_{i+1}$ and $\cup A_{i}=U$ (this is just a quick reminder of the definition of exhaustion).

Definition 3.34. A function $\phi \in \mathcal{C}^{\infty}(U)$ is an exhaustion function if

1. $\phi>0$,
2. the sets $A_{i}=\phi^{-1}([0,1])$ are compact.

Note that this implies that the $A_{i}^{\prime} s$ are an exhaustion.

We use the fact that we can always find a partition of unity to show that we can always find exhaustion functions.

Take a partition of unity $f_{i} \in \mathcal{C}^{\infty}(U)$, and define

$$
\begin{equation*}
\phi=\sum_{i=1}^{\infty} i f_{i} \tag{3.176}
\end{equation*}
$$

This sum converges because only finitely many terms are nonzero.
Consider any point

$$
\begin{equation*}
p \notin \bigcup_{j \leq i} \operatorname{supp} f_{j} . \tag{3.177}
\end{equation*}
$$

Then,

$$
\begin{align*}
1 & =\sum_{k=1}^{\infty} f_{k}(p)  \tag{3.178}\\
& =\sum_{k>i} f_{k}(p),
\end{align*}
$$

so

$$
\begin{align*}
\sum_{\ell=1}^{\infty} \ell f_{\ell}(p) & =\sum_{\ell>i} \ell f_{\ell} \\
& \geq i \sum_{\ell>i} f_{\ell}  \tag{3.179}\\
& =i
\end{align*}
$$

That is, if $p \notin \cup_{j \leq i}$ supp $f_{j}$, then $f(p)>i$. So,

$$
\begin{equation*}
\phi^{-1}([0, i]) \subseteq \bigcup_{j \leq i} \operatorname{supp} f_{j} \tag{3.180}
\end{equation*}
$$

which you should check yourself. The compactness of the r.h.s. implies the compactness of the l.h.s.

Now we look at problem number 4 in section 16 of Munkres. Let $A$ be an arbitrary subset of $\mathbb{R}^{n}$, and let $g: A \rightarrow \mathbb{R}^{k}$ be a map.

Definition 3.35. The function $g$ is $\mathcal{C}^{k}$ on $A$ if for every $p \in A$, there exists a neighborhood $U_{p}$ of $p$ in $\mathbb{R}^{n}$ and a $\mathcal{C}^{k} \operatorname{map} g^{p}: U_{p} \rightarrow \mathbb{R}^{k}$ such that $g^{p}\left|U_{p} \cap A=g\right| U_{p} \cap A$.

Theorem 3.36. If $g: A \rightarrow \mathbb{R}^{k}$ is $\mathcal{C}^{k}$, then there exists a neighborhood $U$ of $A$ in $\mathbb{R}^{n}$ and a $\mathcal{C}^{k}$ map $\tilde{g}: U \rightarrow \mathbb{R}^{k}$ such that $\tilde{g}=g$ on $A$.

Proof. This is a very nice application of partition of unity. Read Munkres for the proof.

## Lecture 16

## 4 Multi-linear Algebra

### 4.1 Review of Linear Algebra and Topology

In today's lecture we review chapters 1 and 2 of Munkres. Our ultimate goal (not today) is to develop vector calculus in $n$ dimensions (for example, the generalizations of grad, div, and curl).

Let $V$ be a vector space, and let $v_{i} \in V, i=1, \ldots, k$.

1. The $v_{i}^{\prime} s$ are linearly independent if the map from $\mathbb{R}^{k}$ to $V$ mapping $\left(c_{1}, \ldots, c_{k}\right)$ to $c_{1} v_{1}+\ldots+c_{k} v_{k}$ is injective.
2. The $v_{i}^{\prime} s$ span $V$ if this map is surjective (onto).
3. If the $v_{i}^{\prime} s$ form a basis, then $\operatorname{dim} V=k$.
4. A subset $W$ of $V$ is a subspace if it is also a vector space.
5. Let $V$ and $W$ be vector spaces. A map $A: V \rightarrow W$ is linear if $A\left(c_{1} v_{1}+c_{2} v_{2}\right)=$ $c_{1} A\left(v_{1}\right)+c_{2} A\left(v_{2}\right)$.
6. The kernel of a linear map $A: V \rightarrow W$ is

$$
\begin{equation*}
\operatorname{ker} A=\{v \in V: A v=0\} . \tag{4.1}
\end{equation*}
$$

7. The image of $A$ is

$$
\begin{equation*}
\operatorname{Im} A=\{A v: v \in V\} \tag{4.2}
\end{equation*}
$$

8. The following is a basic identity:

$$
\begin{equation*}
\operatorname{dim} \operatorname{ker} A+\operatorname{dim} \operatorname{Im} A=\operatorname{dim} V \tag{4.3}
\end{equation*}
$$

9. We can associate linear mappings with matrices. Let $v_{1}, \ldots, v_{n}$ be a basis for $V$, and let $w_{1}, \ldots, w_{m}$ be a basis for $W$. Let

$$
\begin{equation*}
A v_{j}=\sum_{i=1}^{m} a_{i j} w_{j} \tag{4.4}
\end{equation*}
$$

Then we associate the linear map $A$ with the matrix $\left[a_{i j}\right]$. We write this $A \sim$ $\left[a_{i j}\right]$.
10. If $v_{1}, \ldots, v_{n}$ is a basis for $V$ and $u_{j}=\sum a_{i j} w_{j}$ are $n$ arbitrary vectors in $W$, then there exists a unique linear mapping $A: V \rightarrow W$ such that $A v_{j}=u_{j}$.
11. Know all the material in Munkres section $\oint 2$ on matrices and determinants.
12. The quotient space construction. Let $V$ be a vector space and $W$ a subspace. Take any $v \in V$. We define $v+W \equiv\{v+w: w \in W\}$. Sets of this form are called $W$-cosets. One can check that given $v_{1}+W$ and $v_{2}+W$,
(a) If $v_{1}-v_{2} \in W$, then $v_{1}+W=v_{2}+W$.
(b) If $v_{1}-v_{2} \notin W$, then $\left(v_{1}+W\right) \cap\left(v_{2}+W\right)=\phi$.

So every vector $v \in V$ belongs to a unique $W$-coset.
The quotient space $V / W$ is the set of all $W=$ cosets.
For example, let $V=\mathbb{R}^{2}$, and let $W=\{(a, 0): a \in \mathbb{R}\}$. The $W$-cosets are then vertical lines.
The set $V / W$ is a vector space. It satisfies vector addition: $\left(v_{1}+W\right)+\left(v_{2}+W\right)=$ $\left(v_{1}+v_{2}\right)+W$. It also satisfies scaler multiplication: $\lambda(v+W)=\lambda v+W$. You should check that the standard axioms for vector spaces are satisfied.

There is a natural projection from $V$ to $V / W$ :

$$
\begin{equation*}
\pi: V \rightarrow V / W, v \rightarrow v+W \tag{4.5}
\end{equation*}
$$

The map $\pi$ is a linear map, it is surjective, and $\operatorname{ker} \pi=W$. Also, $\operatorname{Im} \pi=V / W$, so

$$
\begin{align*}
\operatorname{dim} V / W & =\operatorname{dim} \operatorname{Im} \pi \\
& =\operatorname{dim} V-\operatorname{dim} \operatorname{ker} \pi  \tag{4.6}\\
& =\operatorname{dim} V-\operatorname{dim} W
\end{align*}
$$

### 4.2 Dual Space

13. The dual space construction: Let $V$ be an $n$-dimensional vector space. Define $V^{*}$ to be the set of all linear functions $\ell: V \rightarrow \mathbb{R}$. Note that if $\ell_{1}, \ell_{2} \in V^{*}$ and $\lambda_{1}, \lambda_{2} \in \mathbb{R}$, then $\lambda_{1} \ell_{1}+\lambda_{2} \ell_{2} \in V^{*}$, so $V^{*}$ is a vector space.
What does $V^{*}$ look like? Let $e_{1}, \ldots, e_{n}$ be a basis of $V$. By item (9), there exists a unique linear map $e_{i}^{*} \in V^{*}$ such that

$$
\left\{\begin{array}{l}
e_{i}^{*}\left(e_{i}\right)=1, \\
e_{i}^{*}\left(e_{j}\right)=0, \text { if } j \neq i .
\end{array}\right.
$$

Claim. The set of vectors $e_{1}^{*}, \ldots, e_{n}^{*}$ is a basis of $V^{*}$.
Proof. Suppose $\ell=\sum c_{i} e_{i}^{*}=0$. Then $0=\ell\left(e_{j}\right)=\sum c_{i} e_{i}^{*}\left(e_{j}\right)=c_{j}$, so $c_{1}=$ $\ldots=c_{n}=0$. This proves that the vectors $e_{i}^{*}$ are linearly independent. Now, if $\ell \in V^{*}$ and $\ell\left(e_{i}\right)=c_{j}$ one can check that $\ell=\sum c_{i} e_{i}^{*}$. This proves that the vectors $e_{i}^{*}$ span $V^{*}$.

The vectors $e_{1}^{*}, \ldots, e_{n}^{*}$ are said to be a basis of $V^{*}$ dual to $e_{1}, \ldots, e_{n}$.
Note that $\operatorname{dim} V^{*}=\operatorname{dim} V$.
Suppose that we have a pair of vectors spaces $V, W$ and a linear map $A: V \rightarrow$ $W$. We get another map

$$
\begin{equation*}
A^{*}: W^{*} \rightarrow V^{*} \tag{4.7}
\end{equation*}
$$

defined by $A^{*} \ell=\ell \circ A$, where $\ell \in W^{*}$ is a linear map $\ell: W \rightarrow \mathbb{R}$. So $A^{*} \ell$ is a linear map $A^{*} \ell: V \rightarrow \mathbb{R}$. You can check that $A^{*}: W^{*} \rightarrow V^{*}$ is linear.

We look at the matrix description of $A^{*}$. Define the following bases:

$$
\begin{align*}
& e_{1}, \ldots, e_{n} \text { a basis of } V  \tag{4.8}\\
& f_{1}, \ldots, f_{n} \text { a basis of } W  \tag{4.9}\\
& e_{1}^{*}, \ldots, e_{n}^{*} \text { a basis of } V^{*}  \tag{4.10}\\
& f_{1}^{*}, \ldots, f_{n}^{*} \text { a basis of } W^{*} . \tag{4.11}
\end{align*}
$$

Then

$$
\begin{align*}
A^{*} f_{j}^{*}\left(e_{i}\right) & =f_{j}^{*}\left(A e_{i}\right) \\
& =f_{j}^{*}\left(\sum_{k} a_{k i} f_{k}\right)  \tag{4.12}\\
& =a_{j i}
\end{align*}
$$

So,

$$
\begin{equation*}
A^{*} f_{j}=\sum_{k} a_{j k} e_{k}^{*}, \tag{4.13}
\end{equation*}
$$

which shows that $A^{*} \sim\left[a_{j i}\right]=\left[a_{i j}\right]^{t}$, the transpose of $A$.

## Lecture 17

Today we begin studying the material that is also found in the Multi-linear Algebra Notes. We begin with the theory of tensors.

### 4.3 Tensors

Let $V$ be a $n$-dimensional vector space. We use the following notation.
Notation.

$$
\begin{equation*}
V^{k}=\underbrace{V \times \cdots \times V}_{k \text { times }} . \tag{4.14}
\end{equation*}
$$

For example,

$$
\begin{align*}
V^{2} & =V \times V  \tag{4.15}\\
V^{3} & =V \times V \times V . \tag{4.16}
\end{align*}
$$

Let $T: V^{k} \rightarrow \mathbb{R}$ be a map.
Definition 4.1. The map $T$ is linear in its ith factor if for every sequence $v_{j} \in V, 1 \leq$ $j \leq n, j \neq i$, the function mapping $v \in V$ to $T\left(v_{1}, \ldots, v_{i-1}, v, v_{i+1}, \ldots, v_{k}\right)$ is linear in $v$.

Definition 4.2. The map $T$ is $k$-linear (or is a $k$-tensor) if it is linear in all $k$ factors.
Let $T_{1}, T_{2}$ be $k$-tensors, and let $\lambda_{1}, \lambda_{2} \in \mathbb{R}$. Then $\lambda_{1} T_{1}+\lambda_{2} T_{2}$ is a $k$-tensor (it is linear in all of its factors).

So, the set of all $k$-tensors is a vector space, denoted by $\mathcal{L}^{k}(V)$, which we sometimes simply denote by $\mathcal{L}^{k}$.

Consider the special case $k=1$. The the set $\mathcal{L}^{1}(V)$ is the set of all linear maps $\ell: V \rightarrow \mathbb{R}$. In other words,

$$
\begin{equation*}
\mathcal{L}^{1}(V)=V^{*} \tag{4.17}
\end{equation*}
$$

We use the convention that

$$
\begin{equation*}
\mathcal{L}^{0}(V)=\mathbb{R} \tag{4.18}
\end{equation*}
$$

Definition 4.3. Let $T_{i} \in \mathcal{L}^{k_{i}}, i=1,2$, and define $k=k_{1}+k_{2}$. We define the tensor product of $T_{1}$ and $T_{2}$ to be the tensor $T_{1} \otimes T_{2}: V^{k} \rightarrow \mathbb{R}$ defined by

$$
\begin{equation*}
T_{1} \otimes T_{2}\left(v_{1}, \ldots, v_{k}\right)=T_{1}\left(v_{1}, \ldots, v_{k_{1}}\right) T_{2}\left(v_{k_{1}+1}, \ldots, v_{k}\right) \tag{4.19}
\end{equation*}
$$

We can conclude that $T_{1} \otimes T_{2} \in \mathcal{L}^{k}$.
We can define more complicated tensor products. For example, let $T_{i} \in \mathcal{L}^{k_{i}}, i=$ $1,2,3$, and define $k=k_{1}+k_{2}+k_{3}$. Then we have the tensor product

$$
\begin{align*}
& T_{1} \otimes T_{2} \otimes T_{3}\left(v_{1}, \ldots, v_{k}\right) \\
& \quad=T_{1}\left(v_{i}, \ldots, v_{k_{1}}\right) T_{2}\left(v_{k_{1}+1}, \ldots, v_{k_{1}+k_{2}}\right) T_{3}\left(v_{k_{1}+k_{2}+1}, \ldots, v_{k}\right) \tag{4.20}
\end{align*}
$$

Then $T_{1} \otimes T_{2} \otimes T_{3} \in \mathcal{L}^{k}$. Note that we could have simply defined

$$
\begin{align*}
T_{1} \otimes T_{2} \otimes T_{3} & =\left(T_{1} \otimes T_{2}\right) \otimes T_{3}  \tag{4.21}\\
& =T_{1} \otimes\left(T_{2} \otimes T_{3}\right),
\end{align*}
$$

where the second equality is the associative law for tensors. There are other laws, which we list here.

- Associative Law: $\left(T_{1} \otimes T_{2}\right) \otimes T_{3}=T_{1} \otimes\left(T_{2} \otimes T_{3}\right)$.
- Right and Left Distributive Laws: Suppose $T_{i} \in \mathcal{L}^{k_{i}}, i=1,2,3$, and assume that $k_{1}=k_{2}$. Then
- Left: $\left(T_{1}+T_{2}\right) \otimes T_{3}=T_{1} \otimes T_{3}+T_{2} \otimes T_{3}$.
- Right: $T_{3} \otimes\left(T_{1}+T_{2}\right)=T_{3} \otimes T_{1}+T_{3} \otimes T_{2}$.
- Let $\lambda$ be a scalar. Then

$$
\begin{equation*}
\lambda\left(T_{1} \otimes T_{2}\right)=\left(\lambda T_{1}\right) \otimes T_{2}=T_{1} \otimes\left(\lambda T_{2}\right) \tag{4.22}
\end{equation*}
$$

Now we look at an important class of $k$-tensors. Remember that $\mathcal{L}^{1}(V)=V^{*}$, and take any 1 -tensors $\ell_{i} \in V^{*}, i=1, \ldots, k$.

Definition 4.4. The tensor $T=\ell_{1} \otimes \cdots \otimes \ell_{k}$ is a decomposable $k$-tensor.
By definition, $T\left(v_{1}, \ldots, v_{k}\right)=\ell_{1}\left(v_{1}\right) \ldots \ell_{k}\left(v_{k}\right)$. That is, $\ell_{1} \otimes \cdots \otimes \ell_{k}\left(v_{1}, \ldots, v_{k}\right)=$ $\ell_{1}\left(v_{1}\right) \ldots \ell_{k}\left(v_{k}\right)$.

Now let us go back to considering $\mathcal{L}^{k}=\mathcal{L}^{k}(V)$.

## Theorem 4.5.

$$
\begin{equation*}
\operatorname{dim} \mathcal{L}^{k}=n^{k} \tag{4.23}
\end{equation*}
$$

Note that for $k=1$, this shows that $\mathcal{L}^{1}(V)=V^{*}$ has dimension $n$.
Proof. Fix a basis $e_{1}, \ldots, e_{n}$ of $V$. This defines a dual basis $e_{i}^{*}, \ldots, e_{n}^{*}$ of $V^{*}, e_{i}^{*}: V \rightarrow$ $\mathbb{R}$ defined by

$$
e_{i}^{*}\left(e_{j}\right)= \begin{cases}1 & \text { if } i=j  \tag{4.24}\\ 0 & \text { if } i \neq j\end{cases}
$$

Definition 4.6. A multi-index $I$ of length $k$ is a set of integers $\left(i_{1}, \ldots, i_{k}\right), 1 \leq i_{r} \leq n$. We define

$$
\begin{equation*}
e_{I}^{*}=e_{i_{1}}^{*} \otimes \cdots \otimes e_{i_{k}}^{*} \in \mathcal{L}^{k} \tag{4.25}
\end{equation*}
$$

Let $J=\left(j_{1}, \ldots, j_{k}\right)$ be a multi-index of length $k$. Then

$$
e_{I}^{*}\left(e_{j_{1}}, \ldots, e_{j_{k}}\right)=e_{i_{1}}^{*}\left(e_{j_{1}}\right) \ldots e_{i_{k}}^{*}\left(e_{j_{k}}\right)= \begin{cases}1 & \text { if } I=J  \tag{4.26}\\ 0 & \text { if } I \neq J\end{cases}
$$

Claim. The $k$-tensors $e_{I}^{*}$ are a basis of $\mathcal{L}^{k}$.
Proof. To prove the claim, we use the following lemma.
Lemma 4.7. Let $T$ be a $k$-tensor. Suppose that $T\left(e_{i_{1}}, \ldots, e_{i_{k}}\right)=0$ for all multiindices $I$. Then $T=0$.

Proof. Define a $(k-1)$-tensor $T_{i}: V^{k-1} \rightarrow \mathbb{R}$ by setting

$$
\begin{equation*}
T_{i}\left(v_{1}, \ldots, v_{k-1}\right)=T\left(v_{1}, \ldots, v_{k-1}, e_{j}\right), \tag{4.27}
\end{equation*}
$$

and let $v_{k}=\sum a_{i} e_{i}$. By linearity, $T\left(v_{1}, \ldots, v_{k}\right)=\sum a_{i} T_{i}\left(v_{1}, \ldots, v_{k-1}\right)$. So, if the lemma is true for the $T_{i}$ 's, then it is true for $T$ by an induction argument (we leave this to the student to prove).

With this lemma we can prove the claim.
First we show that the $e_{I}^{*}$ 's are linearly independent. Suppose that

$$
\begin{equation*}
0=T=\sum c_{I} e_{I}^{*} . \tag{4.28}
\end{equation*}
$$

For any multi-index $J$ of length $k$,

$$
\begin{align*}
0 & =T\left(e_{j_{1}}, \ldots, e_{j_{k}}\right) \\
& =\sum c_{I} e_{I}^{*}\left(e_{j_{1}}, \ldots, e_{j_{k}}\right)  \tag{4.29}\\
& =c_{J} \\
& =0
\end{align*}
$$

So the $e_{I}^{*}$ 's are linearly independent.
Now we show that the $e_{I}^{*}$ 's span $\mathcal{L}^{k}$. Let $T \in \mathcal{L}^{k}$. For every $I$ let $T_{I}=$ $T\left(e_{i_{1}}, \ldots, e_{i_{l}}\right)$, and let $T^{\prime}=\sum T_{I} e_{I}^{*}$. One can check that $\left(T-T^{\prime}\right)\left(e_{j_{1}}, \ldots, e_{j_{k}}\right)=0$ for all multi-indices $J$. Then the lemma tells us that $T=T^{\prime}$, so the $e_{I}^{*}$ 's span $\mathcal{L}^{k}$, which proves our claim.

Since the $e_{I}^{* \prime}$ 's are a basis of $\mathcal{L}^{k}$, we see that

$$
\begin{equation*}
\operatorname{dim} \mathcal{L}^{k}=n^{k} \tag{4.30}
\end{equation*}
$$

which proves our theorem.

### 4.4 Pullback Operators

Let $V, W$ be vector spacers, and let $A: V \rightarrow W$ be a linear map. Let $T \in \mathcal{L}^{k}(W)$, and define a new map $A^{*} T \in \mathcal{L}^{k}(V)$ (called the "pullback" tensor) by

$$
\begin{equation*}
A^{*} T\left(v_{1}, \ldots, v_{k}\right)=T\left(A v_{1}, \ldots, A v_{k}\right) \tag{4.31}
\end{equation*}
$$

You should prove the following claims as an exercise:
Claim. The map $A^{*}: \mathcal{L}^{k}(W) \rightarrow \mathcal{L}^{k}(V)$ is a linear map.
Claim. Let $T_{i} \in \mathcal{L}^{k_{i}}(W), i=1,2$. Then

$$
\begin{equation*}
A^{*}\left(T_{1} \otimes T_{2}\right)=A^{*} T_{1} \otimes A^{*} T_{2} \tag{4.32}
\end{equation*}
$$

Now, let $A: V \rightarrow W$ and $B: W \rightarrow U$ be maps, where $U$ is a vector space. Given $T \in \mathcal{L}^{k}(U)$, we can "pullback" to $W$ by $B^{*} T$, and then we can "pullback" to $V$ by $A^{*}\left(B^{*} T\right)=(B \circ A)^{*} T$.

### 4.5 Alternating Tensors

In this course we will be restricting ourselves to alternating tensors.
Definition 4.8. A permutation of order $k$ is a bijective map

$$
\begin{equation*}
\sigma:\{1, \ldots, k\} \rightarrow\{1, \ldots, k\} . \tag{4.33}
\end{equation*}
$$

The map is a bijection, so $\sigma^{-1}$ exists.
Given two permutations $\sigma_{1}, \sigma_{2}$, we can construct the composite permutation

$$
\begin{equation*}
\sigma_{1} \circ \sigma_{2}(i)=\sigma_{1}\left(\sigma_{2}(i)\right) \tag{4.34}
\end{equation*}
$$

We define

$$
\begin{equation*}
S_{k} \equiv \text { The set of all permutations of }\{1, \ldots, k\} \tag{4.35}
\end{equation*}
$$

There are some special permutations. Fix $1 \leq i<j \leq k$. Let $\tau$ be the permutation such that

$$
\begin{align*}
\tau(i) & =j  \tag{4.36}\\
\tau(j) & =i  \tag{4.37}\\
\tau(\ell) & =\ell, \ell \neq i, j \tag{4.38}
\end{align*}
$$

The permutation $\tau$ is called a transposition.
Definition 4.9. The permutation $\tau$ is an elementary transposition if $j=i+1$.
We state without proof two very useful theorems.

Theorem 4.10. Every permutation can be written as a product $\sigma=\tau_{1} \circ \tau_{2} \circ \cdots \circ \tau_{m}$, where each $\tau_{i}$ is an elementary transposition.

Theorem 4.11. Every permutation $\sigma$ can be written either as a product of an even number of elementary transpositions or as a product of an odd number of elementary transpositions, but not both.

Because of the second theorem, we can define an important invariant of a permutation: the sign of the permutation.

Definition 4.12. If $\sigma=\tau_{1} \circ \cdots \circ \tau_{m}$, where the $\tau_{i}$ 's are elementary transpositions, then the sign of $\sigma$ is

$$
\begin{equation*}
\text { sign of } \sigma=(-1)^{\sigma}=(-1)^{m} \tag{4.39}
\end{equation*}
$$

Note that if $\sigma=\sigma_{1} \circ \sigma_{2}$, then $(-1)^{\sigma}=(-1)^{\sigma_{1}}(-1)^{\sigma_{2}}$. We can see this by letting $\sigma_{1}=\tau_{1} \circ \cdots \circ \tau_{m_{1}}$, and $\sigma_{2}=\tau_{1}^{\prime} \circ \cdots \circ \tau_{m_{2}}^{\prime}$, and noting that $\sigma_{1} \circ \sigma_{2}=\tau_{1} \circ \cdots \circ \tau_{m_{1}} \circ$ $\tau_{1}^{\prime} \circ \cdots \circ \tau_{m_{2}}^{\prime}$.

## Lecture 18

We begin with a quick review of permutations (from last lecture).
A permutation of order $k$ is a bijective map $\sigma:\{1, \ldots, k\} \rightarrow\{1, \ldots, k\}$. We denote by $S_{k}$ the set of permutations of order $k$.

The set $S_{k}$ has some nice properties. If $\sigma \in S_{k}$, then $\sigma^{-1} \in S_{k}$. The inverse permutation $\sigma^{-1}$ is defined by $\sigma^{-1}(j)=i$ if $\sigma(i)=j$. Another nice property is that if $\sigma, \tau \in S_{k}$, then $\sigma \tau \in S_{k}$, where $\sigma \tau(i)=\sigma(\tau(i))$. That is, if $\tau(i)=j$ and $\sigma(j)=k$, then $\sigma \tau(i)=k$.

Take $1 \leq i<j \leq k$, and define

$$
\begin{align*}
\tau_{i, j}(i) & =j  \tag{4.40}\\
\tau_{i, j}(j) & =i  \tag{4.41}\\
\tau_{i, j}(\ell) & =\ell, \ell \neq i, j \tag{4.42}
\end{align*}
$$

The permutation $\tau_{i, j}$ is a transposition. It is an elementary transposition of $j=i+1$. Last time we stated the following theorem.

Theorem 4.13. Every permutation $\sigma$ can be written as a product

$$
\begin{equation*}
\sigma=\tau_{1} \tau_{2} \cdots \tau_{r} \tag{4.43}
\end{equation*}
$$

where the $\tau_{i}$ 's are elementary transpositions.
In the above, we removed the symbol $\circ$ denoting composition of permutations, but the composition is still implied.

Last time we also defined the sign of a permutation
Definition 4.14. The sign of a permutation $\sigma$ is $(-1)^{\sigma}=(-1)^{r}$, where $r$ is as in the above theorem.

Theorem 4.15. The above definition of sign is well-defined, and

$$
\begin{equation*}
(-1)^{\sigma \tau}=(-1)^{\sigma}(-1)^{\tau} \tag{4.44}
\end{equation*}
$$

All of the above is discussed in the Multi-linear Algebra Notes.
Part of today's homework is to show the following two statements:

1. $\left|S_{k}\right|=k$ !. The proof is by induction.
2. $(-1)^{\tau_{i, j}}=-1$. Hint: use induction and $\tau_{i, j}=\left(\tau_{j-1, j}\right)\left(\tau_{i, j-1}\right)\left(\tau_{j-1, j}\right)$, with $i<j$.

We now move back to the study of tensors. Let $V$ be an $n$-dimensional vector space. We define

$$
\begin{equation*}
V^{k}=\underbrace{V \times \cdots \times V}_{k \text { factors }} . \tag{4.45}
\end{equation*}
$$

We define $\mathcal{L}^{k}(v)$ to be the space of all $k$-linear functions $T: V^{k} \rightarrow \mathbb{R}$. If $T_{i} \in \mathcal{L}^{k_{i}}, i=$ 1,2 , and $k=k_{1}+k_{2}$, then $T_{1} \otimes T_{2} \in \mathcal{L}^{k}$. Decomposable $k$-tensors are of the form $T=\ell_{1} \otimes \cdots \otimes \ell_{k}$, where each $\ell_{i} \in \mathcal{L}^{1}=V^{*}$. Note that $\ell_{1} \otimes \cdots \otimes \ell_{k}\left(v_{1}, \ldots, v_{k}\right)=$ $\ell_{1}\left(v_{1}\right) \ldots \ell_{k}\left(v_{k}\right)$.

We define a permutation operation on tensors. Take $\sigma \in S_{k}$ and $T \in \mathcal{L}^{k}(V)$.
Definition 4.16. We define the map $T^{\sigma}: V^{k} \rightarrow \mathbb{R}$ by

$$
\begin{equation*}
T^{\sigma}\left(v_{1}, \ldots, v_{k}\right)=T\left(v_{\sigma^{-1}(1)}, \ldots, v_{\sigma^{-1}(k)}\right) \tag{4.46}
\end{equation*}
$$

Clearly, $T^{\sigma} \in \mathcal{L}^{k}(V)$. We have the following useful formula:

## Claim.

$$
\begin{equation*}
\left(T^{\sigma}\right)^{\tau}=T^{\tau \sigma} . \tag{4.47}
\end{equation*}
$$

Proof.

$$
\begin{align*}
T^{\tau \sigma}\left(v_{1}, \ldots, v_{k}\right) & =T\left(v_{\sigma^{-1}\left(\tau^{-1}(1)\right)}, \ldots, v_{\sigma^{-1}\left(\tau^{-1}(k)\right)}\right) \\
& =T^{\sigma}\left(v_{\tau^{-1}(1)}, \ldots, v_{\tau^{-1}(k)}\right)  \tag{4.48}\\
& =\left(T^{\sigma}\right)^{\tau}\left(v_{1}, \ldots, v_{k}\right)
\end{align*}
$$

Let us look at what the permutation operation does to a decomposable tensor $T=\ell_{1} \otimes \cdots \otimes \ell_{k}$.

$$
\begin{equation*}
T^{\sigma}\left(v_{1}, \ldots, v_{k}\right)=\ell_{1}\left(v_{\sigma^{-1}(1)}\right) \ldots \ell_{k}\left(v_{\sigma^{-1}(k)}\right) . \tag{4.49}
\end{equation*}
$$

The $i$ th factor has the subscript $\sigma^{-1}(i)=j$, where $\sigma(j)=i$, so the the $i$ th factor is $\ell_{\sigma(j)}\left(v_{j}\right)$. So

$$
\begin{align*}
T^{\sigma}\left(v_{1}, \ldots, v_{k}\right) & =\ell_{\sigma(1)}\left(v_{1}\right) \ldots \ell_{\sigma(k)}\left(v_{k}\right)  \tag{4.50}\\
& =\left(\ell_{\sigma(1)} \otimes \cdots \otimes \ell_{\sigma(k)}\right)\left(v_{1}, \ldots, v_{k}\right)
\end{align*}
$$

To summarize,

$$
\left\{\begin{array}{l}
T=\ell_{1} \otimes \cdots \otimes \ell_{k}  \tag{4.51}\\
T^{\sigma}=\ell_{\sigma(1)} \otimes \cdots \otimes \ell_{\sigma(k)} .
\end{array}\right.
$$

Proposition 4.17. The mapping $T \in \mathcal{L}^{k} \rightarrow T^{\sigma} \in \mathcal{L}^{k}$ is linear .
We leave the proof of this as an exercise.
Definition 4.18. A tensor $T \in \mathcal{L}^{k}(V)$ is alternating if $T^{\sigma}=(-1)^{\sigma} T$ for all $\sigma \in S_{k}$.
Definition 4.19. We define

$$
\begin{equation*}
\mathcal{A}^{k}(V)=\text { the set of all alternating } k \text {-tensors. } \tag{4.52}
\end{equation*}
$$

By our previous claim, $\mathcal{A}^{k}$ is a vector space.
The alternating operator Alt can be used to create alternating tensors.
Definition 4.20. Given a $k$-tensor $T \in \mathcal{L}^{k}(V)$, we define the alternating operator Alt : $\mathcal{L}^{k}(V) \rightarrow \mathcal{A}^{k}(V)$ by

$$
\begin{equation*}
\operatorname{Alt}(T)=\sum_{\tau \in S_{k}}(-10)^{\tau} T^{\tau} \tag{4.53}
\end{equation*}
$$

Claim. The alternating operator has the following properties:

1. $\operatorname{Alt}(T) \in \mathcal{A}^{k}(V)$,
2. If $T \in \mathcal{A}^{k}(V)$, then $\operatorname{Alt}(T)=k!T$,
3. $\operatorname{Alt}\left(T^{\sigma}\right)=(-1)^{\sigma} \operatorname{Alt}(T)$,
4. The map Alt: $\mathcal{L}^{k}(V) \rightarrow \mathcal{A}^{k}(V)$ is linear.

Proof. 1.

$$
\begin{equation*}
\operatorname{Alt}(T)=\sum_{\tau}(-1)^{\tau} T^{\tau} \tag{4.54}
\end{equation*}
$$

so

$$
\begin{align*}
\operatorname{Alt}(T)^{\sigma} & =\sum_{\tau}(-1)^{\tau}\left(T^{\tau}\right)^{\sigma} \\
& =\sum_{\tau}(-1)^{\tau} T^{\sigma \tau}  \tag{4.55}\\
& =(-1)^{\sigma} \sum_{\sigma \tau}(-1)^{\sigma \tau} T^{\sigma \tau} \\
& =(-1)^{\sigma} \operatorname{Alt}(T)
\end{align*}
$$

2. 

$$
\begin{equation*}
\operatorname{Alt}(T)=\sum_{\tau}(-1)^{\tau} T^{\tau} \tag{4.56}
\end{equation*}
$$

but $T^{\tau}=(-1)^{\tau} T$, since $T \in \mathcal{A}^{k}(V)$. So

$$
\begin{align*}
\operatorname{Alt}(T) & =\sum_{\tau}(-1)^{\tau}(-1)^{\tau} T  \tag{4.57}\\
& =k!T
\end{align*}
$$

3. 

$$
\begin{align*}
\operatorname{Alt}\left(T^{\sigma}\right) & =\sum_{\tau}(-1)^{\tau}\left(T^{\sigma}\right)^{\tau} \\
& =\sum_{\tau}(-1)^{\tau} T^{\tau \sigma}  \tag{4.58}\\
& =(-1)^{\sigma} \sum_{\tau \sigma}(-1)^{\tau \sigma} T^{\tau \sigma} \\
& =(-1)^{\sigma} \operatorname{Alt}(T) .
\end{align*}
$$

4. We leave the proof as an exercise.

Now we ask ourselves: what is the dimension of $\mathcal{A}^{k}(V)$ ? To answer this, it is best to write a basis.

Earlier we found a basis for $\mathcal{L}^{k}$. We defined $e_{1}, \ldots, e_{n}$ to be a basis of $V$ and $e_{1}^{*}, \ldots, e_{n}^{*}$ to be a basis of $V^{*}$. We then considered multi-indices $I=\left(i_{1}, \ldots, i_{k}\right), 1 \leq$ $i_{r} \leq n$ and defined $\left\{e_{I}^{*}=e_{i_{1}}^{*} \otimes \cdots \otimes e_{i_{k}}^{*}, I\right.$ a multi-index $\}$ to be a basis of $\mathcal{L}^{k}$. For any multi-index $J=\left(j_{1}, \ldots, j_{k}\right)$, we had

$$
e_{I}^{*}\left(e_{j_{1}}, \ldots, e_{j_{k}}\right)= \begin{cases}1 & \text { if } I=J  \tag{4.59}\\ 0 & \text { if } I \neq J\end{cases}
$$

Definition 4.21. A multi-index $I=\left(i_{1}, \ldots, i_{k}\right)$ is repeating if $i_{r}=i_{s}$ for some $r<s$.
Definition 4.22. The multi-index $I$ is strictly increasing if $1 \leq i_{1}<\ldots<i_{k} \leq n$.
Notation. Given $\sigma \in S_{k}$ and $I=\left(i_{1}, \ldots, i_{k}\right)$, we denote $I^{\sigma}=\left(i_{\sigma(1)}, \ldots, i_{\sigma(k)}\right)$.
Remark. If $J$ is a non-repeating multi-index, then there exists a permutation $\sigma$ such that $J=I^{\sigma}$, where $I$ is strictly increasing.

$$
\begin{equation*}
e_{J}^{*}=e_{I^{\sigma}}^{*}=e_{\sigma\left(i_{1}\right)}^{*} \otimes \cdots \otimes e_{\sigma\left(i_{k}\right)}^{*}=\left(e_{I}^{*}\right)^{\sigma} \tag{4.60}
\end{equation*}
$$

Define $\psi_{I}=\operatorname{Alt}\left(e_{I}^{*}\right)$.
Theorem 4.23. 1. $\psi_{I^{\sigma}}=(-1)^{\sigma} \psi_{I}$,
2. If $I$ is repeating, then $\psi_{I}=0$,
3. If $I, J$ are strictly increasing, then

$$
\psi_{I}\left(e_{j_{1}}, \ldots, e_{j_{k}}\right)= \begin{cases}1 & \text { if } I=J  \tag{4.61}\\ 0 & \text { if } I \neq J\end{cases}
$$

Proof. 1.

$$
\begin{align*}
\psi_{I^{\sigma}} & =\operatorname{Alt} e_{I^{\sigma}}^{*} \\
& =\operatorname{Alt}\left(\left(e_{I}^{*}\right)^{\sigma}\right) \\
& =(-1)^{\sigma} \operatorname{Alt} e_{I}^{*}  \tag{4.62}\\
& =(-1)^{\sigma} \psi_{I} .
\end{align*}
$$

2. Suppose that $I$ is repeating. Then $I=I^{\tau}$ for some transposition $\tau$. So $\psi_{I}=$ $(-1)^{\tau} \psi_{I}$. But (as you proved in the homework) $(-1)^{\tau}=-1$, so $\psi_{I}=0$.
3. 

$$
\begin{align*}
\psi_{I} & =\operatorname{Alt}\left(e_{I}^{*}\right) \\
& =\sum_{\tau}(-1)^{\tau} e_{I^{\tau}}^{*}, \tag{4.63}
\end{align*}
$$

so

$$
\begin{align*}
& \psi_{I}\left(e_{j_{1}}, \ldots, e_{j_{k}}\right)=\sum_{\tau}(-1)^{\tau} \underbrace{e_{\tau^{\tau}}^{*}\left(e_{j_{1}}, \ldots, e_{j_{k}}\right)}  \tag{4.64}\\
& \begin{cases}1 & \text { if } I^{\tau}=J, \\
0 & \text { if } I^{\tau} \neq J .\end{cases}
\end{align*}
$$

But $I^{\tau}=J$ only if $\tau$ is the identity permutation (because both $I^{\tau}$ and $J$ are strictly increasing). The only non-zero term in the sum is when $\tau$ is the identity permutation, so

$$
\psi_{I}\left(e_{j_{1}}, \ldots, e_{j_{k}}\right)= \begin{cases}1 & \text { if } I=J  \tag{4.65}\\ 0 & \text { if } I \neq J\end{cases}
$$

Corollary 5. The alternating $k$-tensors $\psi_{I}$, where $I$ is strictly increasing, are a basis of $\mathcal{A}^{k}(V)$.

Proof. Take $T \in \mathcal{A}^{k}(V)$. The tensor $T$ can be expanded as $T=\sum c_{I} e_{I}^{*}$. So

$$
\begin{align*}
\operatorname{Alt}(T) & =k!\sum c_{I} \operatorname{Alt}\left(e_{I}^{*}\right) \\
& =k!\sum c_{I} \psi_{I} \tag{4.66}
\end{align*}
$$

If $I$ is repeating, then $\psi_{I}=0$. If $I$ is non-repeating, then $I=J^{\sigma}$, where $J$ is strictly increasing. Then $\psi_{I}=(-1)^{\sigma} \psi_{J}$.

So, we can replace all multi-indices in the sum by strictly increasing multi-indices,

$$
\begin{equation*}
T=\sum a_{I} \psi_{I}, I \text { 's strictly increasing. } \tag{4.67}
\end{equation*}
$$

Therefore, the $\psi_{I}$ 's span $\mathcal{A}^{k}(V)$. Moreover, the $\psi_{I}$ 's are a basis if and only if the $a_{i}$ 's are unique. We show that the $a_{I}$ 's are unique.

Let $J$ be any strictly increasing multi-index. Then

$$
\begin{align*}
T\left(e_{j_{1}}, \ldots, e_{j_{k}}\right) & =\sum a_{I} \psi\left(e_{j_{1}}, \ldots, e_{j_{k}}\right)  \tag{4.68}\\
& =a_{J}
\end{align*}
$$

by property (3) of the previous theorem. Therefore, the $\psi_{I}$ 's are a basis of $\mathcal{A}^{k}(V)$.

## Lecture 19

We begin with a review of tensors and alternating tensors.
We defined $\mathcal{L}^{k}(V)$ to be the set of $k$-linear maps $T: V^{k} \rightarrow \mathbb{R}$. We defined $e_{1}, \ldots, e_{n}$ to be a basis of $V$ and $e_{1}^{*}, \ldots, e_{n}^{*}$ to be a basis of $V^{*}$. We also defined $\left\{e_{I}^{*}=e_{i_{1}}^{*} \otimes \cdots \otimes e_{i_{k}}^{*}\right\}$ to be a basis of $\mathcal{L}^{k}(V)$, where $I=\left(i_{1}, \ldots, i_{k}\right), 1 \leq i_{r} \leq n$ is a multi-index. This showed that $\operatorname{dim} \mathcal{L}^{k}=n^{k}$.

We defined the permutation operation on a tensor. For $\sigma \in S_{n}$ and $T \in \mathcal{L}^{k}$, we defined $T^{\sigma} \in \mathcal{L}^{k}$ by $T^{\sigma}\left(v_{1}, \ldots, v_{k}\right)=T\left(v_{\sigma^{-1}(1)}, \ldots, v_{\sigma^{-1}(k)}\right)$. Then we defined that $T$ is alternating if $T^{\sigma}=(-1)^{\sigma} T$. We defined $\mathcal{A}^{k}=\mathcal{A}^{k}(V)$ to be the space of all alternating $k$-tensors.

We defined the alternating operator Alt : $\mathcal{L}^{k} \rightarrow \mathcal{A}^{k}$ by $\operatorname{Alt}(T)=\sum(-1)^{\sigma} T^{\sigma}$, and we defined $\psi_{I}=\operatorname{Alt}\left(e_{I}^{*}\right)$, where $I=\left(i_{1}, \ldots, i_{k}\right)$ is a strictly increasing multi-index. We proved the following theorem:

Theorem 4.24. The $\psi_{I}$ 's (where $I$ is strictly increasing) are a basis for $\mathcal{A}^{k}(V)$.
Corollary 6. If $0 \leq k \leq n$, then

$$
\begin{equation*}
\operatorname{dim} \mathcal{A}^{k}=\binom{n}{k}=\frac{n!}{k!(n-k)!} . \tag{4.69}
\end{equation*}
$$

Corollary 7. If $k>n$, then $\mathcal{A}^{k}=\{0\}$.
We now ask what is the kernel of Alt? That is, for which $T \in \mathcal{L}^{k}$ is $\operatorname{Alt}(T)=0$ ?
Let $T \in \mathcal{L}^{k}$ be a decomposable $k$-tensor, $T=\ell_{1} \otimes \cdots \otimes \ell_{k}$, where each $\ell_{i} \in V^{*}$.
Definition 4.25. The $k$-tensor $T$ is redundant if $\ell_{i}=\ell_{i+1}$ for some $1 \leq i \leq k-1$.
We define

$$
\begin{equation*}
\mathcal{I}^{k} \equiv \operatorname{Span}\{\text { redundant } k \text {-tensors }\} \tag{4.70}
\end{equation*}
$$

Claim. If $T \in \mathcal{I}^{k}$, then $\operatorname{Alt}(T)=0$.
Proof. It suffices to prove this for $T=\ell_{1} \otimes \cdots \otimes \ell_{k}$, where $\ell_{1}=\ell_{i+1}$ ( $T$ is redundant). Let $\tau=\tau_{i, i+1} \in S_{k}$. So, $T^{\tau}=T$. But

$$
\begin{align*}
\operatorname{Alt}\left(T^{\tau}\right) & =(-1)^{\tau} \operatorname{Alt}(T) \\
& =-\operatorname{Alt}(T), \tag{4.71}
\end{align*}
$$

so $\operatorname{Alt}(T)=0$.
Claim. Suppose that $T \in \mathcal{I}^{k}$ and $T^{\prime} \in \mathcal{L}^{m}$. Then $T^{\prime} \otimes T \in \mathcal{I}^{k+n}$ and $T \otimes T^{\prime} \in \mathcal{I}^{k+m}$.

Proof. We can assume that $T$ and $T^{\prime}$ are both decomposable tensors.

$$
\begin{align*}
T & =\ell_{1} \otimes \cdots \otimes \ell_{k}, \quad \ell_{i}=\ell_{i+1},  \tag{4.72}\\
T^{\prime} & =\ell_{1}^{\prime} \otimes \cdots \otimes \ell_{m}^{\prime},  \tag{4.73}\\
T \otimes T^{\prime} & =\ell_{1} \otimes \cdots \otimes \underbrace{\ell_{i} \otimes \ell_{i+1}}_{\text {a redundancy }} \otimes \cdots \otimes \ell_{k} \otimes \ell_{1}^{\prime} \otimes \cdots \otimes \ell_{m}^{\prime}  \tag{4.74}\\
& \in \mathcal{I}^{k+m} . \tag{4.75}
\end{align*}
$$

A similar argument holds for $T^{\prime} \otimes T$.
Claim. For each $T \in \mathcal{L}^{k}$ and $\sigma \in S_{k}$, there exists some $w \in \mathcal{I}^{k}$ such that

$$
\begin{equation*}
T=(-1)^{\sigma} T^{\sigma}+W \tag{4.76}
\end{equation*}
$$

Proof. In proving this we can assume that $T$ is decomposable. That is, $T=\ell_{1} \otimes \cdots \otimes$ $\ell_{k}$.

We first check the case $k=2$. Let $T=\ell_{1} \otimes \ell_{2}$. The only (non-identity) permutation is $\sigma=\tau_{1,2}$. In this case, $T=(-1)^{\sigma} T^{\sigma}+W$ becomes $W=T+T^{\sigma}$, so

$$
\begin{align*}
W & =T+T^{\sigma} \\
& =\ell_{1} \otimes \ell_{2}+\ell_{2} \otimes \ell_{1} \\
& =\left(\ell_{1}+\ell_{2}\right) \otimes\left(\ell_{1}+\ell_{2}\right)-\ell_{1} \otimes \ell_{1}-\ell_{2} \otimes \ell_{2}  \tag{4.77}\\
& \in \mathcal{I}^{2}
\end{align*}
$$

We now check the case $k$ is arbitrary. Let $T=\ell_{1} \otimes \cdots \otimes \ell_{k}$ and $\sigma=\tau_{1} \tau_{2} \ldots \tau_{r} \in S_{k}$, where the $\tau_{i}$ 's are elementary transpositions. We will prove that $W \in \mathcal{I}^{k}$ by induction on $r$.

- Case $r=1$ : Then $\sigma=\tau_{i, i+1}$, and

$$
\begin{align*}
W & =T+T^{\sigma} \\
& =\left(\ell_{1} \otimes \cdots \otimes \ell_{k}\right)+\left(\ell_{1} \otimes \cdots \otimes \ell_{k}\right)^{\sigma} \\
& =\ell_{1} \otimes \cdots \otimes \ell_{i-1} \otimes\left(\ell_{i} \otimes \ell_{i+1}+\ell_{i+1} \otimes \ell_{i}\right) \otimes \ell_{i+2} \otimes \cdots \otimes \ell_{k}  \tag{4.78}\\
& \in \mathcal{I}^{k}
\end{align*}
$$

because $\left(\ell_{i} \otimes \ell_{i+1}+\ell_{i+1} \otimes \ell_{i}\right) \in \mathcal{I}^{k}$.

- Induction step $((r-1) \Longrightarrow r)$ : Let $\beta=\tau_{2} \ldots \tau_{r}$, and let $\tau=\tau_{1}$ so that $\sigma=\tau_{1} \tau_{2} \ldots \tau_{r}=\tau \beta$. Then

$$
\begin{equation*}
T^{\sigma}=\left(T^{\beta}\right)^{\tau} \tag{4.79}
\end{equation*}
$$

By induction, we know that

$$
\begin{equation*}
T^{\beta}=(-1)^{\beta} T+W \tag{4.80}
\end{equation*}
$$

for some $W \in \mathcal{I}^{k}$. So,

$$
\begin{align*}
T^{\sigma} & =(-1)^{\beta} T^{\tau}+W^{\tau} \\
& =(-1)^{\beta}(-1)^{\tau} T+W^{\tau}  \tag{4.81}\\
& =(-1)^{\sigma} T+W^{\tau},
\end{align*}
$$

where $W^{\tau}=(-1)^{\tau} W+W^{\prime} \in \mathcal{I}^{k}$.

Corollary 8. For every $T \in \mathcal{L}^{k}$,

$$
\begin{equation*}
\operatorname{Alt}(T)=k!T+W \tag{4.82}
\end{equation*}
$$

for some $W \in \mathcal{I}^{k}$.
Proof.

$$
\begin{equation*}
\operatorname{Alt}(T)=\sum_{\sigma}(-1)^{\sigma} T^{\sigma} \tag{4.83}
\end{equation*}
$$

but we know that $T^{\sigma}=(-1)^{\sigma} T+W_{\sigma}$, for some $W_{\sigma} \in \mathcal{I}^{k}$, so

$$
\begin{align*}
\operatorname{Alt}(T) & =\sum_{\sigma}\left(T+(-1)^{\sigma} W_{\sigma}\right)  \tag{4.84}\\
& =k!T+W,
\end{align*}
$$

where $W=\sum_{\sigma}(-1)^{\sigma} W_{\sigma} \in \mathcal{I}^{k}$.
Theorem 4.26. Every $T \in \mathcal{L}^{k}$ can be written uniquely as a sum

$$
\begin{equation*}
T=T_{1}+T_{2}, \tag{4.85}
\end{equation*}
$$

where $T_{1} \in \mathcal{A}^{k}$ and $T_{2} \in \mathcal{I}^{k}$.
Proof. We know that $\operatorname{Alt}(T)=k!T+W$, for some $W \in \mathcal{I}^{k}$. Solving for $T$, we get

$$
\begin{equation*}
T=\underbrace{\frac{1}{k!} \operatorname{Alt}(T)}_{T_{1}}-\underbrace{\frac{1}{k!} W}_{T_{2}} . \tag{4.86}
\end{equation*}
$$

We check uniqueness:

$$
\begin{equation*}
\operatorname{Alt}(T)=\underbrace{\operatorname{Alt}\left(T_{1}\right)}_{k!T_{1}}+\underbrace{\operatorname{Alt}\left(T_{2}\right)}_{0}, \tag{4.87}
\end{equation*}
$$

so $T_{1}$ is unique, which implies that $T_{2}$ is also unique.
Claim.

$$
\begin{equation*}
\mathcal{I}^{k}=\text { ker Alt } \tag{4.88}
\end{equation*}
$$

Proof. If Alt $T=0$, then

$$
\begin{equation*}
T=-\frac{1}{k!} W, \quad W \in \mathcal{I}^{k} \tag{4.89}
\end{equation*}
$$

so $T \in \mathcal{I}^{k}$.
The space $\mathcal{I}^{k}$ is a subspace of $\mathcal{L}^{k}$, so we can form the quotient space

$$
\begin{equation*}
\Lambda^{k}\left(V^{*}\right) \equiv \mathcal{L}^{k} / \mathcal{I}^{k} \tag{4.90}
\end{equation*}
$$

What's up with this notation $\Lambda^{k}\left(V^{*}\right)$ ? We motivate this notation with the case $k=1$. There are no redundant 1 -tensors, so $\mathcal{I}^{1}=\{0\}$, and we already know that $\mathcal{L}^{1}=V^{*}$. So

$$
\begin{equation*}
\Lambda^{1}\left(V^{*}\right)=V^{*} / \mathcal{I}^{1}=\mathcal{L}^{1}=V^{*} \tag{4.91}
\end{equation*}
$$

Define the map $\pi: \mathcal{L}^{k} \rightarrow \mathcal{L}^{k} / \mathcal{I}^{k}$. The map $\pi$ is onto, and $\operatorname{ker} \pi=\mathcal{I}^{k}$.
Claim. The map $\pi$ maps $\mathcal{A}^{k}$ bijectively onto $\Lambda^{k}\left(V^{*}\right)$.
Proof. Every element of $\Lambda^{k}$ is of the form $\pi(T)$ for some $T \in \mathcal{L}^{k}$. We can write $T=T_{1}+T_{2}$, where $T_{1} \in \mathcal{A}^{k}$ and $T_{2} \in \mathcal{I}^{k}$. So,

$$
\begin{align*}
\pi(T) & =\pi\left(T_{1}\right)+\pi\left(T_{2}\right) \\
& =\pi\left(T_{1}\right)+0  \tag{4.92}\\
& =\pi\left(T_{1}\right)
\end{align*}
$$

So, $\pi$ maps $\mathcal{A}^{k}$ onto $\Lambda^{k}$. Now we show that $\pi$ is one-to-one. If $T \in \mathcal{A}^{k}$ and $\pi(T)=0$, then $T \in \mathcal{I}^{k}$ as well. We know that $\mathcal{A}^{k} \cap \mathcal{I}^{k}=\{0\}$, so $\pi$ is bijective.

We have shown that

$$
\begin{equation*}
\mathcal{A}^{k}(V) \cong \Lambda^{k}\left(V^{*}\right) \tag{4.93}
\end{equation*}
$$

The space $\Lambda^{k}\left(V^{*}\right)$ is not mentioned in Munkres, but sometimes it is useful to look at the same space in two different ways.

## Lecture 20

We begin with a review of last lecture.
Consider a vector space $V$. A tensor $T \in \mathcal{L}^{k}$ is decomposable if $T=\ell_{1} \otimes \cdots \otimes$ $\ell_{k}, \ell_{i} \in \mathcal{L}^{1}=V^{*}$. A decomposable tensor $T$ is redundant of $\ell_{i}=\ell_{i+1}$ for some $i$. We define

$$
\begin{equation*}
\mathcal{I}^{k}=\mathcal{I}^{k}(V)=\operatorname{Span}\{\text { redundant } k \text {-tensors }\} \tag{4.94}
\end{equation*}
$$

Because $\mathcal{I}^{k} \subseteq \mathcal{L}^{k}$, we can take the quotient space

$$
\begin{equation*}
\Lambda^{k}=\Lambda^{k}\left(V^{*}\right)=\mathcal{L}^{k} / \mathcal{I}^{k} \tag{4.95}
\end{equation*}
$$

defining the map

$$
\begin{equation*}
\pi: \mathcal{L}^{k} \rightarrow \Lambda^{k} \tag{4.96}
\end{equation*}
$$

We denote by $\mathcal{A}^{k}(V)$ the set of all alternating $k$-tensors. We repeat the main theorem from last lecture:
Theorem 4.27. The map $\pi$ maps $\mathcal{A}^{k}$ bijectively onto $\Lambda^{k}$. So, $\mathcal{A}^{k} \cong \Lambda^{k}$.
It is easier to understand the space $\mathcal{A}^{k}$, but many theorems are much simpler when using $\Lambda^{k}$. This ends the review of last lecture.

### 4.6 Wedge Product

Now, let $T_{1} \in \mathcal{I}^{k_{1}}$ and $T_{2} \in \mathcal{L}^{k_{2}}$. Then $T_{1} \otimes T_{2}$ and $T_{2} \otimes T_{1}$ are in $\mathcal{I}^{k}$, where $k=k_{1}+k_{2}$. The following is an example of the usefulness of $\Lambda^{k}$.

Let $\mu_{i} \in \Lambda^{k_{i}}, i=1,2$. So, $\mu_{i}=\pi\left(T_{i}\right)$ for some $T_{i} \in \mathcal{L}^{k_{i}}$. Define $k=k_{1}+k_{2}$, so $T_{1} \otimes T_{2} \in \mathcal{L}^{k}$. Then, we define

$$
\begin{equation*}
\pi\left(T_{1} \otimes T_{2}\right)=\mu_{1} \wedge \mu_{2} \in \Lambda^{k} \tag{4.97}
\end{equation*}
$$

Claim. The product $\mu_{i} \wedge \mu_{2}$ is well-defined.
Proof. Take any tensors $T_{i}^{\prime} \in \mathcal{L}^{k_{i}}$ with $\pi\left(T_{i}^{\prime}\right)=\mu_{i}$. We check that

$$
\begin{equation*}
\pi\left(T_{1}^{\prime} \otimes T_{2}^{\prime}\right)=\pi\left(T_{1} \otimes T_{2}\right) \tag{4.98}
\end{equation*}
$$

We can write

$$
\begin{align*}
& T_{1}^{\prime}=T_{1}+W_{1}, \text { where } W_{1} \in \mathcal{I}^{k_{1}}  \tag{4.99}\\
& T_{2}^{\prime}=T_{2}+W_{2}, \text { where } W_{2} \in \mathcal{I}^{k_{2}} \tag{4.100}
\end{align*}
$$

Then,

$$
\begin{equation*}
T_{1}^{\prime} \otimes T_{2}^{\prime}=T_{1} \otimes T_{2}+\underbrace{W_{1} \otimes T_{2}+T_{1} \otimes W_{2}+W_{1} \otimes W_{2}}_{\in \mathcal{I}^{k}} \tag{4.101}
\end{equation*}
$$

so

$$
\begin{equation*}
\mu_{1} \wedge \mu_{2} \equiv \pi\left(T_{1}^{\prime} \otimes T_{2}^{\prime}\right)=\pi\left(T_{1} \otimes T_{2}\right) \tag{4.102}
\end{equation*}
$$

This product $(\wedge)$ is called the wedge product. We can define higher order wedge products. Given $\mu_{i} \in \Lambda^{k_{i}}, i=1,2,3$, where $\mu=\pi\left(T_{i}\right)$, we define

$$
\begin{equation*}
\mu_{1} \wedge \mu_{2} \wedge \mu_{3}=\pi\left(T_{1} \otimes T_{2} \otimes T_{3}\right) \tag{4.103}
\end{equation*}
$$

We leave as an exercise to show the following claim.

## Claim.

$$
\begin{align*}
\mu_{1} \wedge \mu_{2} \wedge \mu_{3} & =\left(\mu_{1} \wedge \mu_{2}\right) \wedge \mu_{3} \\
& =\mu_{1} \wedge\left(\mu_{2} \wedge \mu_{3}\right) \tag{4.104}
\end{align*}
$$

Proof Hint: This triple product law also holds for the tensor product.
We leave as an exercise to show that the two distributive laws hold:
Claim. If $k_{1}=k_{2}$, then

$$
\begin{equation*}
\left(\mu_{1}+\mu_{2}\right) \wedge \mu_{3}=\mu_{1} \wedge \mu_{3}+\mu_{2} \wedge \mu_{3} \tag{4.105}
\end{equation*}
$$

If $k_{2}=k_{3}$, then

$$
\begin{equation*}
\mu_{1} \wedge\left(\mu_{2}+\mu_{3}\right)=\mu_{1} \wedge \mu_{2}+\mu_{1} \wedge \mu_{3} \tag{4.106}
\end{equation*}
$$

Remember that $\mathcal{I}^{1}=\{0\}$, so $\Lambda^{1}=\Lambda^{1} / \mathcal{I}^{1}=\mathcal{L}^{1}=\mathcal{L}^{1}(V)=V^{*}$. That is, $\Lambda^{1}\left(V^{*}\right)=V^{*}$.

Definition 4.28. The element $\mu \in \Lambda^{k}$ is decomposable if it is of the form $\mu=$ $\ell_{1} \wedge \cdots \wedge \ell_{k}$, where each $\ell_{i} \in \Lambda^{1}=V^{*}$.

That means that $\mu=\pi\left(\ell_{1} \otimes \cdots \otimes \ell_{k}\right)$ is the projection of a decomposable $k$-tensor.
Take a permutation $\sigma \in S_{k}$ and an element $\omega \in \Lambda^{k}$ such that $\omega=\pi(T)$, where $T \in \mathcal{L}^{k}$.

## Definition 4.29.

$$
\begin{equation*}
\omega^{\sigma}=\pi\left(T^{\sigma}\right) \tag{4.107}
\end{equation*}
$$

We need to check that this definition does not depend on the choice of $T$.
Claim. Define $\omega^{\sigma}=\pi\left(T^{\sigma}\right)$. Then,

1. The above definition does not depend on the choice of $T$,
2. $\omega^{\sigma}=(-1)^{\sigma} \omega$.

Proof. 1. Last lecture we proved that for $T \in \mathcal{L}^{k}$,

$$
\begin{equation*}
T^{\sigma}=(-1)^{\sigma} T+W \tag{4.108}
\end{equation*}
$$

for some $W \in \mathcal{I}^{k}$. Hence, if $T \in \mathcal{I}^{k}$, then $T^{\sigma} \in \mathcal{I}^{k}$. If $\omega=\pi(T)=\pi\left(T^{\prime}\right)$, then $T^{\prime}-T \in \mathcal{I}^{k}$. Thus, $\left(T^{\prime}\right)^{\sigma}-T^{\sigma} \in \mathcal{I}^{k}$, so $\omega^{\sigma}=\pi\left(\left(T^{\prime}\right)^{\sigma}\right)=\pi\left(T^{\sigma}\right)$.
2.

$$
\begin{equation*}
T^{\sigma}=(-1)^{\sigma} T+W \tag{4.109}
\end{equation*}
$$

for some $W \in \mathcal{I}^{k}$, so

$$
\begin{equation*}
\pi\left(T^{\sigma}\right)=(-1)^{\sigma} \pi(T) \tag{4.110}
\end{equation*}
$$

That is,

$$
\begin{equation*}
\omega^{\sigma}=(-1)^{\sigma} \omega \tag{4.111}
\end{equation*}
$$

Suppose $\omega$ is decomposable, so $\omega=\ell_{1} \wedge \cdots \wedge \ell_{k}, \ell_{i} \in V^{*}$. Then $\omega=\pi\left(\ell_{1} \wedge \cdots \wedge \ell_{k}\right)$, so

$$
\begin{align*}
\omega^{\sigma} & =\pi\left(\left(\ell_{1} \otimes \cdots \otimes \ell_{k}\right)^{\sigma}\right) \\
& =\pi\left(\ell_{\sigma(1)} \otimes \cdots \otimes \ell_{\sigma(k)}\right)  \tag{4.112}\\
& =\ell_{\sigma(1)} \wedge \cdots \wedge \ell_{\sigma(k)} .
\end{align*}
$$

Using the previous claim,

$$
\begin{equation*}
\ell_{\sigma(1)} \wedge \cdots \wedge \ell_{\sigma(k)}=(-1)^{\sigma} \ell_{1} \wedge \cdots \wedge \ell_{k} \tag{4.113}
\end{equation*}
$$

For example, if $k=2$, then $\sigma=\tau_{1,2}$. So, $\ell_{2} \wedge \ell_{1}=-\ell_{1} \wedge \ell_{2}$. In the case $k=3$, we find that

$$
\begin{align*}
\left(\ell_{1} \wedge \ell_{2}\right) \wedge \ell_{3} & =\ell_{1} \wedge\left(\ell_{2} \wedge \ell_{3}\right) \\
& =-\ell_{1} \wedge\left(\ell_{3} \wedge \ell_{2}\right)=-\left(\ell_{1} \wedge \ell_{3}\right) \wedge \ell_{2}  \tag{4.114}\\
& =\ell_{3} \wedge\left(\ell_{1} \wedge \ell_{2}\right)
\end{align*}
$$

This motivates the following claim, the proof of which we leave as an exercise.
Claim. If $\mu \in \Lambda^{2}$ and $\ell \in \Lambda^{1}$, then

$$
\begin{equation*}
\mu \wedge \ell=\ell \wedge \mu \tag{4.115}
\end{equation*}
$$

Proof Hint: Write out $\mu$ as a linear combination of decomposable elements of $\Lambda^{2}$.
Now, suppose $k=4$. Moving $\ell_{3}$ and $\ell_{4}$ the same distance, we find that

$$
\begin{equation*}
\left(\ell_{1} \wedge \ell_{2}\right) \wedge\left(\ell_{3} \wedge \ell_{4}\right)=\left(\ell_{3} \wedge \ell_{4}\right) \wedge\left(\ell_{1} \wedge \ell_{2}\right) \tag{4.116}
\end{equation*}
$$

The proof of the following is an exercise.
Claim. If $\mu \in \Lambda^{2}$ and $\nu \in \Lambda^{2}$, then

$$
\begin{equation*}
\mu \wedge \nu=\nu \wedge \mu \tag{4.117}
\end{equation*}
$$

We generalize the above claims in the following:

Claim. Left $\mu \in \Lambda^{k}$ and $\nu \in \Lambda^{\ell}$. Then

$$
\begin{equation*}
\mu \wedge \nu=(-1)^{k \ell} \nu \wedge \mu \tag{4.118}
\end{equation*}
$$

Proof Hint: First assume $k$ is even, and write out $\mu$ as a product of elements all of degree two. Second, assume that $k$ is odd.

Now we try to find a basis for $\Lambda^{k}\left(V^{*}\right)$. We begin with

$$
\begin{align*}
& e_{1}, \ldots, e_{n} \text { a basis of } V  \tag{4.119}\\
& e_{1}^{*}, \ldots, e_{n}^{*} \text { a basis of } V^{*}  \tag{4.120}\\
& e_{I}^{*}=e_{i_{1}}^{*} \otimes \cdots \otimes e_{i_{k}}^{*}, I=\left(i_{1}, \ldots, i_{k}\right), 1 \leq i_{r} \leq n, \text { a basis of } \mathcal{L}^{k},  \tag{4.121}\\
& \psi_{I}=\operatorname{Alt}\left(e_{I}^{*}\right), \quad \text { 's strictly increasing, a basis of } \mathcal{A}^{k}(V) \tag{4.122}
\end{align*}
$$

We know that $\pi$ maps $\mathcal{A}^{k}$ bijectively onto $\Lambda^{k}$, so $\pi\left(\psi_{I}\right)$, where $I$ is strictly increasing, are a basis of $\Lambda^{k}\left(V^{*}\right)$.

$$
\begin{equation*}
\psi_{I}=\operatorname{Alt} e_{I}^{*}=\sum(-1)^{\sigma}\left(e_{I}^{*}\right)^{\sigma} . \tag{4.123}
\end{equation*}
$$

So,

$$
\begin{align*}
\pi\left(\psi_{I}\right) & =\sum(-1)^{\sigma} \pi\left(\left(e_{I}^{*}\right)^{\sigma}\right) \\
& =\sum(-1)^{\sigma}(-1)^{\sigma} \pi\left(e_{I}^{*}\right)  \tag{4.124}\\
& =k!\pi\left(e_{I}^{*}\right) \\
& \equiv k!\tilde{e}_{I}
\end{align*}
$$

Theorem 4.30. The elements of $\Lambda^{k}\left(V^{*}\right)$

$$
\begin{equation*}
\tilde{e}_{i_{1}}^{*} \wedge \cdots \wedge \tilde{e}_{i_{k}}^{*}, 1 \leq i_{1}<\ldots<i_{k} \leq n \tag{4.125}
\end{equation*}
$$

are a basis of $\Lambda^{k}\left(V^{*}\right)$.
Proof. The proof is above.
Let $V, W$ be vector spaces, and let $A: V \rightarrow W$ be a linear map. We previously defined the pullback operator $A^{*}: \mathcal{L}^{k}(W) \rightarrow \mathcal{L}^{k}(V)$. Also, given $T_{i} \in \mathcal{L}^{k_{i}}(W), i=$ 1,2 , we showed that $A^{*}\left(T_{1} \otimes T_{2}\right)=A^{*} T_{1} \otimes A^{*} T_{2}$. So, if $T=\ell_{1} \otimes \cdots \otimes \ell_{k} \in \mathcal{L}_{k}(W)$ is decomposable, then

$$
\begin{equation*}
A^{*} T=A^{*} \ell_{1} \otimes \cdots \otimes A^{*} \ell_{k}, \quad \ell_{i} \in W^{*} \tag{4.126}
\end{equation*}
$$

If $\ell_{i}=\ell_{i+1}$, then $A^{*} \ell_{1}=A^{*} \ell_{i+1}$. This shows that if $\ell_{1} \otimes \cdots \otimes \ell_{k}$ is redundant, then $A^{*}\left(\ell_{1} \otimes \cdots \otimes \ell_{k}\right)$ is also redundant. So,

$$
\begin{equation*}
A^{*} \mathcal{I}^{k}(W) \subseteq \mathcal{I}^{k}(V) \tag{4.127}
\end{equation*}
$$

Let $\mu \in \Lambda^{k}\left(W^{*}\right)$, so $\mu=\pi(T)$ for some $T \in \mathcal{L}^{k}(W)$. We can pullback to get $\pi\left(A^{*} T\right) \in \Lambda^{k}\left(V^{*}\right)$.

Definition 4.31. $A^{*} \mu=\pi\left(A^{*} T\right)$.
This definition makes sense. If $\mu=\pi(T)=\pi\left(T^{\prime}\right)$, then $T^{\prime}-T \in \mathcal{I}^{k}(W)$. So $A^{*} T^{\prime}-A^{*} T \in \mathcal{I}^{k}(V)$, which shows that $A^{*} \mu=\pi\left(A^{*} T^{\prime}\right)=\pi\left(A^{*} T\right)$.

We ask in the homework for you to show that the pullback operation is linear and that

$$
\begin{equation*}
A^{*}\left(\mu_{1} \wedge \mu_{2}\right)=A^{*} \mu_{1} \wedge A^{*} \mu_{2} \tag{4.128}
\end{equation*}
$$

## Lecture 21

Let $V, W$ be vector spaces, and let $A: V \rightarrow W$ be a linear map. We defined the pullback operation $A^{*}: W^{*} \rightarrow V^{*}$. Last time we defined another pullback operator having the form $A^{*}: \Lambda^{k}\left(W^{*}\right) \rightarrow \Lambda^{k}\left(V^{*}\right)$. This new pullback operator has the following properties:

1. $A^{*}$ is linear.
2. If $\omega_{i} \in \Lambda^{k_{1}}\left(W^{*}\right), i=1,2$, then $A^{*} \omega_{1} \wedge \omega_{2}=A^{*} \omega_{1} \wedge \omega_{2}$.
3. If $\omega$ is decomposable, that is if $\omega=\ell_{1} \wedge \cdots \wedge \ell_{k}$ where $\ell_{i} \in W^{*}$, then $A^{*} \omega=$ $A^{*} \ell_{1} \wedge \cdots \wedge A^{*} \ell_{k}$.
4. Suppose that $U$ is a vector space and that $B: W \rightarrow U$ is a linear map. Then, for every $\omega \in \Lambda^{k}\left(U^{*}\right), A^{*} B^{*} \omega=(B A)^{*} \omega$.

### 4.7 Determinant

Today we focus on the pullback operation in the special case where $\operatorname{dim} V=n$. So, we are studying $\Lambda^{n}\left(V^{*}\right)$, which is called the $n$th exterior power of $V$.

Note that $\operatorname{dim} \Lambda^{n}\left(V^{*}\right)=1$.
Given a linear map $A: V \rightarrow V$, what is the pullback operator

$$
\begin{equation*}
A^{*}: \Lambda^{n}\left(V^{*}\right) \rightarrow \Lambda^{n}\left(V^{*}\right) ? \tag{4.129}
\end{equation*}
$$

Since it is a linear map from a one dimensional vector space to a one dimensional vector space, the pullback operator $A^{*}$ is simply multiplication by some constant $\lambda_{A}$. That is, for all $\omega \in \Lambda^{n}\left(V^{*}\right), A^{*} \omega=\lambda_{A} \omega$.

Definition 4.32. The determinant of $A$ is

$$
\begin{equation*}
\operatorname{det}(A)=\lambda_{A} \tag{4.130}
\end{equation*}
$$

The determinant has the following properties:

1. If $A=I$ is the identity map, then $\operatorname{det}(A)=\operatorname{det}(I)=1$.
2. If $A, B$ are linear maps of $V$ into $V$, then $\operatorname{det}(A B)=\operatorname{det}(A) \operatorname{det}(B)$.

Proof: Let $\omega \in \Lambda^{n}\left(V^{*}\right)$. Then

$$
\begin{align*}
(A B)^{*} \omega & =\operatorname{det}(A B) \omega \\
& =B^{*}\left(A^{*} \omega\right)  \tag{4.131}\\
& =B^{*}(\operatorname{det} A) \omega \\
& =\operatorname{det}(A) \operatorname{det}(B) \omega .
\end{align*}
$$

3. If $A$ is onto, then $\operatorname{det}(A) \neq 0$.

Proof: Suppose that $A: V \rightarrow V$ is onto. Then there exists an inverse linear map $A^{-1}: V \rightarrow V$ such that $A A^{-1}=I$. So, $\operatorname{det}(A) \operatorname{det}\left(A^{-1}\right)=1$.
4. If $A$ is not onto, then $\operatorname{det}(A)=0$.

Proof: Let $W=\operatorname{Im}(A)$. If $A$ is not onto, then $\operatorname{dim} W<\operatorname{dim} V$. Let $B: V \rightarrow W$ be the map $A$ regarded as a map of $V$ into $W$, and let $\iota_{W}: W \rightarrow V$ be inclusion. So, $A=\iota_{W} B$. For all $\omega \in \Lambda^{n}\left(V^{*}\right), A^{*} \omega=B^{*} \iota_{W}^{*} \omega$. Note that $\iota_{W}^{*} \omega \in \Lambda^{n}\left(W^{*}\right)=\{0\}$ because $\operatorname{dim} W<n$. So, $A^{*} \omega=B^{*} \iota_{W}^{*}=0$, which shows that $\operatorname{det}(A)=0$.
Let $W, V$ be $n$-dimensional vector spaces, and let $A: V \rightarrow W$ be a linear map. We have the bases

$$
\begin{align*}
& e_{1}, \ldots, e_{n} \text { basis of } V  \tag{4.132}\\
& e_{1}^{*}, \ldots, e_{n}^{*} \text { dual basis of } V^{*},  \tag{4.133}\\
& f_{1}, \ldots, f_{n} \text { basis of } W  \tag{4.134}\\
& f_{1}^{*}, \ldots, f_{n}^{*} \text { dual basis of } W^{*} . \tag{4.135}
\end{align*}
$$

We can write $A e_{i}=\sum a_{i j} f_{j}$, so that $A$ has the associated matrix $A \sim\left[a_{i j}\right]$. Then $A^{*} f_{j}^{*}=\sum a_{j k} e_{k}^{*}$. Take $\omega=f_{1}^{*} \wedge \cdots \wedge f_{n}^{*} \in \Lambda^{n}\left(W^{*}\right)$, which is a basis vector of $\Lambda^{n}\left(W^{*}\right)$. Let us compute its pullback:

$$
\begin{align*}
A^{*}\left(f_{1}^{*} \wedge \cdots \wedge f_{n}^{*}\right) & =\left(\sum_{k_{1}=1}^{n} a_{1, k_{1}} e_{k_{1}}^{*}\right) \wedge \cdots \wedge\left(\sum_{k_{n}=1}^{n} a_{n, k_{n}} e_{k_{n}}^{*}\right)  \tag{4.136}\\
& =\sum_{k_{1}, \ldots, k_{n}}\left(a_{1, k_{1}} \ldots a_{m, k_{n}}\right) e_{k_{1}}^{*} \wedge \cdots \wedge e_{k_{n}}^{*} .
\end{align*}
$$

Note that if $k_{r}=k_{s}$, where $r \neq s$, then $e_{k_{1}}^{*} \wedge \cdots \wedge e_{k_{n}}^{*}=0$. If there are no repetitions, then there exists $\sigma \in S_{n}$ such that $k_{i}=\sigma(i)$. Thus,

$$
\begin{align*}
A^{*}\left(f_{1}^{*} \wedge \cdots \wedge f_{n}^{*}\right) & =\sum_{\sigma} a_{1, \sigma(1)} \ldots a_{n, \sigma(n)} e_{\sigma(1)}^{*} \wedge \cdots \wedge e_{\sigma(n)}^{*} \\
& =\left(\sum_{\sigma}(-1)^{\sigma} a_{1, \sigma(1)} \ldots a_{n, \sigma(n)}\right) e_{1}^{*} \wedge \cdots \wedge e_{n}^{*} \tag{4.137}
\end{align*}
$$

Therefore,

$$
\begin{equation*}
\operatorname{det}\left[a_{i j}\right]=\sum_{\sigma}(-1)^{\sigma} a_{1, \sigma(1)} \ldots a_{n, \sigma(n)} \tag{4.138}
\end{equation*}
$$

In the case where $W=V$ and each $e_{i}=f_{i}$, we set $\omega=e_{1}^{*} \wedge \cdots \wedge e_{n}^{*}$, and we get $A^{*} \omega=\operatorname{det}\left[a_{i j}\right] \omega$. So, $\operatorname{det}(A)=\operatorname{det}\left[a_{i j}\right]$.

For basic facts about determinants, see Munkres section 2. We will use these results quite a lot in future lectures. We list some of the basic results below.

Let $A=\left[a_{i j}\right]$ be an $n \times n$ matrix.

1. $\operatorname{det}(A)=\operatorname{det}\left(A^{t}\right)$. You should prove this as an exercise. You should explain the following steps:

$$
\begin{align*}
\operatorname{det}(A) & =\sum_{\sigma}(-1)^{\sigma} a_{1, \sigma(1)} \ldots a_{n, \sigma(n)} \\
& =\sum_{\tau}(-1)^{\tau} a_{\tau(1), 1} \ldots a_{\tau(n), n}, \text { where } \tau=\sigma^{-1}  \tag{4.139}\\
& =\operatorname{det}\left(A^{t}\right) .
\end{align*}
$$

2. Let

$$
A=\left[\begin{array}{ll}
B & C  \tag{4.140}\\
0 & D
\end{array}\right]
$$

where $B$ is $k \times k, C$ is $k \times \ell, D$ is $\ell \times \ell$, and $n=k+\ell$. Then

$$
\begin{equation*}
\operatorname{det}(A)=\operatorname{det}(B) \operatorname{det}(D) \tag{4.141}
\end{equation*}
$$

### 4.8 Orientations of Vector Spaces

Let $\ell \subseteq \mathbb{R}^{2}$ be a line through the origin. Then $\ell-\{0\}$ has two connected components. An orientation of $\ell$ is a choice of one of these components.

More generally, given a one-dimensional vector space $\mathbb{L}$, the set $\mathbb{L}$ has two connected components. Choose $v \in \mathbb{L}-\{0\}$. Then the two components are

$$
\begin{equation*}
\left\{\lambda v: \lambda \in \mathbb{R}_{+}\right\} \text {and }\left\{-\lambda v: \lambda \in \mathbb{R}_{+}\right\} . \tag{4.142}
\end{equation*}
$$

Definition 4.33. An orientation of $\mathbb{L}$ is a choice of one of these components, usually labeled $\mathbb{L}_{+}$. We define

$$
\begin{equation*}
v \in \mathbb{L}_{+} \Longleftrightarrow v \text { is positively oriented. } \tag{4.143}
\end{equation*}
$$

Let $V$ be an $n$-dimensional vector space. Then $\Lambda^{n}\left(V^{*}\right)$ is a 1 -dimensional vector space.

Definition 4.34. An orientation of $V$ is an orientation of $\Lambda^{n}\left(V^{*}\right)$. That is, a choice of $\Lambda^{n}\left(V^{*}\right)_{+}$.

Suppose $e_{1}, \ldots, e_{n}$ is a basis of $V$, so $e_{1}^{*}, \ldots, e_{n}^{*}$ is the dual basis of $V^{*}$. Let $\omega=e_{1}^{*} \wedge \cdots \wedge e_{n}^{*} \in \Lambda^{n}\left(V^{*}\right)-\{0\}$.
Definition 4.35. The basis $e_{1}, \ldots, e_{n}$ is positively oriented if $\omega \in \Lambda^{n}\left(V^{*}\right)_{+}$.
Let $f_{1}, \ldots, f_{n}$ be another basis of $V$ and $f_{1}^{*}, \ldots, f_{n}^{*}$ its dual basis. Let $w^{\prime}=$ $f_{1}^{*} \wedge \cdots \wedge f_{n}^{*}$. We ask: How is $\omega^{\prime}$ related to $\omega$ ? The answer: If $f_{j}=\sum a_{i j} e_{i}$, then $\omega^{\prime}=\operatorname{det}\left[a_{i j}\right] \omega$. So, if $e_{1}, \ldots, e_{n}$ is positively oriented, then $f_{1}, \ldots, f_{n}$ is positively oriented if and only if $\operatorname{det}\left[a_{i j}\right]>0$.

Suppose $V$ is an $n$-dimensional vector space and that $W$ is a $k$-dimensional subspace of $V$.

Claim. If $V$ and $V / W$ are given orientations, then $W$ acquires from these orientations a natural subspace orientation.

Idea of proof: Let $\pi: V \rightarrow V / W$ be the canonical map, and choose a basis $e_{1}, \ldots, e_{n}$ of $V$ such that $e_{\ell+1}, \ldots, e_{n}$ is a basis of $W$ and such that $\pi\left(e_{1}\right), \ldots, \pi\left(e_{\ell}\right)$ is a basis of $V / W$, where $\ell=n-k$.

Replacing $e_{1}$ by $-e_{1}$ if necessary, we can assume that $\pi\left(e_{1}\right), \ldots, \pi\left(e_{\ell}\right)$ is an oriented basis of $V / W$. Replacing $e_{n}$ by $-e_{n}$ if necessary, we can assume that $e_{1}, \ldots, e_{n}$ is an oriented basis of $V$. Now, give $W$ the orientation for which $e_{\ell+1}, \ldots, e_{n}$ is an oriented basis of $W$. One should check that this choice of orientation for $W$ is independent of the choice of basis (this is explained in the Multi-linear Algebra notes).

## Lecture 22

In $\mathbb{R}^{3}$ we had the operators grad, div, and curl. What are the analogues in $\mathbb{R}^{n}$ ? Answering this question is the goal of today's lecture.

### 4.9 Tangent Spaces and $k$-forms

Let $p \in \mathbb{R}^{n}$.
Definition 4.36. The tangent space of $p$ in $\mathbb{R}^{n}$ is

$$
\begin{equation*}
T_{p} \mathbb{R}^{n}=\left\{(p, v): v \in \mathbb{R}^{n}\right\} \tag{4.144}
\end{equation*}
$$

We identify the tangent space with $\mathbb{R}^{n}$ via the identification

$$
\begin{align*}
& T_{p} \mathbb{R}^{n} \cong \mathbb{R}^{n}  \tag{4.145}\\
&(p, v) \rightarrow v . \tag{4.146}
\end{align*}
$$

Via this identification, the vector space structure on $\mathbb{R}^{n}$ gives a vector space structure on $T_{p} \mathbb{R}^{n}$.

Let $U$ be an open set in $\mathbb{R}^{n}$, and let $f: U \rightarrow \mathbb{R}^{m}$ be a $\mathcal{C}^{1}$ map. Also, let $p \in U$ and define $q=f(p)$. We define a linear map

$$
\begin{equation*}
d f_{p}: T_{p} \mathbb{R}^{n} \rightarrow T_{q} \mathbb{R}^{m} \tag{4.147}
\end{equation*}
$$

according to the following diagram:

$$
\begin{array}{ccc}
T_{p} \mathbb{R}^{n} & \xrightarrow{d f_{p}} & T_{q} \mathbb{R}^{m} \\
\cong & & \cong \uparrow  \tag{4.148}\\
\mathbb{R}^{n} & \xrightarrow{D f(p)} & \mathbb{R}^{m} .
\end{array}
$$

So,

$$
\begin{equation*}
d f_{p}(p, v)=(q, D f(p) v) \tag{4.149}
\end{equation*}
$$

Definition 4.37. The cotangent space of $\mathbb{R}^{n}$ at $p$ is the space

$$
\begin{equation*}
T_{p}^{*} \mathbb{R}^{n} \equiv\left(T_{p} \mathbb{R}^{n}\right)^{*}, \tag{4.150}
\end{equation*}
$$

which is the dual of the tangent space of $\mathbb{R}^{n}$ at $p$.
Definition 4.38. Let $U$ be an open subset of $\mathbb{R}^{n}$. A $k$-form on $U$ is a function $\omega$ which assigns to every point $p \in U$ an element $\omega_{p}$ of $\Lambda^{k}\left(T_{p}^{*} \mathbb{R}^{n}\right)$ (the $k$ th exterior power of $T_{p}^{*} \mathbb{R}^{n}$ ).

Let us look at a simple example. Let $f \in \mathcal{C}^{\infty}(U), p \in U$, and $c=f(p)$. The map

$$
\begin{equation*}
d f_{p}: T_{p} \mathbb{R}^{n} \rightarrow T_{c} \mathbb{R}=\mathbb{R} \tag{4.151}
\end{equation*}
$$

is a linear map of $T_{p} \mathbb{R}^{n}$ into $\mathbb{R}$. That is, $d f_{p} \in T_{p}^{*} \mathbb{R}^{n}$. So, $d f$ is the one-form on $U$ which assigns to every $p \in U$ the linear map

$$
\begin{equation*}
d f_{p} \in T_{p}^{*} \mathbb{R}^{n}=\Lambda^{1}\left(T_{p}^{*} \mathbb{R}\right) \tag{4.152}
\end{equation*}
$$

As a second example, let $f, g \in \mathcal{C}^{\infty}(U)$. Then $g d f$ is the one-form that maps

$$
\begin{equation*}
p \in U \rightarrow g(p) d f_{p} \in \Lambda^{1}\left(T_{p}^{*} \mathbb{R}^{n}\right) \tag{4.153}
\end{equation*}
$$

As a third example, let $f, g \in \mathcal{C}^{\infty}(U)$. Then $\omega=d f \wedge d g$ is the two-form that maps

$$
\begin{equation*}
p \in U \rightarrow d f_{p} \wedge d g_{p} \tag{4.154}
\end{equation*}
$$

Note that $d f_{p}, d g_{p} \in T_{p}^{*} \mathbb{R}$, so $d f_{p} \wedge d g_{p} \in \Lambda^{2}\left(T_{p}^{*} \mathbb{R}^{n}\right)$.
As a fourth and final example, let $f_{1}, \ldots, f_{k} \in \mathcal{C}^{\infty}(U)$. Then $d f_{1} \wedge \cdots \wedge d f_{k}$ is the $k$-form that maps

$$
\begin{equation*}
p \in U \rightarrow\left(d f_{1}\right)_{p} \wedge \cdots \wedge\left(d f_{k}\right)_{p} . \tag{4.155}
\end{equation*}
$$

Note that each $\left(d f_{i}\right)_{p} \in T_{p}^{*} \mathbb{R}^{n}$, so $\left(d f_{1}\right)_{p} \wedge \cdots \wedge\left(d f_{k}\right)_{p} \in \Lambda^{k}\left(T_{p}^{*} \mathbb{R}^{n}\right)$.
Let us now look at what $k$-forms look like in coordinates. Let $e_{1}, \ldots, d_{n}$ be the standard basis of $\mathbb{R}^{n}$. Let $p \in U$ and let $v_{i}=\left(p, e_{i}\right)$ for each $i$. Then, the vectors $v_{1}, \ldots, v_{n}$ are a basis of $T_{p} \mathbb{R}^{n}$.

Suppose we have a map $f \in \mathcal{C}^{\infty}(U)$. What is $d f_{p}\left(v_{i}\right)$ ?

$$
\begin{equation*}
d f_{p}\left(v_{i}\right)=D e_{i} f(p)=\frac{\partial f}{\partial x_{i}}(p) \tag{4.156}
\end{equation*}
$$

In particular, letting $x_{i}$ be the $i$ th coordinate function,

$$
\left(d x_{i}\right)_{p}\left(v_{j}\right)=\frac{\partial x_{i}}{\partial x_{j}}= \begin{cases}1 & \text { if } i=j  \tag{4.157}\\ 0 & \text { if } i \neq j\end{cases}
$$

So, $\left(d x_{1}\right)_{p}, \ldots,\left(d x_{n}\right)_{p}$ is the basis of $T_{p}^{*} \mathbb{R}^{n}$ dual to $v_{1}, \ldots, v_{n}$.
For any $f \in \mathcal{C}^{\infty}(U)$,

$$
\begin{align*}
d f_{p}\left(v_{j}\right) & =\frac{\partial f}{\partial x_{j}}(p) \\
& =\left(\sum_{i} \frac{\partial f}{\partial x_{i}}(p)\left(d x_{i}\right)_{p}\right)\left(v_{j}\right)  \tag{4.158}\\
\Longrightarrow d f_{p} & =\sum \frac{\partial f}{\partial x_{i}}(p)\left(d x_{i}\right)_{p} \\
\Longrightarrow d f & =\sum \frac{\partial f}{\partial x_{i}} d x_{i}
\end{align*}
$$

Since $\left(d x_{1}\right)_{p}, \ldots,\left(d x_{n}\right)_{p}$ is a basis of $T_{p}^{*} \mathbb{R}^{n}$, the wedge products

$$
\begin{equation*}
\left(d x_{I}\right)_{p}=\left(d x_{i_{1}}\right)_{p} \wedge \cdots \wedge\left(d x_{i_{k}}\right)_{p}, \quad 1 \leq i_{1}<\cdots<i_{k} \leq n \tag{4.159}
\end{equation*}
$$

(I strictly increasing) are a basis of $\Lambda^{k}\left(T_{p}^{*} \mathbb{R}^{n}\right)$.
Therefore, any element $w_{p}$ of $\Lambda^{k}\left(T_{p}^{*} \mathbb{R}^{n}\right)$ can be written uniquely as a sum

$$
\begin{equation*}
\omega_{p}=\sum a_{I}(p)\left(d x_{I}\right)_{p}, \quad a_{I}(p) \in \mathbb{R} \tag{4.160}
\end{equation*}
$$

where the $I$ 's are strictly increasing. Hence, any $k$-form can be written uniquely as a sum

$$
\begin{equation*}
\omega=\sum a_{I} d x_{I}, \quad I \text { strictly increasing, } \tag{4.161}
\end{equation*}
$$

where each $a_{I}$ is a real-valued function on $U$. That is, $a_{I}: U \rightarrow \mathbb{R}$.
Definition 4.39. The $k$-form $\omega$ is $\mathcal{C}^{r}(U)$ if each $a_{I} \in \mathcal{C}^{r}(U)$.
Just to simplify our discussion, from now on we will always take $k$-forms that are $\mathcal{C}^{\infty}$.

Definition 4.40. We define

$$
\begin{equation*}
\Omega^{k}(U)=\text { the set of all } \mathcal{C}^{\infty} k \text {-forms. } \tag{4.162}
\end{equation*}
$$

So, $\omega \in \Omega^{k}(U)$ implies that $\omega=\sum a_{I} d x_{I}$, where $a_{I} \in \mathcal{C}^{\infty}(U)$.
Let us now study some basic operations on $k$-forms.

1. Let $\omega \in \Omega^{k}(U)$ and let $f \in \mathcal{C}^{\infty}(U)$. Then $f \omega \in \Omega^{k}(U)$ is the $k$-form that maps

$$
\begin{equation*}
p \in U \rightarrow f(p) \omega_{p} \in \Lambda^{k}\left(T_{p}^{*} \mathbb{R}^{n}\right) \tag{4.163}
\end{equation*}
$$

2. Let $\omega_{i} \in \Omega^{k}(U), i=1,2$. Then $\omega_{1}+\omega_{2}$ is the $k$-form that maps

$$
\begin{equation*}
p \in U \rightarrow\left(\omega_{1}\right)_{p}+\left(\omega_{2}\right)_{p} \in \Lambda^{k}\left(T_{p}^{*} \mathbb{R}^{n}\right) \tag{4.164}
\end{equation*}
$$

3. Let $\omega_{i} \in \Omega^{k_{i}}(U), i=1,2$, and $k=k_{1}+k_{2}$. Then $w_{1} \wedge \omega_{2} \in \Omega^{k}(U)$ is the $k$-form that maps

$$
\begin{equation*}
p \in U \rightarrow\left(\omega_{1}\right)_{p} \wedge\left(\omega_{2}\right)_{p} \in \Lambda^{k}\left(T_{p}^{*} \mathbb{R}^{n}\right) \tag{4.165}
\end{equation*}
$$

since $\left(\omega_{i}\right)_{p} \in \Lambda^{k_{i}}\left(T_{p}^{*} \mathbb{R}^{n}\right)$.
Definition 4.41. We find it convenient to define $\Lambda^{0}\left(T_{p}^{*} \mathbb{R}^{n}\right)=\mathbb{R}$.

A zero-form $f$ on $U$ is just a real-valued function, so $\Omega^{0}(U)=\mathcal{C}^{\infty}(\mathbb{R})$.
Take $f \in \mathcal{C}^{\infty}(U)$ and $d f \in \Omega^{1}(U)$. This gives an operation

$$
\begin{align*}
d: \Omega^{0}(U) & \rightarrow \Omega^{1}(U),  \tag{4.166}\\
f & \rightarrow d f . \tag{4.167}
\end{align*}
$$

Let $f, g \in \mathcal{C}^{\infty}(U)$ (that is, take $f, g$ to be zero-forms). Then $d(f g)=g d f+f d g$. We can think of this operation as a slightly different notation for the gradient operation.

The maps $d: \Omega^{k}(U) \rightarrow \Omega^{k+1}(U), k=0, \ldots,(n-1)$ give $n$ vector calculus operations.

If $\omega \in \Omega^{k}(U)$, then $\omega$ can be written uniquely as the sum

$$
\begin{equation*}
\omega=\sum a_{I} d x_{I}, I \text { strictly increasing, } \tag{4.168}
\end{equation*}
$$

where each $a_{I} \in \mathcal{C}^{\infty}(U)$. We define

$$
\begin{equation*}
d \omega=\sum d a_{I} \wedge d x_{I} \tag{4.169}
\end{equation*}
$$

This operator is the unique operator with the following properties:

1. For $k=0$, this is the operation we already defined, $d f=\sum \frac{\partial f}{\partial x_{i}} d x_{i}$.
2. If $\omega \in \Omega^{k}$, then $d(d \omega)=0$.
3. If $\omega_{i} \in \Omega^{k_{i}}(U), i=1,2$, then $d\left(\omega_{1} \wedge \omega_{2}\right)=d \omega_{1} \wedge \omega_{2}+(-1)^{k_{1}} \omega_{1} \wedge d \omega_{2}$.

Let $a \in \mathcal{C}^{\infty}(U)$, and $a d x_{I} \in \Omega^{k}(U)$, I strictly increasing. Then

$$
\begin{equation*}
d\left(a d x_{I}\right)=d a \wedge d x_{I} . \tag{4.170}
\end{equation*}
$$

Suppose that $I$ is not strictly increasing. Then

$$
\begin{align*}
d x_{I} & =d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}  \tag{4.171}\\
& =0 \text { if } i_{r}=i_{s} .
\end{align*}
$$

If there are no repetitions, then there exists $\sigma \in S_{k}$ such that $J=I^{\sigma}$ is strictly increasing. Then

$$
\begin{equation*}
d x_{J}=(-1)^{\sigma} d x_{I}, \tag{4.172}
\end{equation*}
$$

so

$$
\begin{align*}
d\left(a d x_{I}\right) & =(-1)^{\sigma} d\left(a d x_{J}\right) \\
& =(-1)^{\sigma} d a \wedge d x_{J}  \tag{4.173}\\
& =d a \wedge d x_{I} .
\end{align*}
$$

Putting this altogether, for every multi-index $I$ of length $k$,

$$
\begin{equation*}
d\left(a d x_{I}\right)=d a \wedge d x_{I} . \tag{4.174}
\end{equation*}
$$

## Lecture 23

Let $U$ be an open set in $\mathbb{R}^{n}$. For each $k=0, \ldots, n-1$, we define the differential operator

$$
\begin{equation*}
d: \Omega^{k}(U) \rightarrow \Omega^{k+1}(U) \tag{4.175}
\end{equation*}
$$

These maps are the $n$ basic vector calculus operations in $n$-dimensional calculus. We review how $d$ is defined.

For $k=0, \Omega^{0}(U)=\mathcal{C}^{\infty}(U)$. Let $f \in \mathcal{C}^{\infty}(U)$, and let $c=f(p)$, where $p \in U$. The mapping $d f_{p}: T_{p} \mathbb{R}^{n} \rightarrow T_{c} \mathbb{R}=\mathbb{R}$ maps $T_{p} \mathbb{R}^{n}$ to $\mathbb{R}$, so $d f_{p} \in T_{p}^{*} \mathbb{R}^{n}$. The map $d f \in \Omega^{1}(U)$ is a one-form that maps $p \in U$ to $d f_{p} \in T_{p}^{*} \mathbb{R}^{n}$. A formula for this in coordinates is

$$
\begin{equation*}
d f=\sum \frac{\partial f}{\partial x_{i}} d x_{i} . \tag{4.176}
\end{equation*}
$$

In $k$ dimensions, $d$ is a map

$$
\begin{equation*}
d: \Omega^{k}(U) \rightarrow \Omega^{k+1}(U) \tag{4.177}
\end{equation*}
$$

Given $\omega \in \Omega^{k}(U), \omega$ can be written uniquely as

$$
\begin{align*}
\omega & =\sum_{I} a_{I} d x_{I} \\
& =\sum_{I} a_{I} d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}, \tag{4.178}
\end{align*}
$$

where $i_{1}<\cdots<i_{k}$ and each $a_{I} \in \mathcal{C}^{\infty}(U)$. Then, we define

$$
\begin{align*}
d \omega & =\sum d a_{I} \wedge d x_{I} \\
& =\sum_{i, I} \frac{\partial a_{I}}{\partial x_{i}} d x_{i} \wedge d x_{I} \tag{4.179}
\end{align*}
$$

where each $I$ is strictly increasing.
The following are some basic properties of the differential operator $d$ :

1. If $\mu \in \Omega^{k}(U)$ and $\nu \in \Omega^{\ell}(U)$, then

$$
\begin{equation*}
d \mu \wedge \nu=d \mu \wedge \nu+(-1)^{k} \mu \wedge d \nu \tag{4.180}
\end{equation*}
$$

2. For and $\omega \in \Omega^{k}(U)$,

$$
\begin{equation*}
d(d \omega)=0 . \tag{4.181}
\end{equation*}
$$

Remark. Let $I$ be any multi-index, and let $a_{I} \in \mathcal{C}^{\infty}(U)$. Then

$$
\begin{equation*}
d\left(a_{I} d x_{I}\right)=d a_{I} \wedge d x_{I} . \tag{4.182}
\end{equation*}
$$

We now prove the above two basic properties of the differential operator.
Claim. If $\mu \in \Omega^{k}(U)$ and $\nu \in \Omega^{\ell}(U)$, then

$$
\begin{equation*}
d \mu \wedge \nu=d \mu \wedge \nu+(-1)^{k} \mu \wedge d \nu \tag{4.183}
\end{equation*}
$$

Proof. Take $\mu=\sum a_{I} d x_{I}$ and $\nu=\sum b_{J} d x_{J}$, where $I, J$ are strictly increasing. Then

$$
\begin{equation*}
\mu \wedge \nu=\sum a_{I} b_{J} \underbrace{d x_{I} \wedge d x_{J}}_{\text {no longer increasing }} \tag{4.184}
\end{equation*}
$$

Then

$$
\begin{align*}
d(\mu \wedge \nu)= & \sum_{i, I, J} \frac{\partial a_{I} b_{J}}{\partial x_{i}} d x_{i} \wedge d x_{I} \wedge d x_{J} \\
= & \sum \frac{\partial a_{I}}{\partial x_{i}} b_{J} d x_{i} \wedge d x_{I} \wedge d x_{J}  \tag{4.185}\\
& +\sum a_{I} \frac{\partial b_{J}}{\partial x_{i}} d x_{i} \wedge d x_{I} \wedge d x_{J}, \quad(I I
\end{align*}
$$

We calculate sums (I) and (II) separately.

$$
\begin{align*}
(I) & =\sum_{i, I, J} \frac{\partial a_{I}}{\partial x_{i}} d x_{i} \wedge d x_{I} \wedge b_{J} d x_{J} \\
& =\left(\sum_{i, I} \frac{\partial a_{I}}{\partial x_{i}} d x_{i} \wedge d x_{I}\right) \wedge \sum_{J} b_{J} d x_{J}  \tag{4.186}\\
& =d \mu \wedge \nu . \\
(I I)= & \sum_{i, I, J} a_{I} \frac{\partial b_{J}}{\partial x_{i}} d x_{i} \wedge d x_{I} \wedge d x_{J} \\
& =(-1)^{k} \sum_{i, I, J} a_{I} d x_{I} \wedge \frac{\partial b_{J}}{\partial x_{i}} d x_{i} \wedge d x_{J}  \tag{4.187}\\
& =(-1)^{k}\left(\sum_{I} a_{I} d x_{I}\right) \wedge \sum_{i, J} \frac{\partial b_{J}}{\partial x_{i}} d x_{i} \wedge d x_{J} \\
& =(-1)^{k} \mu \wedge d \nu .
\end{align*}
$$

So,

$$
\begin{align*}
d(\mu \wedge \nu) & =(I)+(I I) \\
& =d \mu \wedge \nu+(-1)^{k} \mu \wedge d \nu \tag{4.188}
\end{align*}
$$

Claim. For and $\omega \in \Omega^{k}(U)$,

$$
\begin{equation*}
d(d \omega)=0 \tag{4.189}
\end{equation*}
$$

Proof. Let $\omega=\sum a_{I} d x_{I}$, so

$$
\begin{equation*}
d \omega=\sum_{j, I} \frac{\partial a_{I}}{\partial x_{j}} d x_{j} \wedge d x_{I} \tag{4.190}
\end{equation*}
$$

Then,

$$
\begin{equation*}
d(d \omega)=\sum_{i, j, I} \frac{\partial^{2} a_{I}}{\partial x_{i} \partial x_{j}} d x_{i} \wedge d x_{j} \wedge d x_{I} \tag{4.191}
\end{equation*}
$$

Note that if $i=j$, then there is a repeated term in the wedge product, so

$$
\begin{align*}
d(d \omega)= & \sum_{i<j} \frac{\partial^{2} a_{I}}{\partial x_{i} \partial x_{j}} d x_{i} \wedge d x_{j} \wedge d x_{I}  \tag{4.192}\\
& +\sum_{i>j} \frac{\partial^{2} a_{I}}{\partial x_{i} \partial x_{j}} d x_{i} \wedge d x_{j} \wedge d x_{I} \tag{4.193}
\end{align*}
$$

Note that $d x_{i} \wedge d x_{j}=-d x_{j} \wedge d x_{i}$. We relabel the second summand to obtain

$$
\begin{align*}
d(d \omega) & =\sum_{i<j}(\underbrace{\frac{\partial^{2} a_{I}}{\partial x_{i} \partial x_{j}}-\frac{\partial^{2} a_{I}}{\partial x_{j} \partial x_{i}}}_{0}) d x_{i} \wedge d x_{j} \wedge d x_{I}  \tag{4.194}\\
& =0
\end{align*}
$$

Definition 4.42. A $k$-form $\omega \in \Omega^{k}(U)$ is decomposable if $\omega=\mu_{1} \wedge \cdots \wedge \mu_{k}$, where each $\mu_{i} \in \Omega^{1}(U)$.

Theorem 4.43. If $\omega$ is decomposable, then

$$
\begin{equation*}
d \omega=\sum_{i=1}^{k}(-1)^{i-1} \mu_{1} \wedge \cdots \wedge \mu_{i-1} \wedge d \mu_{i} \wedge \mu_{i+1} \wedge \cdots \wedge \mu_{k} \tag{4.195}
\end{equation*}
$$

Proof. The proof is by induction.

The case $k=1$ is obvious. We show that if the theorem is true for $k-1$, then the theorem is true for $k$.

$$
\begin{align*}
d\left(\left(\mu_{1} \wedge \cdots \wedge \mu_{k-1}\right) \wedge \mu_{k}\right)= & \left(d\left(\mu_{1} \wedge \cdots \wedge \mu_{k-1}\right)\right) \wedge \mu_{k} \\
& +(-1)^{k-1}\left(\mu_{1} \wedge \cdots \wedge \mu_{k-1}\right) \wedge d \mu_{k} \\
= & \sum_{i=1}^{k-1}(-1)^{i-1} \mu_{1} \wedge \cdots \wedge d \mu_{i} \wedge \cdots \wedge \mu_{k-1} \wedge \mu_{k}  \tag{4.196}\\
& +(-1)^{k-1}\left(\mu_{1} \wedge \cdots \wedge \mu_{k-1} \wedge \mu_{k}\right) \\
= & \sum_{i=1}^{k}(-1)^{i-1} \mu_{1} \wedge \cdots \wedge d \mu_{i} \wedge \cdots \wedge \mu_{k}
\end{align*}
$$

### 4.10 Pullback Operation on Exterior Forms

Another important operation in the theory of exterior forms is the pullback operator. This operation is not introduced in 18.01 or 18.02 , because vector calculus in not usually taught rigorously.

Let $U$ be open in $\mathbb{R}^{n}$ and $V$ be open in $\mathbb{R}^{m}$, and let $f: U \rightarrow V$ be a $\mathcal{C}^{\infty}$ map. We can write out in components $f=\left(f_{1}, \ldots, f_{n}\right)$, where each $f_{i} \in \mathcal{C}^{\infty}(U)$. Let $p \in U$ and $q=f(p)$.

The pullback of the map $d f_{p}: T_{p} \mathbb{R}^{m} \rightarrow T_{q} \mathbb{R}^{n}$ is

$$
\begin{equation*}
\left(d f_{p}\right)^{*}: \Lambda^{k}\left(T_{q}^{*} \mathbb{R}^{n}\right) \rightarrow \Lambda^{k}\left(T_{p}^{*} \mathbb{R}^{m}\right) \tag{4.197}
\end{equation*}
$$

Suppose you have a $k$-form $\omega$ on $V$.

$$
\begin{align*}
& \omega \in \Omega^{k}(V)  \tag{4.198}\\
& \omega_{q} \in \Lambda^{k}\left(T_{q}^{*} \mathbb{R}^{n}\right) \tag{4.199}
\end{align*}
$$

Then

$$
\begin{equation*}
\left(d f_{p}\right)^{*} w_{q} \in \Lambda^{k}\left(T_{p}^{*} \mathbb{R}^{m}\right) \tag{4.200}
\end{equation*}
$$

Definition 4.44. $f^{*} \omega$ is the $k$-from whose value at $p \in U$ is $\left(d f_{p}\right)^{*} \omega_{q}$.
We consider two examples. Suppose $\phi \in \Omega^{0}(V)=\mathcal{C}^{\infty}(V)$. Then $f^{*} \phi(p)=\phi(q)$, so $f^{*} \phi=\phi \circ f$, where $f: U \rightarrow V$ and $\phi: V \rightarrow \mathbb{R}$.

Again, suppose that $\phi \in \Omega^{0}(V)=\mathcal{C}^{\infty}(V)$. What is $f^{*} d \phi$ ? Let $f(p)=q$. We have the $\operatorname{map} d \phi_{q}: T_{p} \mathbb{R}^{n} \rightarrow T_{c} \mathbb{R}=\mathbb{R}$, where $c=\phi(q)$. So,

$$
\begin{align*}
\left(d f_{p}\right)^{*}(d \phi)_{q} & =d \phi_{q} \circ d f_{p}  \tag{4.201}\\
& =d(\phi \circ f)_{p}
\end{align*}
$$

Therefore,

$$
\begin{equation*}
f^{*} d \phi=d f^{*} \phi \tag{4.202}
\end{equation*}
$$

Suppose that $\mu \in \Omega^{k}(V)$ and $\nu \in \Omega^{e} l l(V)$. Then

$$
\begin{align*}
\left(f^{*}(\mu \wedge \nu)\right)_{p} & =\left(d f_{p}\right)^{*}\left(\mu_{q} \wedge \nu_{q}\right)  \tag{4.203}\\
& =\left(d f_{p}\right)^{*} \mu_{q} \wedge\left(d f_{p}^{*}\right) \nu_{q}
\end{align*}
$$

Hence,

$$
\begin{equation*}
f^{*}(\mu \wedge \nu)=f^{*} \mu \wedge f^{*} \nu \tag{4.204}
\end{equation*}
$$

We now obtain a coordinate formula for $f^{*}$.
Take $\omega \in \Omega^{k}(V)$. We can write $\omega=\sum a_{I} d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}$, where each $a_{I} \in \mathcal{C}^{\infty}(U)$. Then

$$
\begin{align*}
f^{*} \omega & =\sum f^{*} a_{I} f^{*} d x_{i_{1}} \wedge \cdots \wedge f^{*} d x_{i_{k}} \\
& =\sum f^{*} a_{I} d f_{i_{1}} \wedge \cdots \wedge d f_{i_{k}} \tag{4.205}
\end{align*}
$$

where we used the result that $f^{*} d x_{i}=d x_{i} \circ f=d f_{i}$.
Note that $d f_{i}=\sum \frac{\partial f_{i}}{\partial x_{j}} d x_{j}$, where $\frac{\partial f_{i}}{\partial x_{j}} \in \mathcal{C}^{\infty}(U)$. Also, $f^{*} a_{I}=a_{I} \circ f \in \mathcal{C}^{\infty}(U)$, which shows that

$$
\begin{equation*}
f^{*} \omega \in \Omega^{k}(U) \tag{4.206}
\end{equation*}
$$

The following theorem states a very useful property of the pullback operator.
Theorem 4.45. Let $\omega \in \Omega^{k}(V)$. Then,

$$
\begin{equation*}
d f^{*} \omega=f^{*} d \omega \tag{4.207}
\end{equation*}
$$

Proof. We have already checked this for $\omega=\phi \in \mathcal{C}^{\infty}(V), k=0$ already. We now prove the general case.

We can write $\omega=\sum a_{I} d x_{I}$. Then

$$
\begin{equation*}
f^{*} \omega=\sum f^{*} a_{I} d f_{i_{1}} \wedge \cdots \wedge d f_{i_{k}} \tag{4.208}
\end{equation*}
$$

So,

$$
\begin{align*}
d f^{*} \omega= & \sum d f^{*} a_{I} \wedge d f_{i_{1}} \wedge \cdots \wedge d f_{i_{k}}  \tag{4.209}\\
& +\sum f^{*} a_{I} \wedge d\left(d f_{i_{1}} \wedge \cdots \wedge d f_{i_{k}}\right)
\end{align*}
$$

Note that

$$
\begin{equation*}
d\left(d f_{i_{1}} \wedge \cdots \wedge d f_{i_{k}}\right)=\sum_{r=1}^{k}(-1)^{r-1} d f_{i_{1}} \wedge \cdots \wedge d\left(d f_{i_{r}}\right) \wedge \cdots \wedge d f_{i_{k}} \tag{4.210}
\end{equation*}
$$

We know that $d\left(d f_{i_{r}}\right)=0$, so

$$
\begin{align*}
d f^{*} \omega & =\sum_{I} d f^{*} a_{I} \wedge d f_{i_{1}} \wedge \cdots \wedge d f_{i_{k}} \\
& =\sum_{I} f^{*} d a_{I} \wedge f^{*}\left(d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}\right)  \tag{4.211}\\
& =f^{*}\left(\sum d a_{I} \wedge d x_{I}\right) \\
& =f^{*} d \omega .
\end{align*}
$$

## Lecture 24

We review the pullback operation from last lecture. Let $U$ be open in $\mathbb{R}^{m}$ and let $V$ be open in $\mathbb{R}^{n}$. Let $f: U \rightarrow V$ be a $\mathcal{C}^{\infty}$ map, and let $f(p)=q$. From the map

$$
\begin{equation*}
d f_{p}: T_{p} \mathbb{R}^{m} \rightarrow T_{q} \mathbb{R}^{n}, \tag{4.212}
\end{equation*}
$$

we obtain the pullback map

$$
\begin{align*}
\left(d f_{p}\right)^{*}: \Lambda^{k}\left(T_{q}^{*}\right) & \rightarrow \Lambda^{k}\left(T_{p}^{*}\right)  \tag{4.213}\\
\omega \in \Omega^{k}(V) & \rightarrow f^{*} \omega \in \Omega^{k}(U) .
\end{align*}
$$

We define, $f^{*} \omega_{p}=\left(d f_{p}\right)^{*} \omega_{q}$, when $\omega_{q} \in \Lambda^{k}\left(T_{q}^{*}\right)$.
The pullback operation has some useful properties:

1. If $\omega_{i} \in \Omega^{k_{i}}(V), i=1,2$, then

$$
\begin{equation*}
f^{*}\left(\omega_{1} \wedge \omega_{2}\right)=f^{*} \omega_{1} \wedge f^{*} \omega_{2} \tag{4.214}
\end{equation*}
$$

2. If $\omega \in \Omega^{k}(V)$, then

$$
\begin{equation*}
d f^{*} \omega=f^{*} d \omega . \tag{4.215}
\end{equation*}
$$

We prove some other useful properties of the pullback operation.
Claim. For all $\omega \in \Omega^{k}(W)$,

$$
\begin{equation*}
f^{*} g^{*} \omega=(g \circ f)^{*} \omega \tag{4.216}
\end{equation*}
$$

Proof. Let $f(p)=q$ and $g(q)=w$. We have the pullback maps

$$
\begin{align*}
\left(d f_{p}\right)^{*}: \Lambda^{k}\left(T_{q}^{*}\right) & \rightarrow \Lambda^{k}\left(T_{p}^{*}\right)  \tag{4.217}\\
\left(d g_{q}\right)^{*}: \Lambda^{k}\left(T_{w}^{*}\right) & \rightarrow \Lambda^{k}\left(T_{q}^{*}\right)  \tag{4.218}\\
(g \circ f)^{*}: \Lambda^{k}\left(T_{w}^{*}\right) & \rightarrow \Lambda^{k}\left(T_{p}^{*}\right) . \tag{4.219}
\end{align*}
$$

The chain rule says that

$$
\begin{equation*}
(d g \circ f)_{p}=(d g)_{q} \circ(d f)_{p}, \tag{4.220}
\end{equation*}
$$

so

$$
\begin{equation*}
d(g \circ f)_{p}^{*}=\left(d f_{p}\right)^{*}\left(d g_{q}\right)^{*} \tag{4.221}
\end{equation*}
$$

Let $U, V$ be open sets in $\mathbb{R}^{n}$, and let $f: U \rightarrow V$ be a $\mathcal{C}^{\infty}$ map. We consider the pullback operation on $n$-forms $\omega \in \Omega^{n}(V)$. Let $f(0)=q$. Then

$$
\begin{array}{ll}
\left(d x_{i}\right)_{p}, & i=1, \ldots, n, \quad \text { is a basis of } T_{p}^{*}, \text { and } \\
\left(d x_{i}\right)_{q}, & i=1, \ldots, n, \quad \text { is a basis of } T_{q}^{*} . \tag{4.223}
\end{array}
$$

Using $f_{i}=x_{i} \circ f$,

$$
\begin{align*}
\left(d f_{p}\right)^{*}\left(d x_{i}\right)_{q} & =\left(d f_{i}\right)_{p} \\
& =\sum \frac{\partial f_{i}}{\partial x_{j}}(p)\left(d x_{j}\right)_{p} . \tag{4.224}
\end{align*}
$$

In the Multi-linear Algebra notes, we show that

$$
\begin{equation*}
\left(d f_{p}\right)^{*}\left(d x_{1}\right)_{q} \wedge \cdots \wedge\left(d x_{n}\right)_{q}=\operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}(p)\right]\left(d x_{1}\right)_{p} \wedge \cdots \wedge\left(d x_{n}\right)_{p} \tag{4.225}
\end{equation*}
$$

So,

$$
\begin{equation*}
f^{*} d x_{1} \wedge \cdots \wedge d x_{n}=\operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}\right] d x_{1} \wedge \cdots \wedge d x_{n} \tag{4.226}
\end{equation*}
$$

Given $\omega=\phi(x) d x_{1} \wedge \cdots \wedge d x_{n}$, where $\phi \in \mathcal{C}^{\infty}$,

$$
\begin{equation*}
f^{*} \omega=\phi(f(x)) \operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}\right] d x_{1} \wedge \cdots \wedge d x_{n} \tag{4.227}
\end{equation*}
$$

## 5 Integration with Differential Forms

Let $U$ be an open set in $\mathbb{R}^{n}$, and let $\omega \in \Omega^{k}(U)$ be a differential $k$-form.
Definition 5.1. The support of $\omega$ is

$$
\begin{equation*}
\operatorname{supp} \omega=\overline{\left\{p \in U: \omega_{p} \neq 0\right\}} \tag{5.1}
\end{equation*}
$$

Definition 5.2. The $k$-form $\omega$ is compactly supported if supp $\omega$ is compact. We define

$$
\begin{equation*}
\Omega_{c}^{k}(U)=\text { the space of all compactly supported } k \text {-forms. } \tag{5.2}
\end{equation*}
$$

Note that

$$
\begin{equation*}
\Omega_{c}^{0}(U)=\mathcal{C}_{0}^{\infty}\left(\mathbb{R}^{n}\right) \tag{5.3}
\end{equation*}
$$

Given $\omega \in \Omega_{c}^{n}(U)$, we can write

$$
\begin{equation*}
\omega=\phi(x) d x_{1} \wedge \cdots \wedge d x_{n} \tag{5.4}
\end{equation*}
$$

where $\phi \in \mathcal{C}_{0}^{\infty}(U)$.

## Definition 5.3.

$$
\begin{equation*}
\int_{U} \omega \equiv \int_{U} \phi=\int_{U} \phi(x) d x_{1} \ldots d x_{n} \tag{5.5}
\end{equation*}
$$

We are going to state and prove the change of variables theorem for integrals of differential $k$-forms. To do so, we first need the notions of orientation preserving and orientation reversing.

Let $U, V$ be open sets in $\mathbb{R}^{n}$. Let $f: U \rightarrow V$ be a $\mathcal{C}^{\infty}$ diffeomorphism. That is, for every $p \in U, D f(p): \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is bijective. We associate $D f(p)$ with the matrix

$$
\begin{equation*}
D f(p) \cong\left[\frac{\partial f_{i}}{\partial x_{j}}(p)\right] \tag{5.6}
\end{equation*}
$$

The map $f$ is a diffeomorphism, so

$$
\begin{equation*}
\operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}(p)\right] \neq 0 \tag{5.7}
\end{equation*}
$$

So, if $U$ is connected, then this determinant is either positive everywhere or negative everywhere.

Definition 5.4. The map $f$ is orientation preserving if det $>0$ everywhere. The map $f$ is orientation reversing if det $<0$ everywhere.

The following is the change of variables theorem:
Theorem 5.5. If $\omega \in \Omega_{c}^{n}(V)$, then

$$
\begin{equation*}
\int_{U} f^{*} \omega=\int_{V} \omega \tag{5.8}
\end{equation*}
$$

if $f$ is orientation preserving, and

$$
\begin{equation*}
\int_{U} f^{*} \omega=-\int_{V} \omega \tag{5.9}
\end{equation*}
$$

if $f$ is orientation reversing.
In Munkres and most texts, this formula is written in slightly uglier notation. Let $\omega=\phi(x) d x_{1} \wedge \cdots \wedge d x_{n}$, so

$$
\begin{equation*}
f^{*} \omega=\phi(f(x)) \operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}\right] d x_{1} \wedge \cdots \wedge d x_{n} . \tag{5.10}
\end{equation*}
$$

The theorem can be written as following:
Theorem 5.6. If $f$ is orientation preserving, then

$$
\begin{equation*}
\int_{V} \phi=\int_{U} \phi \circ f \operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}\right] . \tag{5.11}
\end{equation*}
$$

This is the coordinate version of the theorem.
We now prove a useful theorem found in the Supplementary Notes (and Spivak) called Sard's Theorem.

Let $U$ be open in $\mathbb{R}^{n}$, and let $f: U \rightarrow \mathbb{R}^{n}$ be a $\mathcal{C}^{1}(U)$ map. For every $p \in U$, we have the map $D f(p): \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$. We say that $p$ is a critical point of $f$ if $D f(p)$ is not bijective. Denote

$$
\begin{equation*}
C_{f}=\text { the set of all critical points of } f \tag{5.12}
\end{equation*}
$$

Sard's Theorem. The image $f\left(C_{f}\right)$ is of measure zero.
Proof. The proof is in the Supplementary Notes.
As an example of Sard's Theorem, let $c \in \mathbb{R}^{n}$ and let $f: U \rightarrow \mathbb{R}^{n}$ be the map defined by $f(x)=c$. Note that $D f(p)=0$ for all $p \in U$, so $C_{f}=U$. The set $C_{f}=U$ is not a set of measure zero, but $f\left(C_{f}\right)=\{c\}$ is a set of measure zero.

As an exercise, you should prove the following claim:
Claim. Sard's Theorem is true for maps $f: U \rightarrow \mathbb{R}^{n}$, where $U$ is an open, connected subset of $\mathbb{R}$.

Proof Hint: Let $f \in \mathcal{C}^{\infty}(U)$ and define $g=\frac{\partial f}{\partial x}$. The map $g$ is continuous because $f \in \mathcal{C}^{1}(U)$. Let $I=[a, b] \subseteq U$, and define $\ell=b-a$. The continuity of $g$ implies that $g$ is uniformly continuous on $I$. That is, for every $\epsilon>0$, there exists a number $N>0$ such that $|g(x)-g(y)|<\epsilon$ whenever $x, y \in I$ and $|x-y|<\ell / N$.

Now, slice $I$ into $N$ equal subintervals. Let $I_{r}, r=1, \ldots, k \leq N$ be the subintervals intersecting $C_{f}$. Prove the following lemma:
Lemma 5.7. If $x, y \in I_{r}$, then $|f(x)-f(y)|<\epsilon \ell / N$.
Proof Hint: Find $c \in I_{r}$ such that $f(x)-f(y)=(x-y) g(c)$. There exists $c_{0} \in I_{r} \cap C_{f}$ if and only if $g\left(c_{0}\right)=0$. So, we can take

$$
\begin{equation*}
|g(c)|=\left|g(c)-g\left(c_{0}\right)\right| \leq \epsilon \tag{5.13}
\end{equation*}
$$

Then $|f(x)-f(y)| \leq \epsilon \ell / N$.
From the lemma, we can conclude that

$$
\begin{equation*}
f\left(I_{r}\right) \equiv J_{r} \tag{5.14}
\end{equation*}
$$

is of length less than $\epsilon \ell / N$. Therefore,

$$
\begin{equation*}
f\left(C_{f} \cap I\right) \subset \bigcup_{r=1}^{k} J_{r} \tag{5.15}
\end{equation*}
$$

is of length less than

$$
\begin{equation*}
\frac{\epsilon \ell}{N} k \leq \frac{\epsilon \ell N}{N}=\epsilon \ell \tag{5.16}
\end{equation*}
$$

Letting $\epsilon \rightarrow 0$, we find that $F\left(C_{f} \cap I\right)$ is of measure zero.
To conclude the proof, let $I_{m}, m=1,2,3, \ldots$, be an exhaustion of $U$ by closed intervals $I_{1} \subset I_{2} \subset I_{3} \subset \cdots$ such that $\bigcup I_{m}=U$. We have shown that $f\left(C_{f} \cap I_{m}\right)$ is measure zero. So, $f\left(C_{f}\right)=\bigcup f\left(C_{f} \cap I_{m}\right)$ implies that $f\left(C_{f}\right)$ is of measure zero.

## Lecture 25

### 5.1 The Poincare Lemma

Let $U$ be an open subset of $\mathbb{R}^{n}$, and let $\omega \in \Omega^{k}(U)$ be a $k$-form. We can write $\omega=\sum a_{I} d x_{I}, I=\left(i_{1}, \ldots, i_{k}\right)$, where each $a_{I} \in \mathcal{C}^{\infty}(U)$. Note that

$$
\begin{equation*}
\omega \in \Omega_{c}^{k} \Longleftrightarrow a_{I} \in \mathcal{C}_{0}^{\infty}(U) \text { for each } I \tag{5.17}
\end{equation*}
$$

We are interested in $\omega \in \Omega_{c}^{n}(U)$, which are of the form

$$
\begin{equation*}
\omega=f d x_{1} \wedge \cdots \wedge d x_{n} \tag{5.18}
\end{equation*}
$$

where $f \in \mathcal{C}_{0}^{\infty}(U)$. We define

$$
\begin{equation*}
\int_{U} \omega=\int_{U} f=\int_{U} f d x \tag{5.19}
\end{equation*}
$$

the Riemann integral of $f$ over $U$.
Our goal over the next couple lectures is to prove the following fundamental theorem known as the Poincare Lemma.

Poincare Lemma. Let $U$ be a connected open subset of $\mathbb{R}^{n}$, and let $\omega \in \Omega_{c}^{n}(U)$. The following conditions are equivalent:

1. $\int_{U} \omega=0$,
2. $\omega=d \mu$, for some $\mu \in \Omega_{c}^{n-1}(U)$.

In today's lecture, we prove this for $U=\operatorname{Int} Q$, where $Q=\left[a_{1}, b_{1}\right] \times \cdots \times\left[a_{n}, b_{n}\right]$ is a rectangle.

Proof. First we show that (2) implies (1).

## Notation.

$$
\begin{equation*}
d x_{1} \wedge \cdots \wedge{\widehat{d x_{i}}} \wedge \cdots \wedge d x_{n} \equiv d x_{1} \wedge \cdots \wedge d x_{i-1} \wedge d x_{i+1} \wedge \cdots \wedge d x_{n} \tag{5.20}
\end{equation*}
$$

Let $\mu \in \Omega_{c}^{n-1}(U)$. Specifically, define

$$
\begin{equation*}
\mu=\sum_{i} f_{i} d x_{1} \wedge \cdots \wedge \widehat{d x_{i}} \wedge \cdots \wedge d x_{n} \tag{5.21}
\end{equation*}
$$

where each $f_{i} \in \mathcal{C}_{0}^{\infty}(U)$. Every $\mu \in \Omega_{c}^{n-1}(U)$ can be written this way.
Applying $d$ we obtain

$$
\begin{equation*}
d \mu=\sum_{i} \sum_{j} \frac{\partial f_{i}}{\partial x_{j}} d x_{j} \wedge d x_{1} \wedge \cdots \wedge{\widehat{d x_{i}}} \wedge \cdots \wedge d x_{n} \tag{5.22}
\end{equation*}
$$

Notice that if $i \neq j$, then the $i, j$ th summand is zero, so

$$
\begin{align*}
d \mu & =\sum_{i} \frac{\partial f_{i}}{\partial x_{i}} d x_{i} \wedge d x_{1} \wedge \cdots \wedge \widehat{d x_{i}} \wedge \cdots \wedge d x_{n}  \tag{5.23}\\
& =\sum(-1)^{i-1} \frac{\partial f_{i}}{\partial x_{i}} d x_{1} \wedge \cdots \wedge d x_{n}
\end{align*}
$$

Integrate to obtain

$$
\begin{equation*}
\int_{U} d \mu=\sum(-1)^{i-1} \int_{U} \frac{\partial f_{i}}{\partial x_{i}} \tag{5.24}
\end{equation*}
$$

Note that

$$
\begin{equation*}
\int_{a_{i}}^{b_{i}} \frac{\partial f_{i}}{\partial x_{i}} d x_{i}=\left.f_{i}(x)\right|_{x_{i}=a_{i}} ^{x_{i}=b_{i}}=0-0=0 \tag{5.25}
\end{equation*}
$$

because $f$ is compactly supported in $U$. It follows from the Fubini Theorem that

$$
\begin{equation*}
\int_{U} \frac{\partial f_{i}}{\partial x_{i}}=0 \tag{5.26}
\end{equation*}
$$

Now we prove the other direction, that (1) implies (2). Before our proof we make some remarks about functions of one variable.

Suppose $I=(a, b) \subseteq \mathbb{R}$, and let $g \in \mathcal{C}_{0}^{\infty}(I)$ such that supp $g \subseteq[c, d]$, where $a<c<d<b$. Also assume that

$$
\begin{equation*}
\int_{a}^{b} g(s) d s=0 \tag{5.27}
\end{equation*}
$$

Define

$$
\begin{equation*}
h(x)=\int_{a}^{x} g(s) d s \tag{5.28}
\end{equation*}
$$

where $a \leq x \leq b$.
Claim. The function $h$ is also supported on $c, d$.
Proof. If $x>d$, then we can write

$$
\begin{equation*}
h(x)=\int_{a}^{b} g(s) d s-\int_{x}^{b} g(s) d s \tag{5.29}
\end{equation*}
$$

where the first integral is zero by assumption, and the second integral is zero because the integrand is zero.

Now we begin our proof that (1) implies (2).
Let $\omega \in \Omega^{n}(U)$, where $U=Q$, and assume that

$$
\begin{equation*}
\int_{U} \omega=0 \tag{5.30}
\end{equation*}
$$

We will use the following inductive lemma:

Lemma 5.8. For all $0 \leq k \leq n+1$, there exists $\mu \in \Omega_{c}^{n-1}(U)$ and $f \in \mathcal{C}_{0}^{\infty}(U)$ such that

$$
\begin{equation*}
\omega=d \mu+f d x_{1} \wedge \cdots \wedge d x_{n} \tag{5.31}
\end{equation*}
$$

and

$$
\begin{equation*}
\int f\left(x_{1}, \ldots, x_{n}\right) d x_{k} \ldots d x_{n}=0 \tag{5.32}
\end{equation*}
$$

Note that the hypothesis for $k=0$ and $\mu=0$ says that $\int \omega=0$, which is our assumption (1). Also note that the hypothesis for $k=n+1, f=0$, and $\omega=d \mu$ is the same as the statement (2). So, if we can show that (the lemma is true for $k$ ) implies (the lemma is true for $k+1$ ), then we will have shown that (1) implies (2) in Poincare's Lemma. We now show this.

Assume that the lemma is true for $k$. That is, we have

$$
\begin{equation*}
\omega=d \mu+f d x_{1} \wedge \cdots \wedge d x_{n} \tag{5.33}
\end{equation*}
$$

and

$$
\begin{equation*}
\int f\left(x_{1}, \ldots, x_{n}\right) d x_{k} \ldots d x_{n}=0 \tag{5.34}
\end{equation*}
$$

where $\mu \in \Omega_{c}^{n-1}(U)$, and $f \in \mathcal{C}_{0}^{\infty}(\mathbb{R})$.
We can assume that $\mu$ and $f$ are supported on $\operatorname{Int} Q^{\prime}$, where $Q^{\prime} \subseteq \operatorname{Int} Q$ and $Q^{\prime}=\left[c_{1}, d_{1}\right] \times \cdots \times\left[c_{n}, d_{n}\right]$.

Define

$$
\begin{equation*}
g\left(x_{1}, \ldots, x_{k}\right)=\int f\left(x_{1}, \ldots, x_{n}\right) d_{k+1} \ldots d x_{n} \tag{5.35}
\end{equation*}
$$

Note that $g$ is supported on the interior of $\left[c_{1}, d_{1}\right] \times \cdots \times\left[c_{k}, d_{k}\right]$. Also note that

$$
\begin{equation*}
\int_{a_{k}}^{b_{k}} g\left(x_{1}, \ldots, x_{k-1}, s\right) d s=\int f\left(x_{1}, \ldots, x_{n}\right) d x_{k} \ldots d x_{n}=0 \tag{5.36}
\end{equation*}
$$

by our assumption that the lemma holds true for $k$.
Now, define

$$
\begin{equation*}
h\left(x_{1}, \ldots, x_{k}\right)=\int_{a_{k}}^{x_{k}} g\left(x_{1}, \ldots, x_{k-1}, s\right) d s \tag{5.37}
\end{equation*}
$$

From our earlier remark about functions of one variable, $h$ is supported on $c_{k} \leq x_{k} \leq$ $d_{k}$. Also, note that $h$ is supported on $c_{i} \leq x_{i} \leq d_{i}$, for $1 \leq i \leq k-1$. We conclude therefore that $h$ is also supported on $\left[c_{1}, d_{1}\right] \times \cdots \times\left[c_{k}, d_{k}\right]$.

Both $g$ and its "anti-derivative" are supported.

$$
\begin{equation*}
\frac{\partial h}{\partial x_{k}}=g . \tag{5.38}
\end{equation*}
$$

Let $\ell=n-k$, and consider $\rho=\rho\left(x_{k+1}, \ldots, x_{n}\right) \in \mathcal{C}_{0}^{\infty}\left(\mathbb{R}^{\ell}\right)$. Assume that $\rho$ is supported on the rectangle $\left[c_{k+1}, d_{k+1}\right] \times \cdots \times\left[c_{n}, d_{n}\right]$ and that

$$
\begin{equation*}
\int \rho d x_{k+1} \ldots d x_{n}=1 \tag{5.39}
\end{equation*}
$$

We can always find such a function, so we just fix one such function.
Define

$$
\begin{equation*}
\nu=(-1)^{k} h\left(x_{1}, \ldots, x_{k}\right) \rho\left(x_{k+1}, \ldots, x_{n}\right) d x_{1} \wedge \cdots \wedge \widehat{d x_{k}} \wedge \cdots \wedge d x_{n} \tag{5.40}
\end{equation*}
$$

The form $\nu$ is supported on $Q^{\prime}=\left[c_{1}, d_{1}\right] \times \cdots \times\left[c_{n}, d_{n}\right]$.
Now we compute $d \nu$,

$$
\begin{equation*}
d \nu=(-1)^{k} \sum_{j} \frac{\partial}{\partial x_{j}}(h \rho) d x_{j} \wedge d x_{1} \wedge \cdots \wedge \widehat{d x_{k}} \wedge \cdots \wedge d x_{n} \tag{5.41}
\end{equation*}
$$

Note that if $j \neq k$, then the summand is zero, so

$$
\begin{align*}
d \nu & =(-1)^{k} \frac{\partial h}{\partial x_{k}} \rho d x_{k} \wedge d x_{1} \wedge \cdots \wedge \widehat{d x_{k}} \wedge \cdots \wedge d x_{n} \\
& =(-1) \frac{\partial h}{\partial x_{k}} \rho d x_{1} \wedge \cdots \wedge d x_{n}  \tag{5.42}\\
& =-g \rho d x_{1} \wedge \cdots \wedge d x_{n}
\end{align*}
$$

Now, define

$$
\begin{equation*}
\mu_{\text {new }}=\mu-\nu \tag{5.43}
\end{equation*}
$$

and

$$
\begin{align*}
& \quad f_{\text {new }}=f\left(x_{1}, \ldots, x_{n}\right)-g\left(x_{1}, \ldots, x_{k}\right) \rho\left(x_{k+1}, \ldots, x_{n}\right)  \tag{5.44}\\
& \omega=d \mu_{\text {new }}+f_{\text {new }} d x_{1} \wedge \cdots \wedge d x_{n} \\
& =d \mu+\left(g\left(x_{1}, \ldots, x_{k}\right) \rho\left(x_{k+1}, \ldots, x_{n}\right)-f\left(x_{1}, \ldots, x_{k}\right)-g \rho\right) d x_{1} \wedge \cdots \wedge d x_{n}  \tag{5.45}\\
& =d \mu+f d x_{1} \wedge \cdots \wedge d x_{n} \\
& =\omega
\end{align*}
$$

Note that

$$
\begin{align*}
\int f_{\text {new }}= & \int f_{\text {new }}\left(x_{1}, \ldots, x_{n}\right) d x_{k+1} \ldots d x_{n} \\
= & \int f\left(x_{1}, \ldots, x_{n}\right) d x_{k+1} \ldots d x_{n}  \tag{5.46}\\
& -g\left(x_{1}, \ldots, x_{k}\right) \int \rho\left(x_{k+1}, \ldots, x_{n}\right) d x_{k+1} \ldots d x_{n} \\
= & g\left(x_{1}, \ldots, x_{k}\right)-g\left(x_{1}, \ldots, x_{k}\right)=0
\end{align*}
$$

which implies that the lemma is true for $k+1$.
Remark. In the above proof, we implicitly assumed that if $f \in \mathcal{C}_{0}^{\infty}\left(\mathbb{R}^{n}\right)$, then

$$
\begin{equation*}
g\left(x_{1}, \ldots, x_{k}\right)=\int f\left(x_{1}, \ldots, x_{n}\right) d x_{k+1} \ldots d x_{m} \tag{5.47}
\end{equation*}
$$

is in $\mathcal{C}_{0}^{\infty}\left(\mathbb{R}^{k}\right)$. We checked the support, but we did not check that $g$ is in $\mathcal{C}^{\infty}\left(\mathbb{R}^{k}\right)$. The proof of this is in the Supplementary Notes.

## Lecture 26

We continue our study of forms with compact support. Let us begin with a review. Let $U \in \mathbb{R}^{n}$ be open, and let

$$
\begin{equation*}
\omega=\sum_{I} f_{I}\left(x_{1}, \ldots, x_{n}\right) d x_{I} \tag{5.48}
\end{equation*}
$$

where $I=\left(i_{1}, \ldots, i_{k}\right)$ is strictly increasing and $d x_{I}=d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}$. Then
$\omega$ is compactly supported $\Longleftrightarrow$ every $f_{I}$ is compactly supported.
By definition,

$$
\begin{equation*}
\operatorname{supp} f_{I}=\overline{\left\{x \in U: f_{I}(x) \neq 0\right\}} \tag{5.50}
\end{equation*}
$$

We assume that the $f_{I}$ 's are $\mathcal{C}^{2}$ maps.

## Notation.

$$
\begin{equation*}
\Omega_{c}^{k}(U)=\text { space of compactly supported differentiable } k \text {-forms on } U \text {. } \tag{5.51}
\end{equation*}
$$

Now, let $\omega \in \Omega_{c}^{n}(U)$ defined by

$$
\begin{equation*}
\omega=f\left(x_{1}, \ldots, x_{n}\right) d x_{1} \wedge \cdots \wedge d x_{n} \tag{5.52}
\end{equation*}
$$

where $f \in \Omega_{c}^{0}(U)$. Then

$$
\begin{equation*}
\int_{\mathbb{R}^{n}} \omega=\int_{\mathbb{R}^{n}} f\left(x_{1}, \ldots, x_{n}\right) d x_{1} \wedge \cdots \wedge d x_{n} \tag{5.53}
\end{equation*}
$$

Last time we proved the Poincare Lemma for open rectangles $R$ in $\mathbb{R}^{n}$. We assumed that $\omega \in \Omega_{c}^{n}(\operatorname{Int} R)$. That is, we assumed that $\omega \in \Omega_{c}^{n}\left(\mathbb{R}^{n}\right)$ such that supp $\omega \subset \operatorname{Int} R$. We showed that for such $\omega$ the following two conditions are equivalent:

1. $\int_{\mathbb{R}^{n}} \omega=0$,
2. There exists a $\mu \in \Omega_{c}^{n-1}(\operatorname{Int} R)$ such that $d \mu=0$.

Definition 5.9. Whenever $\omega \in \Omega^{k}(U)$ and $\omega=d \mu$ for some $\mu \in \Omega^{k-1}(U)$, we say that $\omega$ is exact.

Definition 5.10. Whenever $\omega \in \Omega^{k}(U)$ such that $d \omega=0$, we say that $\omega$ is closed.
Observe that

$$
\begin{equation*}
\omega \in \Omega_{c}^{n}(U) \Longrightarrow d \omega=0 \tag{5.54}
\end{equation*}
$$

Now we prove the Poincare Lemma for open connected subsets of $\mathbb{R}^{n}$.

Poincare Lemma. Let $U$ be a connected open subset of $\mathbb{R}^{n}$, and let $\omega \in \Omega_{c}^{n}(U)$. The following conditions are equivalent:

1. $\int_{U} \omega=0$,
2. $\omega=d \mu$, for some $\mu \in \Omega_{c}^{n-1}(U)$.

Proof. We prove this more general case by reducing the proof to the case where $U$ is a rectangle, which we proved in the previous lecture.

First we prove that (2) implies (1). We can choose a family of rectangles $\left\{R_{i}, i \in\right.$ $\mathbb{N}\}$ such that

$$
\begin{equation*}
U=\bigcup_{i \in \mathbb{N}} \operatorname{Int} R_{i} \tag{5.55}
\end{equation*}
$$

Since the support of $\mu$ is compact, the set supp $\mu$ is covered by finitely many of the rectangles.

We take a partition of unity $\left\{\phi_{i}, i \in \mathbb{N}\right\}$ subordinate to $\left\{R_{i}\right\}$, so that

$$
\begin{equation*}
\mu=\sum_{i=1}^{N} \underbrace{\phi_{i} \mu}) \text { supported on Int } R_{i} \tag{5.56}
\end{equation*}
$$

Then

$$
\begin{equation*}
\int d \mu=\sum_{i} \int_{\mathbb{R}^{n}} d\left(\phi_{i} \mu\right) \tag{5.57}
\end{equation*}
$$

Each term on the r.h.s is zero by the Poincare Lemma we proved last lecture.
We now prove the other direction, that (1) implies (2). It is equivalent to show that if $\omega_{1}, \omega_{2} \in \Omega_{c}^{n}(U)$ such that

$$
\begin{equation*}
\int \omega_{1}=\int \omega_{2} \tag{5.58}
\end{equation*}
$$

then $\omega_{1} \sim \omega_{2}$, meaning that there exists a form $\mu \in \Omega_{c}^{n-1}(U)$ such that $\omega_{1}=\omega_{2}+d \mu$.
Choose a partition of unity $\left\{\phi_{i}\right\}$ as before. Then

$$
\begin{equation*}
\omega=\sum_{i=1}^{M} \underbrace{\phi_{i} \omega}_{\text {supported on Int } R_{i}} \tag{5.59}
\end{equation*}
$$

Let

$$
\begin{equation*}
\int \omega=c \in \mathbb{R} \tag{5.60}
\end{equation*}
$$

and

$$
\begin{equation*}
\int \phi_{i} \omega=c_{i} . \tag{5.61}
\end{equation*}
$$

Choose a form $\omega_{0}$ such that

$$
\begin{equation*}
\int \omega_{0}=1 \tag{5.62}
\end{equation*}
$$

and such that supp $\omega_{0} \subseteq Q_{0}=R_{j}$ for some $j$. Then

$$
\begin{equation*}
\int \underbrace{\phi_{i} \omega}_{\text {supported in } R_{i}}=\int \underbrace{c_{i} \omega_{0}}_{\text {supported in } Q_{0}} \tag{5.63}
\end{equation*}
$$

We want to show that there exists $\mu_{i} \in \Omega_{c}^{n-1}(U)$ such that $\phi_{i} \omega=c_{i} \omega_{i}+d \mu_{i}$.
Now we use the fact that $U$ is connected. We use the following lemma.
Lemma 5.11. Let $U$ be connected. Given rectangles $R_{i}$ such that supp $\phi_{i} \omega \subset \operatorname{Int} R_{i}$, and given a fixed rectangle $Q_{0}$ and any point $x \in U$, there exists a finite sequence of rectangles $R_{0}, \ldots, R_{N}$ with the following properties: $Q_{0}=R_{0}, x \in \operatorname{Int} R_{N}$, and (Int $\left.R_{i}\right) \cap\left(\right.$ Int $\left.R_{i+1}\right)$ is non-empty.

We omit the proof of this lemma.
Now, define $\omega_{i}=\phi_{i} \omega$, so

$$
\begin{equation*}
\int \omega_{i}=\int c_{i} \omega_{0} \tag{5.64}
\end{equation*}
$$

Note that

$$
\begin{align*}
\operatorname{supp}\left(c_{i} \omega_{0}\right) & \subseteq \operatorname{Int}\left(Q_{0}\right)  \tag{5.65}\\
\operatorname{supp}\left(\omega_{i}\right) & \subseteq \operatorname{Int}\left(R_{i}\right) . \tag{5.66}
\end{align*}
$$

Choose forms $\nu_{i}$ such that supp $\nu_{i} \subseteq \operatorname{Int} R_{i} \cap \operatorname{Int} R_{i+1}$ and such that

$$
\begin{equation*}
\int \nu_{i}=1 \tag{5.67}
\end{equation*}
$$

This implies that

$$
\begin{equation*}
\operatorname{supp}\left(\nu_{i}-\nu_{i+1}\right) \subseteq \operatorname{Int} R_{i+1} \tag{5.68}
\end{equation*}
$$

By definition,

$$
\begin{equation*}
\int\left(\nu_{i}-\nu_{i+1}\right)=0 \tag{5.69}
\end{equation*}
$$

By the Poincare Lemma we proved last time, $\nu_{i} \sim \nu_{i+1}$, so there exists $\mu_{i} \in \Omega_{c}^{n-1}(U)$ such that $\nu_{i}=\nu_{i+1}+d \mu_{i}$.

So,

$$
\begin{equation*}
c_{i} \omega_{0} \sim c_{i} \nu_{0} \sim c_{i} \nu_{1} \sim \ldots \sim c_{i} \nu_{N} \sim \phi_{i} \omega \tag{5.70}
\end{equation*}
$$

### 5.2 Proper Maps and Degree

We introduce a class of functions that remain compactly supported under the pullback operation.

Definition 5.12. Let $U \subseteq \mathbb{R}^{n}$ and $V \subseteq \mathbb{R}^{k}$, and let $f: U \rightarrow V$ be a continuous map. The map $f$ is proper if for all compact subsets $K \subseteq V$, the set $f^{-1}(K)$ is compact.

Let $U \subseteq \mathbb{R}^{n}$ and $V \subseteq \mathbb{R}^{k}$, and let $f: U \rightarrow V$ be a continuous map. Also let $\omega \in \Omega^{k}(V)$. The map

$$
\begin{equation*}
f^{*}: \Omega^{k}(V) \rightarrow \Omega^{k}(U) \tag{5.71}
\end{equation*}
$$

is defined such that

$$
\begin{equation*}
\omega=g\left(y_{1}, \ldots, y_{n}\right) d y_{i_{1}} \wedge \ldots \wedge d y_{i_{k}} \rightarrow f^{*} \omega=g(f(x)) d f_{i_{1}} \wedge \cdots d f_{i_{k}} \tag{5.72}
\end{equation*}
$$

So,

$$
\begin{equation*}
f^{-1}(\operatorname{supp} \omega) \supseteq \operatorname{supp}\left(f^{*} \omega\right) . \tag{5.73}
\end{equation*}
$$

If $f$ is proper and $\omega \in \Omega_{c}^{n}(V)$, then $\operatorname{supp}\left(f^{*} \omega\right)$ is compact, in which case the map $f^{*}$ is actually of the form

$$
\begin{equation*}
f^{*}: \Omega_{c}^{k}(V) \rightarrow \Omega_{c}^{k}(U) \tag{5.74}
\end{equation*}
$$

That is, $\omega \in \Omega_{c}^{n}(V) \rightarrow f^{*} \omega \in \Omega_{c}^{n}(U)$. So, it makes sense to take the integral

$$
\begin{equation*}
\int_{U} f^{*} \omega=(\operatorname{deg} f) \int_{V} \omega \tag{5.75}
\end{equation*}
$$

Theorem 5.13. Let $U, V$ be connected open subsets of $\mathbb{R}^{n}$, and let $f: U \rightarrow V$ be $a$ $\mathcal{C}^{\infty}$ map. For all $\omega \in \Omega_{c}^{n}(V)$,

$$
\begin{equation*}
\int_{U} f^{*} \omega=(\operatorname{deg} f) \int_{V} \omega \tag{5.76}
\end{equation*}
$$

Proof. Take $\omega_{0} \in \Omega_{c}^{n}(V)$ such that

$$
\begin{equation*}
\int \omega_{0}=1 \tag{5.77}
\end{equation*}
$$

Define

$$
\begin{equation*}
\operatorname{deg} f \equiv \int f^{*} \omega_{0} \tag{5.78}
\end{equation*}
$$

and suppose that

$$
\begin{equation*}
\int \omega=c . \tag{5.79}
\end{equation*}
$$

Then

$$
\begin{equation*}
\int \omega=\int c \omega_{0} . \tag{5.80}
\end{equation*}
$$

By the Poincare Lemma, $\omega \sim c \omega_{0}$. That is, there exists $\mu \in \Omega_{c}^{n-1}(V)$ such that $\omega=c \omega_{0}+d \mu$. Then

$$
\begin{align*}
f^{*} \omega & =f^{*}\left(c \omega_{0}\right)+f^{*}(d \mu) \\
& =f^{*}\left(c \omega_{0}\right)+d\left(f^{*} \mu\right), \tag{5.81}
\end{align*}
$$

which shows that $f^{*} \omega \sim f^{*}\left(c \omega_{0}\right)$. Putting this altogether,

$$
\begin{align*}
\int f^{*} \omega & =\int f^{*}\left(c \omega_{0}\right) \\
& =c \int f^{*} \omega_{0}  \tag{5.82}\\
& =c \operatorname{deg} f \\
& =\left(\int \omega\right) \operatorname{deg} f
\end{align*}
$$

We had $\omega=g\left(y_{1}, \ldots, y_{n}\right) d y_{1} \wedge \cdots \wedge d y_{n}$, so

$$
\begin{align*}
f^{*} \omega & =g(f(x)) d f_{1} \wedge \cdots \wedge d f_{m} \\
& =g(f(x)) \operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}\right] d x_{1} \wedge \cdots \wedge d x_{n} \tag{5.83}
\end{align*}
$$

where we used the fact that

$$
\begin{equation*}
d f_{i}=\sum_{j=1}^{n} \frac{\partial f_{i}}{\partial x_{j}} d x_{j} \tag{5.84}
\end{equation*}
$$

Restated in coordinates, the above theorem says that

$$
\begin{align*}
\int_{U} g(f(x)) \operatorname{det}(D f) d x_{1} \wedge \cdots \wedge d x_{n} & \\
& =(\operatorname{deg} f) \int_{V} g\left(y_{1}, \ldots, y_{n}\right) d y_{1} \wedge \cdots \wedge d y_{n} \tag{5.85}
\end{align*}
$$

Claim. Given proper maps $f: V \rightarrow W$ and $g: U \rightarrow V$, where $U, V, W$ are connected open subsets of $\mathbb{R}^{n}$,

$$
\begin{equation*}
\operatorname{deg}(f g)=(\operatorname{deg} g)(\operatorname{deg} f) \tag{5.86}
\end{equation*}
$$

Proof. Note that $(f \circ g)^{*}=g^{*} \circ f^{*}$, so

$$
\begin{align*}
\operatorname{deg}(f \circ g) \int_{W} \omega & =\int_{U}(f \circ g)^{*} \omega \\
& =\int_{V} g^{*}\left(f^{*} \omega\right) \\
& =(\operatorname{deg} g) \int_{V} f^{*} \omega  \tag{5.87}\\
& =(\operatorname{deg} g)(\operatorname{deg} f) \int_{W} \omega .
\end{align*}
$$

## Lecture 27

We proved the following Poincare Lemma:
Poincare Lemma. Let $U$ be a connected open subset of $\mathbb{R}^{n}$, and let $\omega \in \Omega_{c}^{n}(U)$. The following conditions are equivalent:

1. $\int_{U} \omega=0$,
2. $\omega=d \mu$, for some $\mu \in \Omega_{c}^{n-1}(U)$.

We first proved this for the case $U=\operatorname{Int} Q$, where $Q$ was a rectangle. Then we used this result to generalize to arbitrary open connected sets. We discussed a nice application: proper maps and degree.

Let $U, V$ be open subsets of $\mathbb{R}^{n}$, and let $f: U \rightarrow V$ be a $\mathcal{C}^{\infty}$ map. The map $f$ is proper if for every compact set $C \subseteq V$, the pre-image $f^{-1}(C)$ is also compact. Hence, if $f$ is proper, then

$$
\begin{equation*}
f^{*} \Omega_{c}^{k}(V) \subseteq \Omega_{c}^{k}(U) \tag{5.88}
\end{equation*}
$$

That is, if $\omega \in \Omega_{c}^{k}(V)$, then $f^{*} \omega \in \Omega_{c}^{k}(U)$, for all $k$.
When $k=n$,

$$
\begin{equation*}
\omega \in \Omega_{c}^{n}(V) \tag{5.89}
\end{equation*}
$$

In which case, we compare

$$
\begin{equation*}
\int_{v} \omega \text { and } \int_{U} f^{*} \omega \tag{5.90}
\end{equation*}
$$

Using the Poincare Lemma, we obtain the following theorem.
Theorem 5.14. There exists a constant $\gamma_{f}$ with the property that for all $\omega \in \Omega_{c}^{n}(V)$,

$$
\begin{equation*}
\int_{U} f^{*} \omega=\gamma_{f} \int_{V} \omega \tag{5.91}
\end{equation*}
$$

We call this constant the degree of $f$,

## Definition 5.15.

$$
\begin{equation*}
\operatorname{deg}(f)=\gamma_{f} \tag{5.92}
\end{equation*}
$$

Let $U, V, W$ be open connected subsets of $\mathbb{R}^{n}$, and let $f: U \rightarrow V$ and $g: V \rightarrow W$ be proper $\mathcal{C}^{\infty}$ maps. Then the map $g \circ f: U \rightarrow W$ is proper, and

$$
\begin{equation*}
\operatorname{deg}(g \circ f)=\operatorname{deg}(f) \operatorname{deg}(g) . \tag{5.93}
\end{equation*}
$$

Proof Hint: For all $\omega \in \Omega_{c}^{n}(W),(g \circ f)^{*} \omega=f^{*}\left(g^{*} \omega\right)$.

We give some examples of the degree of various maps. Let $f=T_{a}$, the transposition by $a$. That is, let $f(x)=x+a$. From \#4 in section 4 of the Supplementary Notes, the map $T_{a}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is proper. One can show that $\operatorname{deg}\left(T_{a}\right)=1$.

As another example, let $f=A: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ be a bijective liner map. Then

$$
\operatorname{deg} A= \begin{cases}1 & \text { if } \operatorname{det} A>0  \tag{5.94}\\ -1 & \text { if } \operatorname{det} A<0\end{cases}
$$

We now study the degree as it pertains to orientation preserving and orientation reversing maps.

Let $U, V$ be connected open sets in $\mathbb{R}^{n}$, and let $f: U \rightarrow V$ be a diffeomorphism. Take $p \in U$. Then $D f(p): \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is one-to-one and onto. The map $f$ is orientation preserving if $\operatorname{det} D f(p)>0$ for all $p \in U$, and the map $f$ is orientation reversing if $\operatorname{deg} D f(p)<0$ for all $p \in U$.

Theorem 5.16. If $f$ is orientation preserving, then $\operatorname{deg}(f)=1$; if $f$ is orientation reversing, then $\operatorname{deg}(f)=-1$.

Proof. Let $a \in U$ and $b=f(a)$. Define

$$
\begin{equation*}
f_{\text {old }}=f \tag{5.95}
\end{equation*}
$$

and define

$$
\begin{equation*}
f_{\text {new }}=T_{-b} \circ f_{\text {old }} \circ T_{a}, \tag{5.96}
\end{equation*}
$$

where $T_{-b}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ and $T_{a}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ are transpositions by $-b$ and $a$, respectively.
By the formula $\operatorname{deg}(g \circ f)=\operatorname{deg}(f) \operatorname{deg}(g)$,

$$
\begin{align*}
\operatorname{deg}\left(f_{\text {new }}\right) & =\operatorname{deg}\left(T_{-b}\right) \operatorname{deg}\left(f_{\text {old }}\right) \operatorname{deg}\left(T_{a}\right)  \tag{5.97}\\
& =\operatorname{deg}(\text { old }) .
\end{align*}
$$

By replacing $f$ with $f_{\text {new }}$, we can assume that $0 \in U$ and $f(0)=0$.
We can make yet another simplification, that $D f(0)=I$, the identity. To see this, let $D f(0)=A$, where $A: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$. Taking our new $f$, we redefine $f_{\text {old }}=f$, and we redefine $f_{\text {new }}=A^{-1} \circ f_{\text {old }}$. Then,

$$
\begin{equation*}
\operatorname{deg}\left(f_{\text {new }}\right)=\operatorname{deg}(A) \operatorname{deg} \operatorname{deg}\left(f_{\text {old }}\right) \tag{5.98}
\end{equation*}
$$

where

$$
\begin{align*}
\operatorname{deg} A & =\operatorname{deg}\left(D f_{\text {old }}\right) \\
& = \begin{cases}1 & \text { if } D f_{\text {old }} \text { is orient. preserving } \\
-1 & \text { if } D f_{\text {old }} \text { is orient. reversing. }\end{cases} \tag{5.99}
\end{align*}
$$

We again replace $f$ with $f_{\text {new }}$. It suffices to prove the theorem for this new $f$. To summarize, we can assume that

$$
\begin{equation*}
0 \in U, \quad f(0)=0 . \quad \text { and } \quad D f(0)=I \tag{5.100}
\end{equation*}
$$

Consider $g(x)=x-f(x)$ (so $f(x)=x-g(x))$. Note that $(D g)(0)=I-I=0$. If we write $g=\left(g_{1}, \ldots, g_{n}\right)$, then

$$
\begin{equation*}
\left[\frac{\partial g_{i}}{\partial x_{j}}(0)\right]=0 . \tag{5.101}
\end{equation*}
$$

So, each $\frac{\partial g_{i}}{\partial x_{j}}(0)=0$.
Lemma 5.17. There exists $\delta>0$ such that for all $|x|<\delta$,

$$
\begin{equation*}
|g(x)| \leq \frac{|x|}{2} \tag{5.102}
\end{equation*}
$$

Proof. So far, we know that $g(0)=0-f(0)=0$, and $\frac{\partial g_{i}}{\partial x_{j}}(0)=0$. By continuity, there exists $\delta>0$ such that

$$
\begin{equation*}
\left|\frac{\partial g_{i}}{\partial x_{j}}(x)\right| \leq \frac{1}{2 n} \tag{5.103}
\end{equation*}
$$

for all $|x|<\delta$. Using the Mean-value Theorem, for all $|x|<\delta$,

$$
\begin{align*}
g_{i}(x) & =g_{i}(x)-g_{i}(0) \\
& =\sum \frac{\partial g_{i}}{\partial x_{j}}(c) x_{j} \tag{5.104}
\end{align*}
$$

where $c=t_{0} x$ for some $0<t_{0}<1$. So,

$$
\begin{align*}
\left|g_{i}(x)\right| & \leq \sum_{i=1}^{n} \frac{1}{2 n}\left|x_{i}\right| \\
& \leq \frac{1}{2} \max \left\{\left|x_{i}\right|\right\}  \tag{5.105}\\
& =\frac{1}{2}|x|
\end{align*}
$$

Define $\tilde{f}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ as follows. Let $\rho \in \mathcal{C}_{0}^{\infty}\left(\mathbb{R}^{n}\right)$, defined to have the follow properties

$$
\rho(x)= \begin{cases}1 & \text { if }|x|<\delta / 2  \tag{5.106}\\ 0 & \text { if }|x|>\delta \\ 0 \leq \rho(x) \leq 1 & \text { otherwise }\end{cases}
$$

Remember that $f(x)=x-g(x)$. Define

$$
\tilde{f}= \begin{cases}x-\rho(x) g(x) & \text { if }|x|<\delta  \tag{5.107}\\ x & \text { if }|x|>\delta\end{cases}
$$

Claim. The map $\tilde{f}$ has the following properties:

1. $\tilde{f}=f(x)$ for all $|x|<\frac{\delta}{2}$,
2. $\tilde{f}=x$ for all $|x|>\delta$,
3. $|\tilde{f}(x)| \geq \frac{|x|}{2}$,
4. $|\tilde{f}(x)| \leq 2|x|$.

Proof. We only proof properties (3) and (4). First we prove property (3). We have $\tilde{f}(x)=x-\rho(x) g(x)=x$ when $|x| \geq \delta$, so $|\tilde{f}(x)|=|x|$ when $|x| \geq \delta$. For $|x|<\delta$, we have

$$
\begin{align*}
|\tilde{f}(x)| & \geq|x|-\rho(x)|g(x)| \\
& =|x|-|g(x)| \\
& \geq|x|-\frac{|x|}{2}  \tag{5.108}\\
& =\frac{|x|}{2} .
\end{align*}
$$

We now prove property (4). We have $\tilde{f}(x)=x-\rho(x) g(x)$, so $|\tilde{f}(x)|=|x|$ for $x \geq \delta$. For $x<\delta$, we have

$$
\begin{align*}
|\tilde{f}(x)| & \leq|x|+\rho(x)|g(x)| \\
& \leq|x|+\frac{1}{2}|x|  \tag{5.109}\\
& \leq 2|x| .
\end{align*}
$$

Let $Q_{r} \equiv\left\{x \in \mathbb{R}^{n}:|x| \leq r\right\}$. The student should check that

$$
\begin{equation*}
\text { Property }(3) \Longrightarrow \tilde{f}^{-1}\left(Q_{r}\right) \subseteq Q_{2 r} \tag{5.110}
\end{equation*}
$$

and that

$$
\begin{equation*}
\operatorname{Property}(4) \Longrightarrow \tilde{f}^{-1}\left(\mathbb{R}^{n}-Q_{2 r}\right) \subseteq \mathbb{R}^{n}-Q_{r} \tag{5.111}
\end{equation*}
$$

Notice that $\tilde{f}^{-1}\left(Q_{r}\right) \subseteq Q_{2 r} \Longrightarrow \tilde{f}$ is proper.
Now we turn back to the map $f$. Remember that $f: U \rightarrow V$ is a diffeomorphism and that $f(0)=0$. So, the set $f\left(\operatorname{Int} Q_{\delta / 2}\right)$ is an open neighborhood of 0 in $\mathbb{R}^{n}$. Take

$$
\begin{equation*}
\omega \in \Omega_{c}^{n}\left(f\left(\operatorname{Int} Q_{\delta / 2}\right) \cap \operatorname{Int} Q_{\delta / 4}\right) \tag{5.112}
\end{equation*}
$$

such that

$$
\begin{equation*}
\int_{\mathbb{R}^{n}} \omega=1 \tag{5.113}
\end{equation*}
$$

Then,

$$
\begin{equation*}
f^{*} \omega \in \Omega_{c}^{n}\left(Q_{\delta / 2}\right) \tag{5.114}
\end{equation*}
$$

and

$$
\begin{equation*}
\tilde{f}^{*} \omega \in \Omega_{c}^{n}\left(Q_{\delta / 2}\right), \tag{5.115}
\end{equation*}
$$

by Equation 5.110 . This shows that $f^{*} \omega=\tilde{f}^{*} \omega$. Hence,

$$
\begin{align*}
\int_{U} f^{*} \omega & =\int_{U} \tilde{f}^{*} \omega=\operatorname{deg}(f) \int_{V} \omega  \tag{5.116}\\
& =\operatorname{deg}(\tilde{f}) \int_{V} \omega
\end{align*}
$$

where

$$
\begin{equation*}
\int_{V} \omega=1 \tag{5.117}
\end{equation*}
$$

Therefore,

$$
\begin{equation*}
\operatorname{deg}(f)=\operatorname{deg}(\tilde{f}) \tag{5.118}
\end{equation*}
$$

Now, let us use Equation 5.111. Choose $\omega \in \Omega_{c}^{n}\left(\mathbb{R}^{n}-Q_{2 \delta}\right)$. So,

$$
\begin{equation*}
f^{*} \omega \in \Omega_{c}^{n}\left(\mathbb{R}^{n}-Q_{\delta}\right) \tag{5.119}
\end{equation*}
$$

Again we take

$$
\begin{equation*}
\int_{\mathbb{R}^{n}} \omega=1 \tag{5.120}
\end{equation*}
$$

By property (2), $\tilde{f}=I$ on $\mathbb{R}^{n}-Q_{\delta}$, so

$$
\begin{equation*}
\tilde{f}^{*} \omega=\omega \tag{5.121}
\end{equation*}
$$

Integrating,

$$
\begin{equation*}
\int_{\mathbb{R}}^{n} \tilde{f}^{*} \omega=\operatorname{deg}(\tilde{f}) \int_{\mathbb{R}}^{n} \omega=\int_{\mathbb{R}}^{n} \omega . \tag{5.122}
\end{equation*}
$$

Therefore,

$$
\begin{equation*}
\operatorname{deg}(f)=\operatorname{deg}(\tilde{f})=1 \tag{5.123}
\end{equation*}
$$

## Lecture 28

Let $U, V$ be connected open sets of $\mathbb{R}^{n}$, and let $f: U \rightarrow V$ be a diffeomorphism. Then

$$
\operatorname{deg}(f)= \begin{cases}+1 & \text { if } f \text { is orient. preserving }  \tag{5.124}\\ -1 & \text { if } f \text { is orient. reversing. }\end{cases}
$$

We showed that given any $\omega \in \Omega_{c}^{n}(V)$,

$$
\begin{equation*}
\int_{U} f^{*} \omega= \pm \int_{V} \omega \tag{5.125}
\end{equation*}
$$

Let $\omega=\phi(x) d x_{1} \wedge \cdots \wedge d x_{n}$, where $\phi \in \mathcal{C}_{0}^{\infty}(V)$. Then

$$
\begin{equation*}
f^{*} \omega=\phi(f(x)) \operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}(x)\right] d x_{1} \wedge \cdots \wedge d x_{n} \tag{5.126}
\end{equation*}
$$

so,

$$
\begin{equation*}
\int_{U} \phi(f(x)) \operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}\right] d x= \pm \int_{V} \phi(x) d x \tag{5.127}
\end{equation*}
$$

Notice that

$$
\begin{align*}
f \text { is orientation preserving } & \Longleftrightarrow \operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}(x)\right]>0  \tag{5.128}\\
f \text { is orientation reversing } & \Longleftrightarrow \operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}(x)\right]<0 \tag{5.129}
\end{align*}
$$

So, in general,

$$
\begin{equation*}
\int_{U} \phi(f(x))\left|\operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}(x)\right]\right| d x \tag{5.130}
\end{equation*}
$$

As usual, we assumed that $f \in \mathcal{C}^{\infty}$.
Remark. The above is true for $\phi \in \mathcal{C}_{0}^{1}$, a compactly supported continuous function. The proof of this is in section 5 of the Supplementary Notes. The theorem is true even if only $f \in \mathcal{C}^{1}$ (the notes prove it for $f \in \mathcal{C}^{2}$ ).

Today we show how to compute the degree in general.
Let $U, V$ be connected open sets in $\mathbb{R}^{n}$, and let $f: U \rightarrow V$ be a proper $\mathcal{C}^{\infty}$ map.
Claim. Let $B$ be a compact subset of $V$, and let $A=f^{-1}(B)$. If $U_{0}$ is an open subset of $U$ with $A \subseteq U_{0}$, then there exists an open subset $V_{0}$ of $V$ with $B \subseteq V_{0}$ such that $f^{-1}\left(V_{0}\right) \subseteq U_{0}$.

Proof. Let $C \subseteq V$ be a compact set with $B \subseteq \operatorname{Int} C$, and let $W=f^{-1}(C)-U_{0}$. The set $W$ is compact, so the set $f(W)$ is also compact. Moreover, $f(W) \cap B=\phi$ since $f^{-1}(B) \subseteq U_{0}$.

Now, let $V_{0}=\operatorname{Int} C-f(W)$. This set is open, and

$$
\begin{align*}
f^{-1}\left(V_{0}\right) & \subseteq f^{-1}(\operatorname{Int} C)-W  \tag{5.131}\\
& \subseteq U_{0}
\end{align*}
$$

Claim. If $X \subseteq U$ is closed, then $f(X)$ is closed in $V$.
Proof. Take any point $p \in V-f(x)$. Then $f^{-1}(p) \in U-X$. Apply the previous result with $B=\{p\}, A=f^{-1}(p)$, and $U_{0}=U-X$. There exists an open set $V_{0} \ni p$ such that $f^{-1} \subseteq U-X$. The set $V_{0} \cap f(X)=\phi$, so $V-f(X)$ is open in $V$.

We now remind you of Sard's Theorem. Let $f: U \rightarrow V$ be a proper $\mathcal{C}^{\infty}$ map. We define the critical set

$$
\begin{equation*}
C_{f}=\{p \in U: D f(p) \text { is not bijective }\} . \tag{5.132}
\end{equation*}
$$

The set $C_{f}$ is closed. The set $f\left(C_{f}\right)$ in $V$ is a set of measure zero. The set $f\left(C_{f}\right)$ is closed as well, since $f$ is proper.

Definition 5.18. A point $q \in V$ is a regular value of $f$ if $q \in V-f\left(C_{f}\right)$.
Sard's Theorem basically says that there are "lots" of regular values.
Lemma 5.19. If $q$ is a regular value, then $f^{-1}(q)$ is a finite set.
Proof. First, $p \in f^{-1}(q) \Longrightarrow p \notin C_{f}$. So, $D f(p): \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is bijective. By the IFT, the map $f$ is a diffeomorphism of a neighborhood $U_{p}$ of $p \in U$ onto a neighborhood of $q$. In particular, since $f$ is one-to-one and onto,

$$
\begin{equation*}
U_{p} \cap f^{-1}(q)=\{p\} \tag{5.133}
\end{equation*}
$$

Consider the collection $\left\{U_{p}: p \in f^{-1}(q)\right\}$, which is an open cover of $f^{-1}(q)$. The H-B Theorem tells us that there exists a finite subcover $\left\{U_{p_{i}}, i=1, \ldots, N\right\}$. Hence,

$$
\begin{equation*}
f^{-1}(q)=\left\{p_{1}, \ldots, p_{N}\right\} . \tag{5.134}
\end{equation*}
$$

Theorem 5.20. The degree of $f$ is

$$
\begin{equation*}
\operatorname{deg}(f)=\sum_{i=1}^{N} \sigma_{p_{i}}, \tag{5.135}
\end{equation*}
$$

where

$$
\sigma_{p_{i}}= \begin{cases}+1 & \text { if } D f\left(p_{i}\right) \text { is orient. preserving }  \tag{5.136}\\ -1 & \text { if } D f\left(p_{i}\right) \text { is orient. reversing. }\end{cases}
$$

So, to calculate the degree, you just pick any regular value $q$ and "count" the number of points in the pre-image of $q$, keeping track of the value of $\sigma_{p_{i}}$.

Proof. For each $p_{i} \in f^{-1}(q)$, let $U_{p_{i}}$ be an open neighborhood of $p_{i}$ such that $f$ maps $U_{p_{i}}$ diffeomorphically onto a neighborhood of $q$. We can assume that the $U_{p_{i}}$ 's do not intersect.

Now, choose a neighborhood $V_{0}$ of $q$ such that

$$
\begin{equation*}
f^{-1}\left(V_{0}\right) \subseteq \bigcup U_{p_{i}} \tag{5.137}
\end{equation*}
$$

Next, replace each $U_{p_{i}}$ by $U_{p_{i}} \cap f^{-1}\left(V_{0}\right)$. So, we can assume the following:

1. $f$ is a diffeomorphism of $U_{p_{i}}$ onto $V_{0}$,
2. $f^{-1}\left(V_{0}\right)=\bigcup U_{p_{i}}$,
3. The $U_{p_{i}}$ 's don't intersect.

Choose $\omega \in \Omega_{c}^{n}\left(V_{0}\right)$ such that

$$
\begin{equation*}
\int_{V} \omega=1 . \tag{5.138}
\end{equation*}
$$

Then,

$$
\begin{align*}
\int_{U} f^{*} \omega & =\sum_{i} \int_{U_{p_{i}}} f^{*} \omega \\
& =\sum_{i} \sigma_{p_{i}} \int_{V_{0}} \omega  \tag{5.139}\\
& =\sum_{i} \sigma_{p_{i}} .
\end{align*}
$$

But,

$$
\begin{equation*}
\int_{U} f^{*} \omega=(\operatorname{deg} f) \int_{U} \omega=\operatorname{deg} f \tag{5.140}
\end{equation*}
$$

so

$$
\begin{equation*}
\sum \sigma_{p_{i}}=\operatorname{deg} f \tag{5.141}
\end{equation*}
$$

This is a very nice theorem that is not often discussed in textbooks.
The following is a useful application of this theorem. Suppose $f^{-1}(q)$ is empty, so $q \notin f(U)$. Then $q \notin f\left(C_{f}\right)$, so $q$ is a regular value. Therefore,

$$
\begin{equation*}
\operatorname{deg}(f)=0 \tag{5.142}
\end{equation*}
$$

This implies the following useful theorem.

Theorem 5.21. If $\operatorname{deg}(f) \neq 0$, then $f: U \rightarrow V$ is onto.
This theorem can let us know if a system of non-linear equations has a solution, simply by calculating the degree. The way to think about this is as follows. Let $f=\left(f_{1}, \ldots, f_{n}\right)$ and let $q=\left(c_{1}, \ldots, c_{n}\right) \in V$. If $q \in f(U)$ then there exists a solution $x \in U$ to the system of non-linear equations

$$
\begin{equation*}
f_{i}(x)=c_{i}, \quad i=1, \ldots, n \tag{5.143}
\end{equation*}
$$

## Lecture 29

We have been studying the important invariant called the degree of $f$. Today we show that the degree is a "topological invariant."

### 5.3 Topological Invariance of Degree

Recall that given a subset $A$ of $\mathbb{R}^{m}$ and a function $F: A \rightarrow \mathbb{R}^{\ell}$, we say that $F$ is $\mathcal{C}^{\infty}$ if it extends to a $\mathcal{C}^{\infty}$ map on a neighborhood of $A$.

Let $U$ be open in $\mathbb{R}^{n}$, let $V$ be open in $\mathbb{R}^{k}$, and let $A=U \times[0,1]$.
Definition 5.22. Let $f_{0}, f_{1}: U \rightarrow V$ be $\mathcal{C}^{\infty}$ maps. The maps $f_{0}$ and $f_{1}$ are homotopic if there is a $\mathcal{C}^{\infty} \operatorname{map} F: U \times[0,1] \rightarrow V$ such that $F(p, 0)=f_{0}(p)$ and $F(p, 1)=f_{1}(p)$ for all $p \in U$.

Let $f_{t}: U \rightarrow V$ be the map defined by

$$
\begin{equation*}
f_{t}(p)=F(p, t) . \tag{5.144}
\end{equation*}
$$

Note that $F \in \mathcal{C}^{\infty} \Longrightarrow f_{t} \in \mathcal{C}^{\infty}$. So, $f_{t}: U \rightarrow V$, where $0 \leq t \leq 1$, gives a family of maps parameterized by $t$. The family of maps $f_{t}$ is called a $\mathcal{C}^{\infty}$ deformation of $f_{0}$ into $f_{1}$.

Definition 5.23. The map $F$ is a proper homotopy if for all compact sets $A \subseteq V$, the pre-image $F^{-1}(A)$ is compact.

Denote by $\pi$ the map $\pi: U \times[0,1] \rightarrow U$ that sends $(p, t) \rightarrow t$. Let $A \subseteq V$ be compact. Then $B=\pi\left(F^{-1}(A)\right)$ is compact, and for all $t, f_{t}^{-1}(A) \subseteq B$. As a consequence, each $f_{t}$ is proper.

We concentrate on the case where $U, V$ are open connected subsets of $\mathbb{R}^{n}$ and $f_{0}, f_{1}: U \rightarrow V$ are proper $\mathcal{C}^{\infty}$ maps. We now prove that the degree is a topological invariant.

Theorem 5.24. If $f_{0}$ and $f_{1}$ are homotopic by a proper homotopy, then

$$
\begin{equation*}
\operatorname{deg}\left(f_{0}\right)=\operatorname{deg}\left(f_{1}\right) \tag{5.145}
\end{equation*}
$$

Proof. Let $\omega \in \Omega_{c}^{n}(V)$ and let supp $\omega=A$. Let $F: U \times I \rightarrow V$ be a proper homotopy between $f_{0}$ and $f_{1}$. Take $B=\pi\left(F^{-1}(A)\right)$, which is compact. For all $t \in[0,1]$, $f_{t}^{-1}(A) \subseteq B$.

Let us compute $f_{t}^{*} \omega$. We can write $\omega=\phi(x) d x_{1} \wedge \cdots \wedge d x_{n}$, where $\operatorname{supp} \phi \subseteq A$. So,

$$
\begin{equation*}
f_{t}^{*} \omega=\phi(F(x, t)) \operatorname{det}\left[\frac{\partial F_{i}}{\partial x_{j}}(x, t)\right] d x_{1} \wedge \cdots \wedge d x_{n} \tag{5.146}
\end{equation*}
$$

and

$$
\begin{align*}
\int_{U} f_{t}^{*} \omega & =\operatorname{deg}\left(f_{t}\right) \int_{V} \omega \\
& =\int_{U} \phi(F(x, t)) \operatorname{det}\left[\frac{\partial F_{i}}{\partial x_{j}}(x, t)\right] d x_{1} \ldots d x_{n} \tag{5.147}
\end{align*}
$$

Notice that the integrand is supported in the compact set $B$ for all $t$, and it is $\mathcal{C}^{\infty}$ as a function of $x$ and $t$. By Exercise $\# 2$ in section 2 of the Supplementary Notes, this implies that the integral is $\mathcal{C}^{\infty}$ in $t$. From Equation 5.147, we can conclude that $\operatorname{deg}\left(f_{t}\right)$ is a $\mathcal{C}^{\infty}$ function of $t$.

Now here is the trick. Last lecture we showed that $\operatorname{deg}\left(f_{t}\right)$ is an integer. Since $\operatorname{deg}\left(f_{t}\right)$ is continuous, it must be a constant $\operatorname{deg}\left(f_{t}\right)=$ constant.

We consider a simple application of the above theorem. Let $U=V=\mathbb{R}^{2}$, and think of $\mathbb{R}^{2}=\mathbb{C}$. We make the following associations:

$$
\begin{align*}
i^{2} & =-1  \tag{5.148}\\
z & =x+i y  \tag{5.149}\\
\bar{z} & =x-i y  \tag{5.150}\\
z \bar{z} & =|z|^{2}=x^{2}+y^{2}  \tag{5.151}\\
d z & =d x+i d y  \tag{5.152}\\
d \bar{z} & =d x-i d y  \tag{5.153}\\
d z \wedge d \bar{z} & =-2 i d x \wedge d y  \tag{5.154}\\
d x \wedge d y & =\frac{1}{2} i d z \wedge d \bar{z} . \tag{5.155}
\end{align*}
$$

Consider a map $f: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$, thinking of $\mathbb{R}^{2}=\mathbb{C}$, defined by

$$
\begin{equation*}
f(z)=z^{n}+\sum_{i=0}^{n-1} c_{i} z^{i}, c_{i} \in \mathbb{C} \tag{5.156}
\end{equation*}
$$

Claim. The map $f$ is proper.
Proof. Let $C=\sum\left|c_{i}\right|$. For $|z|>1$,

$$
\begin{equation*}
\left|\sum_{i=0}^{n-1} c_{i} z^{i}\right| \leq C|z|^{n-1} \tag{5.157}
\end{equation*}
$$

So,

$$
\begin{align*}
|f(z)| & \geq|z|^{n}-\left|\sum c_{i} z^{i}\right| \\
& =|z|^{n}-C|z|^{n-1}  \tag{5.158}\\
& =|z|^{n}\left(1-\frac{C}{|z|}\right) .
\end{align*}
$$

For $|z|>2 C$,

$$
\begin{equation*}
|f(z)| \geq \frac{|z|^{n}}{2} \tag{5.159}
\end{equation*}
$$

So, if $R>1$ and $R>2 C$, then $f^{-1}\left(B_{R}\right) \subseteq B_{R_{1}}$, where $R_{1}^{n} / 2 \leq R$ (and where $B_{r}$ denotes the ball of radius $r$ ). So $f$ is proper.

Now, let us define a homotopy $F: \mathbb{C} \times[0,1] \rightarrow \mathbb{C}$ by

$$
\begin{equation*}
F(z, t)=z^{n}+t \sum_{i=0}^{n-1} c_{i} z^{i} \tag{5.160}
\end{equation*}
$$

We claim that $F^{-1}\left(B_{R}\right) \subseteq B_{R_{1}} \times[0,1]$, by exactly the same argument as above. So $F$ is proper.

Notice that

$$
\begin{align*}
& F(z, 1)=f_{1}(z)=f(z)  \tag{5.161}\\
& F(z, 0)=f_{0}(z)=z^{n} \tag{5.162}
\end{align*}
$$

So, by the above theorem, $\operatorname{deg}(f)=\operatorname{deg}\left(f_{0}\right)$.
Let us compute $\operatorname{deg}\left(f_{0}\right)$ by brute force. We have $f_{0}(z)=z^{n}$, so

$$
\begin{align*}
& f_{0}^{*} d z=d z^{n}=n z^{n-1} d z  \tag{5.163}\\
& f_{0}^{*} d \bar{z}=d \bar{z}^{n}=n \bar{z}^{n-1} d \bar{z} \tag{5.164}
\end{align*}
$$

Using the associations defined above,

$$
\begin{align*}
f_{0}^{*}(d x \wedge d y) & =\frac{i}{2} f_{0}^{*}(d z \wedge d \bar{z}) \\
& =\frac{i}{2} f_{0}^{*} d z \wedge f_{0}^{*} d \bar{z}  \tag{5.165}\\
& =\frac{i}{2} n^{2}|z|^{2(n-1)} d z \wedge d \bar{z} \\
& +n^{2}|z|^{2 n-2} d x \wedge d y
\end{align*}
$$

Let $\phi \in \mathcal{C}_{0}^{\infty}(\mathbb{R})$ such that

$$
\begin{equation*}
\int_{0}^{\infty} \phi(s) d s=1 \tag{5.166}
\end{equation*}
$$

Let $\omega=\phi\left(|z|^{2}\right) d x \wedge d y$. We calculate $\int_{\mathbb{R}^{2}} \omega$. Let us use polar coordinates, where

$$
r=\sqrt{x^{2}+y^{2}}=|z|
$$

$$
\begin{align*}
\int_{\mathbb{R}^{2}} \omega & =\int_{\mathbb{R}^{2}} \phi\left(|z|^{2}\right) d x d y \\
& =\int_{\mathbb{R}^{2}} \phi\left(r^{2}\right) r d r d \theta \\
& =2 \pi \int_{o}^{\infty} \phi\left(r^{2}\right) r d r  \tag{5.167}\\
& +2 \pi \int_{0}^{\infty} \phi(s) \frac{d s}{2} \\
& =\pi
\end{align*}
$$

Now we calculate $\int f_{0}^{*} \omega$. First, we note that

$$
\begin{equation*}
f_{0}^{*} \omega=\phi\left(|z|^{2 n}\right) n^{2}|z|^{2 n-2} d x \wedge d y \tag{5.168}
\end{equation*}
$$

So,

$$
\begin{align*}
\int f_{0}^{*} \omega & =n^{2} \int_{0}^{\infty} \phi\left(r^{2 n}\right) r^{2 n-2} r d r d \theta \\
& =n^{2}(2 \pi) \int_{0}^{\infty} \phi\left(r^{2 n}\right) r^{2 n-1} d r  \tag{5.169}\\
& =n^{2}(2 \pi) \int_{0}^{\infty} \phi(s) \frac{d s}{2 n} \\
& =n \pi
\end{align*}
$$

To summarize, we have calculated that

$$
\begin{equation*}
\int_{\mathbb{R}^{2}} \omega=\pi \quad \text { and } \quad \int_{\mathbb{R}^{2}} f_{0}^{*} \omega=n \pi . \tag{5.170}
\end{equation*}
$$

Therefore,

$$
\begin{equation*}
\operatorname{deg}\left(f_{0}\right)=\operatorname{deg}(f)=n \tag{5.171}
\end{equation*}
$$

A better way to do the above calculation is in the homework: problem $\# 6$ of section 6 of the Supplementary Notes.

Last lecture we showed that if $\operatorname{deg}(f) \neq 0$, then the map $f$ is onto. Applying this to the above example, we find that the algebraic equation

$$
\begin{equation*}
z^{n}+\sum_{i=0}^{n-1} c_{i} z^{i}=0 \tag{5.172}
\end{equation*}
$$

has a solution. This is known as the Fundamental Theorem of Algebra.

## Lecture 30

## 6 Manifolds

### 6.1 Canonical Submersion and Canonical Immersion Theorems

As part of today's homework, you are to prove the canonical submersion and immersion theorems for linear maps. We begin today's lecture by stating these two theorems.

Let $A: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ be a linear map, and let $\left[a_{i j}\right]$ be its associated matrix. We have the transpose map $A^{t}: \mathbb{R}^{m} \rightarrow \mathbb{R}^{n}$ with the associated matrix $\left[a_{j i}\right]$.

Definition 6.1. Let $k<n$. Define the canonical submersion map $\pi$ and the canonical immersion map $\iota$ as follows:

Canonical submersion:

$$
\begin{equation*}
\pi: \mathbb{R}^{n} \rightarrow \mathbb{R}^{k}, \quad\left(x_{1}, \ldots, x_{n}\right) \rightarrow\left(x_{1}, \ldots, x_{k}\right) \tag{6.1}
\end{equation*}
$$

Canonical immersion:

$$
\begin{equation*}
\iota: \mathbb{R}^{k} \rightarrow \mathbb{R}^{n}, \quad\left(x_{1}, \ldots, x_{k}\right) \rightarrow\left(x_{1}, \ldots, x_{k}, 0, \ldots, 0\right) \tag{6.2}
\end{equation*}
$$

Canonical Submersion Thoerem. Let $A: \mathbb{R}^{n} \rightarrow \mathbb{R}^{k}$ be a linear map, and suppose that $A$ is onto. Then there exists a bijective linear map $B: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ such that $A \circ B=\pi$.

Proof Hint: Show that there exists a basis $v_{1}, \ldots, v_{n}$ of $\mathbb{R}^{n}$ such that $A v_{i}=e_{i}, i=$ $1, \ldots, k$, (the standard basis of $\mathbb{R}^{k}$ ) and $A v_{i}=0$ for all $i>k$. Then let $B: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ be the linear map $B e_{i}=v_{i}, i=1, \ldots, n$, where $e_{i}, \ldots, e_{n}$ is the standard basis of $\mathbb{R}^{n}$.

Canonical Immersion Thoerem. As before, let $k<n$. Let $A: \mathbb{R}^{k} \rightarrow \mathbb{R}^{n}$ be $a$ one-to-one linear map. Then there exists a bijective linear map $B: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ such that $B \circ A=\iota$.

Proof Hint: Note that $B \circ A=\iota \Longleftrightarrow A^{t} B^{t}=\pi$. Use the Canonical Submersion Theorem.

Now we prove non-linear versions of these two theorems.
Let $U$ be an open set in $\mathbb{R}^{n}$, and let $f: U \rightarrow \mathbb{R}^{k}$ be a $\mathcal{C}^{\infty}$ map. Let $p \in U$.
Definition 6.2. The map $f$ is a submersion at $p$ if $D f(p): \mathbb{R}^{n} \rightarrow \mathbb{R}^{k}$ is onto.

Canonical Submersion Thoerem. Assume that $f$ is a submersion at $p$ and that $f(p)=0$. Then there exists a neighborhood $U_{0}$ of $p$ in $U$, a neighborhood $V$ of 0 in $\mathbb{R}^{n}$, and a diffeomorphism $g: V \rightarrow U_{0}$ such that $f \circ g=\pi$.

Proof. Let $T_{p}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ be the translation defined by $x \rightarrow x+p$. Replacing $f$ by $f \circ T_{p}$ we can assume that $p=0$ and $f(0)=0$.

Let $A=(D f)(0)$, where $A: \mathbb{R}^{n} \rightarrow \mathbb{R}^{k}$ is onto by the assumption that $f$ is a submersion. So, there exists a bijective linear map $B: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ such that $A \circ B=\pi$. Replacing $f$ by $f \circ B$ we can assume that $D f(0)=\pi$.

Define a map $h: U \rightarrow \mathbb{R}^{n}$ by

$$
\begin{equation*}
h\left(x_{1}, \ldots, x_{n}\right)=\left(f\left(x_{1}, \ldots, x_{k}\right) ; x_{k+1}, \ldots, x_{n}\right) . \tag{6.3}
\end{equation*}
$$

Note that (1) $D h(0)=I$; and (2) $\pi h=f$. By (1), the function $h$ maps a neighborhood $U_{0}$ of 0 in $U$ diffeomorphically onto a neighborhood $V$ of 0 in $\mathbb{R}^{n}$. By (2), we have $\pi=f \circ h^{-1}$. Take $g=h^{-1}$.

There is a companion theorem having to do with immersions.
Definition 6.3. Let $U$ be an open subset of $\mathbb{R}^{k}$, and let $f: U \rightarrow \mathbb{R}^{n}$ be a $\mathcal{C}^{\infty}$ map. Let $p \in U$. The map $f$ is an immersion at $p$ if $(D f)(p): \mathbb{R}^{k} \rightarrow \mathbb{R}^{n}$ is injective (one-to-one).

Canonical Immersion Thoerem. Let $U$ be a neighborhood of 0 in $\mathbb{R}^{k}$, and let $f: U \rightarrow \mathbb{R}^{n}$ be a $\mathcal{C}^{\infty}$ map. Assume that $f$ is an immersion at 0 . Then there exists $a$ neighborhood $V$ of $f(0)=p$ in $\mathbb{R}^{n}$, a neighborhood $W$ of 0 in $\mathbb{R}^{k}$, and a diffeomorphism $g: V \rightarrow W$ such that $\iota^{-1}(W) \subseteq U$ and $g \circ f=\iota$.

Proof. Replacing $f$ by $T_{p} \circ f$, we can assume that $f(0)=0$. Let $A=D f(0)$, so $A: \mathbb{R}^{k} \rightarrow \mathbb{R}^{n}$ is injective. There exists a linear map $B: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ such that $B A=\iota$. Replacing $f$ by $B \circ f$, we can assume that $D f(0)=\iota$.

Let $\ell=n-k$. Since $U \subseteq \mathbb{R}^{k}$, we get $U \times \mathbb{R}^{\ell} \subseteq \mathbb{R}^{k} \times \mathbb{R}^{\ell}=\mathbb{R}^{n}$. Define a map $h: U \times \mathbb{R}^{\ell} \rightarrow \mathbb{R}^{n}$ by

$$
\begin{equation*}
h\left(x_{1}, \ldots, x_{n}\right)=f\left(x_{1}, \ldots, x_{k}\right)+\left(0, \ldots, 0, x_{k+1}, \ldots, x_{n}\right) . \tag{6.4}
\end{equation*}
$$

One can check that (1) Dh(0) $=I$; and (2) $h \circ \iota=f$.
By (1), the function $h$ maps a neighborhood $W$ of 0 in $U \times \mathbb{R}^{\ell}$ diffeomorphically onto a neighborhood $V$ of 0 in $\mathbb{R}^{n}$. Moreover, $W \subseteq U \times \mathbb{R}^{\ell}$, so $\iota^{-1}(W) \subseteq U$.

By (2), we obtain the canonical immersion map $\iota=h^{-1} \circ f$. Take $g=h^{-1}$.

### 6.2 Definition of Manifold

Now we move on to the study of manifolds.
Let $X$ be a subset of $\mathbb{R}^{n}$, let $Y$ be a subset of $\mathbb{R}^{m}$, and let $f: X \rightarrow Y$ be a continuous map. We define that the map $f$ is a $\mathcal{C}^{\infty}$ map if for every point $p \in X$ there exists a neighborhood $U_{p}$ of $p$ in $\mathbb{R}^{n}$ and a $\mathcal{C}^{\infty}$ map $g_{p}: U_{p} \rightarrow \mathbb{R}^{n}$ such that $g_{p} \mid X \cap U_{p}=f$.

We showed in the homework that if $f: X \rightarrow Y$ is a $\mathcal{C}^{\infty}$ map, then there exists a neighborhood $U$ of $X$ in $\mathbb{R}^{n}$ and a $\mathcal{C}^{\infty}$ map $g: U \rightarrow \mathbb{R}^{n}$ extending $f$.

Definition 6.4. A map $f: X \rightarrow Y$ is a diffeomorphism if it is one-to-one, onto, a $\mathcal{C}^{\infty}$ map, and $f^{-1}: Y \rightarrow X$ is $\mathcal{C}^{\infty}$.

Let $X$ be a subset of $\mathbb{R}^{N}$.
Definition 6.5. The set $X$ is an $n$-dimensional manifold if for every point $p \in X$ there exists a neighborhood $V$ of $p$ in $\mathbb{R}^{N}$, an open set $U$ in $\mathbb{R}^{m}$, and a diffeomorphism $f: U \rightarrow V \cap X$. The collection $(f, U, X)$ is called a parameterization of $X$ at $p$.

This definition does not illustrate how manifolds come up in nature. Usually manifolds come up in the following scenario.

Let $W$ be open in $\mathbb{R}^{N}$, and let $f_{i}: W \rightarrow \mathbb{R}, i=1, \ldots, \ell$ be $\mathcal{C}^{\infty}$ functions. Suppose you want to study the solution space of

$$
\begin{equation*}
f_{i}\left(x_{1}, \ldots, x_{N}\right)=0, \quad i=1, \ldots, \ell \tag{6.5}
\end{equation*}
$$

Then you consider the mapping $f: W \rightarrow \mathbb{R}^{\ell}$ defined by

$$
\begin{equation*}
f(x)=\left(f_{1}(x), \ldots, f_{\ell}(x)\right) \tag{6.6}
\end{equation*}
$$

Claim. If for every $p \in W$ the map $f$ is a submersion of $p$, then Equation 6.6 defines a $k$-dimensional manifold, where $k=N-\ell$.

## Lecture 31

### 6.3 Examples of Manifolds

We begin with a review of the definition of a manifold.
Let $X$ be a subset of $\mathbb{R}^{n}$, let $Y$ be a subset of $\mathbb{R}^{m}$, and let $f: X \rightarrow Y$ be a continuous map.

Definition 6.6. The map $f$ is $\mathcal{C}^{\infty}$ if for every $p \in X$, there exists a neighborhood $U_{p}$ of $p$ in $\mathbb{R}^{n}$ and a $\mathcal{C}^{\infty}$ map $g_{p}: U_{p} \rightarrow \mathbb{R}^{m}$ such that $g_{p}=f$ on $U_{p} \cap X$.

Claim. If $f: X \rightarrow Y$ is continuous, then there exists a neighborhood $U$ of $X$ in $\mathbb{R}^{n}$ and a $\mathcal{C}^{\infty}$ map $g: U \rightarrow \mathbb{R}^{m}$ such that $g=f$ on $U \cap X$.

Definition 6.7. The map $f: X \rightarrow Y$ is a diffeomorphism if it is one-to-one, onto, and both $f$ and $f^{-1}$ are $\mathcal{C}^{\infty}$ maps.

We define the notion of a manifold.
Definition 6.8. A subset $X$ of $\mathbb{R}^{N}$ is an $n$-dimensional manifold if for every $p \in X$, there exists a neighborhood $V$ of $p$ in $\mathbb{R}^{N}$, an open set $U$ in $\mathbb{R}^{n}$, and a diffeomorphism $\phi: U \rightarrow X \cap V$.

Intuitively, the set $X$ is an $n$-dimensional manifold if locally near every point $p \in X$, the set $X$ "looks like an open subset of $\mathbb{R}^{n}$."

Manifolds come up in practical applications as follows:
Let $U$ be an open subset of $\mathbb{R}^{N}$, let $k<N$, and let $f: \mathbb{R}^{N} \rightarrow \mathbb{R}^{k}$ be a $\mathcal{C}^{\infty}$ map. Suppose that 0 is a regular value of $f$, that is, $f^{-1}(0) \cap C_{f}=\phi$.

Theorem 6.9. The set $X=f^{-1}(0)$ is an $n$-dimensional manifold, where $n=N-k$.
Proof. If $p \in f^{-1}(0)$, then $p \notin C_{f}$. So the map $D f(p): \mathbb{R}^{N} \rightarrow \mathbb{R}^{k}$ is onto. The map $f$ is a submersion at $p$.

By the canonical submersion theorem, there exists a neighborhood $V$ of 0 in $\mathbb{R}^{n}$, a neighborhood $U_{0}$ of $p$ in $U$, and a diffeomorphism $g: V \rightarrow U$ such that

$$
\begin{equation*}
f \circ g=\pi \tag{6.7}
\end{equation*}
$$

Recall that $\mathbb{R}^{N}=\mathbb{R}^{\ell} \times \mathbb{R}^{n}$ and $\pi: \mathbb{R}^{N} \rightarrow \mathbb{R}^{k}$ is the map that sends

$$
\begin{equation*}
(x, y) \in \mathbb{R}^{k} \times \mathbb{R}^{n} \rightarrow \mathbb{R}^{k} \tag{6.8}
\end{equation*}
$$

Hence, $\pi^{-1}(0)=\{0\} \times \mathbb{R}^{n}=\mathbb{R}^{n}$. By Equation 6.7, the function $g$ maps $V \cap \pi^{-1}(0)$ diffeomorphically onto $U_{0} \cap f^{-1}(0)$. But $V \cap \pi^{-1}(0)$ is a neighborhood of 0 in $\mathbb{R}^{n}$ and $U_{0} \cap f^{-1}(0)$ is a neighborhood of $p$ in $X$.

We give three examples of applications of the preceding theorem.

1. We consider the $n$-sphere $S^{n}$. Define a map

$$
\begin{equation*}
f: \mathbb{R}^{n+1} \rightarrow \mathbb{R}, \quad f(x)=x_{1}^{2}+\ldots+x_{n+1}^{2}-1 \tag{6.9}
\end{equation*}
$$

The derivative is $(D f)(x)=2\left[x_{1}, \ldots, x_{n+1}\right]$, so $C_{f}=\{0\}$. If $a \in f^{-1}(0)$, then $\sum a_{i}^{2}=1$, so $a \notin C_{f}$. Thus, the set $f^{-1}(0)=S^{n}$ is an $n$-dimensional manifold.
2. Let $g: \mathbb{R}^{n} \rightarrow \mathbb{R}^{k}$ be a $\mathcal{C}^{\infty}$ map. Define

$$
\begin{equation*}
X=\operatorname{graph} g=\left\{(x, y) \in \mathbb{R}^{n} \times \mathbb{R}^{k}: y=g(x)\right\} \tag{6.10}
\end{equation*}
$$

Note that $X \subseteq \mathbb{R}^{n} \times \mathbb{R}^{k}=\mathbb{R}^{n+k}$.
Claim. The set $X$ is an $n$-dimensional manifold.
Proof. Define a map $f: \mathbb{R}^{n} \times \mathbb{R}^{k} \rightarrow \mathbb{R}^{k}$ by

$$
\begin{equation*}
f(x, y)=y-g(x) \tag{6.11}
\end{equation*}
$$

Note that $D f(x, y)=\left[-D g(x), I_{k}\right]$. This is always of rank $k$, so $C_{f}=\phi$. Hence, the graph $g$ is an $n$-dimensional manifold.
3. The following example comes from Munkres section 24, exercise \#6. Let

$$
\begin{equation*}
\mathcal{M}_{n}=\text { the set of all } n \times n \text { matrices, } \tag{6.12}
\end{equation*}
$$

so

$$
\begin{equation*}
\mathcal{M}_{n} \cong \mathbb{R}^{n^{2}} \tag{6.13}
\end{equation*}
$$

With any element $\left[a_{i j}\right]$ in $\mathcal{M}_{n}$ we associate a vector

$$
\begin{equation*}
\left(a_{11}, \ldots, a_{1 n}, a_{21}, \ldots, a_{2 n}, \ldots\right) \tag{6.14}
\end{equation*}
$$

Now, let

$$
\begin{equation*}
\mathcal{S}_{n}=\left\{A \in \mathcal{M}_{n}: A=A^{t}\right\} \tag{6.15}
\end{equation*}
$$

so

$$
\begin{equation*}
\mathcal{S}_{n} \cong \mathbb{R}^{\frac{n(n+1)}{2}} \tag{6.16}
\end{equation*}
$$

With any element $\left[a_{i j}\right]$ in $\mathcal{S}_{n}$ we associate a vector

$$
\begin{equation*}
\left(a_{11}, \ldots, a_{1 n}, a_{22}, a_{23}, \ldots, a_{2 n}, a_{33}, a_{34}, \ldots\right) \tag{6.17}
\end{equation*}
$$

The above association avoids the "redundancies" $a_{12}=a_{21}, a_{31}=a_{13}, a_{32}=a_{23}$, etc.
Define

$$
\begin{equation*}
O(n)=\left\{A \in \mathcal{M}_{n}: A^{t} A=I\right\} \tag{6.18}
\end{equation*}
$$

which is the set of orthogonal $n \times n$ matrices.
As an exercise, the student should prove the following claim.

Claim. The set $O(n) \subseteq \mathcal{M}_{n}$ is an $\frac{n(n-1)}{2}$-dimensional manifold.
Proof Hint: First hint: Let $f: \mathcal{M}_{n} \rightarrow \mathcal{S}_{n}$ be the map defined by

$$
\begin{equation*}
f(A)=A^{t} A-I \tag{6.19}
\end{equation*}
$$

so $O(n)=f^{-1}(0)$. Show that $f^{-1}(0) \cap C_{f}=\phi$. The main idea is to show that if $A \notin f^{-1}(0)$, then the map $D f(A): \mathcal{M}_{n} \rightarrow \mathcal{S}_{n}$ is onto.
Second hint: Note that $D f(A)$ is the map the sends $B \in \mathcal{M}_{n}$ to $A^{t} B+B^{t} A$.
Manifolds are often defined by systems of non-linear equations:
Let $f: \mathbb{R}^{N} \rightarrow \mathbb{R}^{k}$ be a continuous map, and suppose that $C_{f} \cap f^{-1}(0)=\phi$. Then $X=f^{-1}(0)$ is an $n$-dimensional manifold. Suppose that $f=\left(f_{1}, \ldots, f_{k}\right)$. Then $X$ is defined by the system of equations

$$
\begin{equation*}
f_{i}\left(x_{1}, \ldots, x_{N}\right)=0, \quad i=1, \ldots, k . \tag{6.20}
\end{equation*}
$$

This system of equations is called non-degenerate, since for every $x \in X$ the matrix

$$
\begin{equation*}
\left[\frac{\partial f_{i}}{\partial x_{j}}(x)\right] \tag{6.21}
\end{equation*}
$$

is of rank $k$.
Claim. Every $n$-dimensional manifold $X \subseteq \mathbb{R}^{N}$ can be described locally by a system of $k$ non-degenerate equations of the type above.

Proof Idea: Let $X \subseteq \mathbb{R}^{N}$ be an $n$-dimensional manifold. Let $p \in X$, let $U$ be an open subset of $\mathbb{R}^{n}$, and let $V$ be an open neighborhood of $p$ in $\mathbb{R}^{N}$. Let $\phi: I \rightarrow V \cap X$ be a diffeomorphism. Modifying $\phi$ by a translation if necessary we can assume that $0 \in U$ and $\phi(0)=p$. We can think of $\phi$ as a map $\phi: U \rightarrow \mathbb{R}^{N}$ mapping $U$ into $X$.
Claim. The linear map $(D \phi)(0): \mathbb{R}^{n} \rightarrow \mathbb{R}^{N}$ is injective.
Proof. The map $\phi^{-1}: V \cap X \rightarrow U$ is a $\mathcal{C}^{\infty}$ map, so (shrinking $V$ if necessary) we can assume there is a $\mathcal{C}^{\infty} \operatorname{map} \psi: V \rightarrow U$ with $\psi=\phi^{-1}$ on $V \cap X$. Since $\phi$ maps $U$ onto $V \cap X$, we have $\psi \circ \phi=\phi^{-1} \circ \phi=I=$ the identity map of $U$ onto itself. Thus,

$$
\begin{equation*}
I=D(\psi \circ \phi)(0)=(D \psi)(p)(D \phi)(0) . \tag{6.22}
\end{equation*}
$$

That is, $D \psi(p)$ is a "left inverse" of $D \phi(0)$. So, $D \phi(0)$ is injective.

We can conclude that $\phi: U \rightarrow \mathbb{R}^{N}$ is an immersion at 0 . The canonical immersion theorem tells us that there exists a neighborhood $U_{0}$ of 0 in $U$, a neighborhood $V_{p}$ of $p$ in $V$, and a $\mathcal{C}^{\infty} \operatorname{map} g: V_{p} \rightarrow \mathbb{R}^{N}$ mapping $p$ onto 0 and mapping $V_{p}$ diffeomorphically onto a neighborhood $\mathcal{O}$ of 0 in $\mathbb{R}^{N}$ such that

$$
\begin{equation*}
\iota^{-1}(\mathcal{O})=U_{0} \tag{6.23}
\end{equation*}
$$

and

$$
\begin{equation*}
g \circ \phi=\iota \tag{6.24}
\end{equation*}
$$

on $U_{0}$. Here, the map $\iota$ is the canonical submersion map $\iota: \mathbb{R}^{n} \rightarrow \mathbb{R}^{N}$ that maps $\left(x_{1}, \ldots, x_{n}\right) \rightarrow\left(x_{1}, \ldots, x_{n}, 0, \ldots, 0\right)$.

By Equation 6.24, the function $g$ maps $\phi\left(U_{0}\right)$ onto $\iota\left(U_{0}\right)$. However, by Equation 6.23 , the set $\iota\left(U_{0}\right)$ is the subset of $\mathcal{O}$ defined by the equations

$$
\begin{equation*}
x_{i}=0, \quad i=n+1, \ldots, N . \tag{6.25}
\end{equation*}
$$

So, if $g=\left(g_{1}, \ldots, g_{N}\right)$, then $\phi\left(U_{0}\right)=X \cap V_{p}$ is defined by the equations

$$
\begin{equation*}
g_{i}=0, \quad i=n+1, \ldots, N . \tag{6.26}
\end{equation*}
$$

Moreover, the $N \times N$ matrix

$$
\begin{equation*}
\left[\frac{\partial g_{i}}{\partial x_{j}}(x)\right] \tag{6.27}
\end{equation*}
$$

is of rank $N$ at every point $x \in V_{p}$, since $g: V_{p} \rightarrow \mathcal{O}$ is a diffeomorphism. Hence, the last $N-n$ row vectors of this matrix

$$
\begin{equation*}
\left(\frac{\partial g_{i}}{\partial x_{1}}, \ldots, \frac{\partial g_{i}}{\partial x_{N}}\right), i=n+1, \ldots, N \tag{6.28}
\end{equation*}
$$

are linearly independent at every point $x \in V_{p}$.
Now let $k=N-n$ and let $f_{i}=g_{i+n}, i=1, \ldots, k$. Then $X \cap V_{p}$ is defined by the equations

$$
\begin{equation*}
f_{i}(x)=0, \quad i=1, \ldots, k \tag{6.29}
\end{equation*}
$$

and the $k \times N$ matrix

$$
\begin{equation*}
\left[\frac{\partial f_{i}}{\partial x_{k}}(x)\right] \tag{6.30}
\end{equation*}
$$

is of rank $k$ at all points $x \in V_{p}$. In other words, the system of equations 6.29 is non-degenerate.

## Lecture 32

### 6.4 Tangent Spaces of Manifolds

We generalize our earlier discussion of tangent spaces to tangent spaces of manifolds. First we review our earlier treatment of tangent spaces.

Let $p \in \mathbb{R}^{n}$. We define

$$
\begin{equation*}
T_{p} \mathbb{R}^{n}=\left\{(p, v): v \in \mathbb{R}^{n}\right\} \tag{6.31}
\end{equation*}
$$

Of course, we associate $T_{p} \mathbb{R}^{n} \cong \mathbb{R}^{n}$ by the map $(p, v) \rightarrow v$.
If $U$ is open in $\mathbb{R}^{n}, V$ is open in $\mathbb{R}^{k}$, and $f:(U, p) \rightarrow(V, q)$ (meaning that $f$ maps $U \rightarrow V$ and $\left.p \rightarrow p_{1}\right)$ is a $\mathcal{C}^{\infty}$ map, then we have the map $d f_{p}: T_{p} \mathbb{R}^{n} \rightarrow T_{q} \mathbb{R}^{k}$. Via the identifications $T_{p} \mathbb{R}^{n} \cong \mathbb{R}^{n}$ and $T_{p} \mathbb{R}^{k} \cong \mathbb{R}^{k}$, the map $d f_{p}$ is just the map $D f(p): \mathbb{R}^{n} \rightarrow \mathbb{R}^{k}$. Because these two maps can be identified, we can use the chain rule for $\mathcal{C}^{\infty}$ maps. Specifically, if $f:(U, p) \rightarrow(V, q)$ and $g:(V, q) \rightarrow\left(\mathbb{R}^{\ell}, w\right)$, then

$$
\begin{equation*}
d(g \circ f)_{p}=(d g)_{q} \circ(d f)_{p} \tag{6.32}
\end{equation*}
$$

because $(D g)(q)(D f(p))=(D g \circ f)(p)$.
You might be wondering: Why did we make everything more complicated by using $d f$ instead of $D f$ ? The answer is because we are going to generalize from Euclidean space to manifolds.

Remember, a set $X \subseteq \mathbb{R}^{N}$ is an $n$-dimensional manifold if for every $p \in X$, there exists a neighborhood $V$ of $p$ in $\mathbb{R}^{N}$, an open set $U$ in $\mathbb{R}^{n}$, and a diffeomorphism $\phi: U \rightarrow V \cap X$. The map $\phi: U \rightarrow V \cap X$ is called a parameterization of $X$ at $p$.

Let us think of $\phi$ as a map $\phi: U \rightarrow \mathbb{R}^{N}$ with $\operatorname{Im} \phi \subseteq X$.
Claim. Let $\phi^{-1}(p)=q$. Then the $\operatorname{map}(d \phi)_{q}: T_{q} \mathbb{R}^{n} \rightarrow T_{p} \mathbb{R}^{N}$ is one-to-one.
Reminder of proof: The map $\phi^{-1}: V \cap X \rightarrow U$ is a $\mathcal{C}^{\infty}$ map. So, shrinking $V$ if necessary, we can assume that this map extends to a map $\psi: V \rightarrow U$ such that $\psi=\phi^{-1}$ on $X \cap V$. Then note that for any $u \in U$, we have $\psi(\phi(u))=\phi^{-1}(\phi(u))=u$. So, $\psi \circ \phi=\mathrm{id}_{U}=$ the identity on $U$.

Using the chain rule, and letting $\phi(q)=p$, we get

$$
\begin{align*}
d(\psi \circ \phi)_{q} & =(d \psi)_{o} \circ(d \phi)_{q}  \tag{6.33}\\
& =\left(d\left(\mathrm{id}_{U}\right)\right)_{q} .
\end{align*}
$$

So, $(d \phi)_{q}$ is injective.
Today we define for any $p \in X$ the tangent space $T_{p} X$, which will be a vector subspace $T_{p} X \subseteq T_{p} \mathbb{R}^{N}$. The tangent space will be like in elementary calculus, that is, a space tangent to some surface.

Let $\phi: U \rightarrow V \cap X$ be a parameterization of $X$, and let $\phi(q)=p$. The above claim tells us that $(d \phi)_{q}: T_{q} \mathbb{R}^{n} \rightarrow T_{p} \mathbb{R}^{N}$ is injective.

Definition 6.10. We define the tangent space of a manifold $X$ to be

$$
\begin{equation*}
T_{p} X=\operatorname{Im}(d \phi)_{q} \tag{6.34}
\end{equation*}
$$

Because $(d \phi)_{q}$ is injective, the space $T_{p} X$ is $n$-dimensional.
We would like to show that the space $T_{p} X$ does not depend on the choice of parameterization $\phi$. To do so, we will make use of an equivalent definition for the tangent space $T_{p} X$.

Last time we showed that given $p \in X \subseteq \mathbb{R}^{N}$, and $k=N-n$, there exists a neighborhood $V$ of $p$ in $\mathbb{R}^{N}$ and a $\mathcal{C}^{\infty}$ map $f: V \rightarrow \mathbb{R}^{k}$ mapping $f(p)=0$ such that $X \cap V=f^{-1}(0)$. Note that $f^{-1}(0) \cap C_{f}=\phi$ (where here $\phi$ is the empty set).

We motivate the second definition of the tangent space. Since $p \in f^{-1}(0)$, the point $p \notin C_{f}$. So, the map $d f_{p}: T_{p} \mathbb{R}^{N} \rightarrow T_{0} \mathbb{R}^{k}$ is surjective. So, the kernel of $d f_{p}$ in $T_{p} \mathbb{R}^{N}$ is of dimension $N-k=n$.
Definition 6.11. An alternate definition for the tangent space of a manifold is

$$
\begin{equation*}
T_{p} X=\operatorname{ker} d f_{p} \tag{6.35}
\end{equation*}
$$

Claim. These two definitions for the tangent space $T_{p} X$ are equivalent.
Proof. Let $\phi: U \rightarrow V \cap X$ be a parameterization of $X$ at $p$ with $\phi(p)=q$. The function $f: V \rightarrow \mathbb{R}^{k}$ has the property that $f^{-1}(0)=X \cap V$. So, $f \circ \phi \equiv 0$. Applying the chain rule,

$$
\begin{equation*}
\left(d f_{p}\right) \circ\left(d \phi_{q}\right)=d(0)=0 . \tag{6.36}
\end{equation*}
$$

So, $\operatorname{Im} d \phi_{q}=\operatorname{ker} d f_{p}$.
We can now explain why the tangent space $T_{p} X$ is independent of the chosen parameterization. We have two definitions for the tangent space. The first does not depend on the choice of $\phi$, and the second does not depend on choice of $f$. Therefore, the tangent space depends on neither.
Lemma 6.12. Let $W$ be an open subset of $\mathbb{R}^{\ell}$, and let $g: W \rightarrow \mathbb{R}^{n}$ be a $\mathcal{C}^{\infty}$ map. Suppose that $g(W) \subseteq X$ and that $g(w)=p$, where $w \in W$. Then $(d g)_{W} \subseteq T_{p} X$.
Proof Hint: We leave the proof as an exercise. As above, we have a map $f: V \rightarrow \mathbb{R}^{k}$ such that $X \cap V=f^{-1}(0)$ and $T_{p} X=\operatorname{ker} d f_{p}$. Let $W_{1}=g^{-1}(V)$, and consider the map $f \circ g: W_{1} \rightarrow \mathbb{R}^{k}$. As before, $f \circ g=0$, so $d f_{p} \circ d g_{w}=0$.

Suppose that $X \subseteq \mathbb{R}^{N}$ is an $n$-dimensional manifold and $Y \subseteq \mathbb{R}^{\ell}$ is an $m$ dimensional manifold. Let $f: X \rightarrow Y$ be a $\mathcal{C}^{\infty}$ map, and let $f(p)=q$. We want to define a linear map

$$
\begin{equation*}
d f_{p}: T_{p} X \rightarrow T_{q} Y \tag{6.37}
\end{equation*}
$$

Let $v$ be a neighbor hood of $p$ in $\mathbb{R}^{N}$, and let $g: V \rightarrow \mathbb{R}^{\ell}$ be a map such that $g=f$ on $V \cap X$. By definition $T_{p} X \subseteq T_{p} \mathbb{R}^{N}$, so we have

$$
\begin{equation*}
d g_{p}: T_{p} \mathbb{R}^{N} \rightarrow T_{q} \mathbb{R}^{k} \tag{6.38}
\end{equation*}
$$

We define the map $d f_{p}$ to be the restriction of $d g_{p}$ to the tangent space $T_{p} X$.

## Definition 6.13.

$$
\begin{equation*}
d f_{p}=d g_{p} \mid T_{p} X \tag{6.39}
\end{equation*}
$$

There are two questions about this definition that should have us worried:

1. Is $\operatorname{Im} d g_{p}\left(T_{p} X\right)$ a subset of $T_{q} Y$ ?
2. Does this definition depend on the choice of $g$ ?

We address these two questions here:

1. Is $\operatorname{Im} d g_{p}\left(T_{p} X\right)$ a subset of $T_{q} Y$ ?

Let $U$ be an open subset of $\mathbb{R}^{N}$, let $q=f(p)$, and let $\phi: U \rightarrow X \cap V$ be a parameterization of $X$ at $p$. As before, let us think of $\phi$ as a map $\phi: U \rightarrow \mathbb{R}^{N}$ with $\phi(U) \subseteq X$.

By definition, $T_{p} X=\operatorname{Im}(d \phi)_{r}$, where $\phi(r)=p$. So, given $v \in T_{p} X$, one can always find $w \in T_{r} \mathbb{R}^{n}$ with $v=(d \phi)_{r} w$.
Now, is it true that $(d g)_{p}(v) \in T_{q} Y$ ? We have

$$
\begin{align*}
(d g)_{p} v & =(d g)_{p}(d \phi)_{r}(w)  \tag{6.40}\\
& =d(g \circ \phi)_{r}(w),
\end{align*}
$$

and the map $(g \circ \phi)$ is of the form $g \circ \phi: U \rightarrow Y$, so

$$
\begin{equation*}
d(g \circ \phi)_{r}(w) \in T_{q} Y \tag{6.41}
\end{equation*}
$$

2. Does the definition depend on the choice of $g$ ?

Consider two such maps $g_{1}, g_{2}: V \rightarrow \mathbb{R}^{\ell}$. The satisfy $g_{1}=g_{2}=f$ on $X \cap V$. Then, with $v, w$ as above,

$$
\begin{align*}
& \left(d g_{1}\right)_{p}(v)=d\left(g_{1} \circ \phi\right)_{r}(w)  \tag{6.42}\\
& \left(d g_{2}\right)_{p}(v)=d\left(g_{2} \circ \phi\right)_{r}(w) . \tag{6.43}
\end{align*}
$$

Since $g_{1}=g_{2}$ on $X \cap V$, we have

$$
\begin{equation*}
g_{1} \circ \phi=g_{2} \circ \phi=f \circ \phi \tag{6.44}
\end{equation*}
$$

Hence,

$$
\begin{equation*}
d\left(g_{1} \circ \phi\right)_{r}(w)=d\left(g_{2} \circ \phi\right)_{r}(w) \tag{6.45}
\end{equation*}
$$

As an exercise, show that the chain rule also generalizes to manifolds as follows: Suppose that $X_{1}, X_{2}, X_{3}$ are manifolds with $X_{i} \subseteq \mathbb{R}^{N_{i}}$, and let $f: X_{1} \rightarrow X_{2}$ and $g: X_{2} \rightarrow X_{3}$ be $\mathcal{C}^{\infty}$ maps. Let $f(p)=q$ and $g(q)=r$.

Show the following claim.

## Claim.

$$
\begin{equation*}
d(g \circ f)_{p}=\left(d g_{q}\right) \circ(d f)_{q} . \tag{6.46}
\end{equation*}
$$

Proof Hint: Let $V_{1}$ be a neighborhood of $p$ in $\mathbb{R}^{N_{1}}$, and let $V_{2}$ be a neighborhood of $q$ in $\mathbb{R}^{N_{2}}$. Let $\tilde{f}: V_{1} \rightarrow V_{2}$ be an extension of $f$ to $V_{1}$, and let $\tilde{g}: V_{2} \rightarrow \mathbb{R}^{N_{3}}$ be an extension of $g$ to $V_{2}$.

The chain rule for $f, g$ follows from the chain rule for $\tilde{f}, \tilde{g}$.

## Lecture 33

### 6.5 Differential Forms on Manifolds

Let $U \subseteq \mathbb{R}^{n}$ be open. By definition, a $k$-form $\omega$ on $U$ is a function which assigns to each point $p \in U$ an element $\omega_{p} \in \Lambda^{k}\left(T_{p}^{*} \mathbb{R}^{n}\right)$.

We now define the notion of a $k$-form on a manifold. Let $X \subseteq \mathbb{R}^{N}$ be an $n$ dimensional manifold. Then, for $p \in X$, the tangent space $T_{p} X \subseteq T_{p} \overline{\mathbb{R}}^{N}$.

Definition 6.14. A $k$-form $\omega$ on $X$ is a function on $X$ which assigns to each point $p \in X$ an element $\omega_{p} \in \Lambda^{k}\left(\left(T_{p} X\right)^{*}\right)$.

Suppose that $f: X \rightarrow \mathbb{R}$ is a $\mathcal{C}^{\infty}$ map, and let $f(p)=a$. Then $d f_{p}$ is of the form

$$
\begin{equation*}
d f_{p}: T_{p} X \rightarrow T_{a} \mathbb{R} \cong \mathbb{R} \tag{6.47}
\end{equation*}
$$

We can think of $d f_{p} \in\left(T_{p} X\right)^{*}=\Lambda^{1}\left(\left(T_{p} X\right)^{*}\right)$. So, we get a one-form $d f$ on $X$ which maps each $p \in X$ to $d f_{p}$.

Now, suppose

$$
\begin{align*}
& \mu \text { is a } k \text {-form on } X \text {, and }  \tag{6.48}\\
& \nu \text { is an } \ell \text {-form on } X \text {. } \tag{6.49}
\end{align*}
$$

For $p \in X$, we have

$$
\begin{align*}
& \mu_{p} \in \Lambda^{k}\left(T_{p}^{*} X\right) \text { and }  \tag{6.50}\\
& \nu_{p} \in \Lambda^{\ell}\left(T_{p}^{*} X\right) . \tag{6.51}
\end{align*}
$$

Taking the wedge product,

$$
\begin{equation*}
\mu_{p} \wedge \nu_{p} \in \Lambda^{k+\ell}\left(T_{p}^{*} X\right) \tag{6.52}
\end{equation*}
$$

The wedge product $\mu \wedge \nu$ is the $(k+\ell)$-form mapping $p \in X$ to $\mu_{p} \wedge \nu_{p}$.
Now we consider the pullback operation. Let $X \subseteq \mathbb{R}^{N}$ and $Y \subseteq \mathbb{R}^{\ell}$ be manifolds, and let $f: X \rightarrow Y$ be a $\mathcal{C}^{\infty}$ map. Let $p \in X$ and $a=f(p)$. We have the map

$$
\begin{equation*}
d f_{p}: T_{p} X \rightarrow T_{a} Y \tag{6.53}
\end{equation*}
$$

From this we get the pullback

$$
\begin{equation*}
\left(d f_{p}\right)^{*}: \Lambda^{k}\left(T_{a}^{*} Y\right) \rightarrow \Lambda^{k}\left(T_{p}^{*} X\right) \tag{6.54}
\end{equation*}
$$

Let $\omega$ be a $k$-form on $Y$. Then $f^{*} \omega$ is defined by

$$
\begin{equation*}
\left(f^{*} \omega\right)_{p}=\left(d f_{p}\right)^{*} \omega_{q} \tag{6.55}
\end{equation*}
$$

Let $f: X \rightarrow Y$ and $g: Y \rightarrow Z$ be $\mathcal{C}^{\infty}$ maps on manifolds $X, Y, Z$. Let $\omega$ be a $k$-form. Then

$$
\begin{equation*}
(g \circ f)^{*} \omega=f^{*}\left(g^{*} \omega\right), \tag{6.56}
\end{equation*}
$$

where $g \circ f: X \rightarrow Z$.
So far, the treatment of $k$-forms for manifolds has been basically the same as our earlier treatment of $k$-forms. However, the treatment for manifolds becomes more complicated when we study $\mathcal{C}^{\infty}$ forms.

Let $U$ be an open subset of $\mathbb{R}^{n}$, and let $\omega$ be a $k$-form on $U$. We can write

$$
\begin{equation*}
\omega=\sum a_{I}(x) d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}, \quad I=\left(i_{1}, \ldots, i_{k}\right) \tag{6.57}
\end{equation*}
$$

By definition, we say that $\omega \in \Omega^{k}(U)$ if each $A_{I} \in \mathcal{C}^{\infty}(U)$.
Let $V$ be an open subset of $\mathbb{R}^{k}$, and let $f: U \rightarrow V$ be a $\mathcal{C}^{\infty}$ map. Let $\omega \in \Omega^{k}(V)$. Then $f^{*} \omega \in \Omega^{k}(U)$. Now, we want to define what we mean by a $\mathcal{C}^{\infty}$ form on a manifold.

Let $X \subseteq \mathbb{R}^{n}$ be an $n$-dimensional manifold, and let $p \in X$. There exists an open set $U$ in $\mathbb{R}^{N}$, a neighborhood $V$ of $p$ in $\mathbb{R}^{N}$, and a diffeomorphism $\phi: U \rightarrow V \cap X$. The diffeomorphism $\phi$ is a parameterization of $X$ at $p$.

We can think of $\phi$ in the following two ways:

1. as a map of $U$ onto $V \cap X$, or
2. as a map of $U$ onto $V$, whose image is contained in $X$.

The second way of thinking about $\phi$ is actually the map $\iota_{X} \circ \phi$, where $\iota_{X}: X \rightarrow \mathbb{R}^{N}$ is the inclusion map. Note that $\iota_{X}: X \rightarrow \mathbb{R}^{N}$ is $\mathcal{C}^{\infty}$, because it extends to the identity $\operatorname{map} I: \mathbb{R}^{N} \rightarrow \mathbb{R}^{N}$.

We give two equivalent definitions for $\mathcal{C}^{\infty} k$-forms. Let $\omega$ be a $k$-form on $X$.
Definition 6.15. The $k$-form $\omega$ is $\mathcal{C}^{\infty}$ at $p$ if there exists a $k$-form $\tilde{\omega} \in \Omega^{k}(V)$ such that $\iota_{X}^{*} \tilde{w}=\omega$.

Definition 6.16. The $k$-form $\omega$ is $\mathcal{C}^{\infty}$ at $p$ if there exists a diffeomorphism $\phi: U \rightarrow$ $V \cap U$ such that $\phi^{*} \omega \in \Omega^{k}(U)$.

The first definition depends only on the choice of $\tilde{\omega}$, and the second definition depends only on the choice of $\phi$. So, if the definitions are equivalent, then neither definition depends on the choice of $\tilde{\omega}$ or the choice of $\phi$.

We show that these two definitions are indeed equivalent.
Claim. The above two definitions are equivalent.
Proof. First, we show that (def 6.15) $\Longrightarrow$ (def 6.16). Let $\omega=\iota_{X}^{*} \tilde{\omega}$. Then $\phi^{*} \omega=$ $\left(\iota_{X} \circ \phi\right)^{*} \tilde{\omega}$. The map $\iota \circ \phi: U \rightarrow V$ is $\mathcal{C}^{\infty}$, and $\tilde{\omega} \in \Omega^{k}(v)$, so $\phi^{*} \omega=\left(\iota_{X} \circ \phi\right)^{*} \tilde{\omega} \in \Omega^{k}(U)$.

Second, we show that (def 6.16) $\Longrightarrow$ (def 6.15). Let $\phi: U \rightarrow V \cap U$ be a diffeomorphism. Then $\phi^{-1}: V \cap X \rightarrow U$ can be extended to $\psi: V \rightarrow U$, where $\psi$ is $\mathcal{C}^{\infty}$. On $V \cap X$, the map $\phi=\iota_{X}^{*} \tilde{\omega}$, where $\tilde{\omega}=\psi^{*}\left(\phi^{*} \omega\right)$. It is easy to show that $\tilde{\omega}$ is $\mathcal{C}^{\infty}$.

Definition 6.17. The $k$-form $\omega$ is $\mathcal{C}^{\infty}$ if $\omega$ is $\mathcal{C}^{\infty}$ at $p$ for every point $p \in X$.
Notation. If $\omega$ is $\mathcal{C}^{\infty}$, then $\omega \in \Omega^{k}(X)$.
Theorem 6.18. If $\omega \in \Omega^{k}(X)$, then there exists a neighborhood $W$ of $X$ in $\mathbb{R}^{N}$ and a $k$-form $\tilde{\omega} \in \Omega^{k}(W)$ such that $\iota_{X}^{*} \tilde{\omega}=w$.

Proof. Let $p \in X$. There exists a neighborhood $V_{p}$ of $p$ in $\mathbb{R}^{N}$ and a $k$-form $\omega^{p} \in$ $\Omega^{k}\left(V_{p}\right)$ such that $\iota_{X}^{*} \omega^{p}=\omega$ on $V_{p} \cap X$.

Let

$$
\begin{equation*}
W \subseteq \bigcup_{p \in X} V_{p} \tag{6.58}
\end{equation*}
$$

The collection of sets $\left\{V_{p}: p \in X\right\}$ is an open cover of $W$. Let $\rho_{1}, i=1,2,3, \ldots$, be a partition of unity subordinate to this cover. So, $\rho_{i} \in \mathcal{C}_{0}^{\infty}(W)$ and supp $\rho_{i} \subset V_{p}$ for some $p$. Let

$$
\tilde{\omega}_{i}= \begin{cases}\rho_{i} \omega^{p} & \text { on } V_{p}  \tag{6.59}\\ 0 & \text { elsewhere }\end{cases}
$$

Notice that

$$
\begin{align*}
\iota_{X}^{*} \tilde{\omega}_{i} & =\iota_{X}^{*} \rho_{i} \iota_{X}^{*} \omega^{p}  \tag{6.60}\\
& =\left(\iota_{X}^{*} \rho_{i}\right) \omega .
\end{align*}
$$

Take

$$
\begin{equation*}
\tilde{\omega}=\sum_{i=1}^{\infty} \tilde{\omega}_{i} . \tag{6.61}
\end{equation*}
$$

This sum makes sense since we used a partition of unity. From the sum, we can see that $\tilde{w} \in \Omega^{k}(W)$. Finally,

$$
\begin{align*}
\iota_{X}^{*} \tilde{w} & =\left(\iota_{X}^{*} \sum \rho_{i}\right) \omega  \tag{6.62}\\
& =\omega .
\end{align*}
$$

Theorem 6.19. Let $X \subseteq \mathbb{R}^{N}$ and $Y \subseteq \mathbb{R}^{\ell}$ be manifolds, and let $f: X \rightarrow Y$ be a $\mathcal{C}^{\infty}$ map. If $\omega \in \Omega^{k}(X)$, then $f^{*} \omega \in \Omega^{k}(Y)$.

Proof. Take an open set $W$ in $\mathbb{R}^{\ell}$ such that $W \supset Y$, and take $\tilde{\omega} \in \Omega^{k}(W)$ such that $\iota_{X}^{*} \tilde{\omega}=\omega$. Take any $p \in X$ and $\phi: U \rightarrow V$ a parameterization of $X$ at $p$.

We show that the pullback $\phi^{*}\left(f^{*} \omega\right)$ is in $\Omega^{k}(U)$. We can write

$$
\begin{align*}
\phi^{*}\left(f^{*} \omega\right) & =\phi^{*} f^{*}\left(\iota_{X}^{*} \tilde{w}\right)  \tag{6.63}\\
& =(\iota \circ f \circ \phi)^{*} \tilde{\omega},
\end{align*}
$$

where in the last step we used the chain rule.
The form $\tilde{\omega} \in \Omega^{k}(W)$, where $W$ is open in $\mathbb{R}^{\ell}$, so $\iota \circ f \circ \phi: U \rightarrow W$. The theorem that we proved on Euclidean spaces shows that the r.h.s of Equation 6.63 is in $\Omega^{k}(U)$.

The student should check the following claim:
Claim. If $\mu, \nu \in \Omega^{k}(Y)$, then

$$
\begin{equation*}
f^{*}(\mu \wedge \nu)=f^{*} \mu \wedge f^{*} \nu \tag{6.64}
\end{equation*}
$$

The differential operation $d$ is an important operator on $k$-forms on manifolds.

$$
\begin{equation*}
d: \Omega^{k}(X) \rightarrow \Omega^{k+1}(X) \tag{6.65}
\end{equation*}
$$

Let $X \subseteq \mathbb{R}^{N}$ be a manifold, and let $\omega \in \Omega^{k}(X)$. There exists an open neighborhood $W$ of $X$ in $\mathbb{R}^{N}$ and a $k$-form $\tilde{\omega} \in \Omega^{k}(W)$ such that $\iota_{X}^{*} \tilde{\omega}=\omega$.

Definition 6.20. $d \omega=\iota_{X}^{*} d \tilde{\omega}$.
Why is this definition well-defined? It seems to depend on the choice of $\tilde{\omega}$.
Take a parameterization $\phi: U \rightarrow V \cap X$ of $X$ at $p$. Then

$$
\begin{align*}
\phi^{*} \iota_{X}^{*} d \tilde{\omega} & =\left(\iota_{X} \circ \phi\right)^{*} d \tilde{\omega} \\
& =d\left(\iota_{X} \circ \phi\right)^{*} \omega \\
& =d \phi^{*}\left(\iota_{X}^{*} \tilde{\omega}\right)  \tag{6.66}\\
& =d \phi^{*} \omega .
\end{align*}
$$

So,

$$
\begin{equation*}
\phi^{*} \iota_{X}^{*} d \tilde{\omega}=d \phi^{*} \omega . \tag{6.67}
\end{equation*}
$$

Take the inverse mapping $\phi^{-1}: V \cap X \rightarrow U$ and take the pullback $\left(\phi^{-1}\right)^{*}$ of each side of Equation 6.67, to obtain

$$
\begin{equation*}
\iota_{X}^{*} d \tilde{\omega}=\left(\phi^{-1}\right)^{*} d \phi^{*} \omega \tag{6.68}
\end{equation*}
$$

The r.h.s does not depend on $\tilde{\omega}$, so neither does the l.h.s.
To summarize this lecture, everything we did with $k$-forms on Euclidean space applies to $k$-forms on manifolds.

## Lecture 34

### 6.6 Orientation of Manifolds

Let $X$ be an $n$-dimensional manifold in $\mathbb{R}^{N}$. Assume that $X$ is a closed subset of $\mathbb{R}^{N}$. Let $f: X \rightarrow \mathbb{R}$ be a $\mathcal{C}^{\infty}$ map.

Definition 6.21. We remind you that the support of $f$ is defined to be

$$
\begin{equation*}
\operatorname{supp} f=\overline{\{x \in X: f(x) \neq 0\}} \tag{6.69}
\end{equation*}
$$

Since $X$ is closed, we don't have to worry about whether we are taking the closure in $X$ or in $\mathbb{R}^{n}$.

Note that

$$
\begin{equation*}
f \in \mathcal{C}_{0}^{\infty}(X) \Longleftrightarrow \operatorname{supp} f \text { is compact. } \tag{6.70}
\end{equation*}
$$

Let $\omega \in \Omega^{k}(X)$. Then

$$
\begin{equation*}
\operatorname{supp} \omega=\overline{\left\{p \in X: \omega_{p} \neq 0\right\}} \tag{6.71}
\end{equation*}
$$

We use the notation

$$
\begin{equation*}
\omega \in \Omega_{c}^{k}(X) \Longleftrightarrow \operatorname{supp} \omega \text { is compact. } \tag{6.72}
\end{equation*}
$$

We will be using partitions of unity, so we remind you of the definition:
Definition 6.22. A collection of functions $\left\{\rho_{i} \in \mathcal{C}_{0}^{\infty}(X): i=1,2,3, \ldots\right\}$ is a partition of unity if

1. $0 \leq \rho_{i}$,
2. For every compact set $A \subseteq X$, there exists $N>0$ such that $\operatorname{supp} \rho_{i} \cap A=\phi$ for all $i>N$,
3. $\sum \rho_{i}=1$.

Suppose the collection of $\operatorname{sets} \mathcal{U}=\left\{U_{\alpha}: \alpha \in I\right\}$ is a covering of $X$ by open subsets $U_{\alpha}$ of $X$.

Definition 6.23. The partition of unity $\rho_{i}, i=1,2,3, \ldots$, is subordinate to $\mathcal{U}$ if for every $i$, there exists $\alpha \in I$ such that supp $\rho_{i} \subseteq U_{\alpha}$.

Claim. Given a collection of sets $\mathcal{U}=\left\{U_{\alpha}: \alpha \in I\right\}$, there exists a partition of unity subordinate to $\mathcal{U}$.

Proof. For each $\alpha \in I$, let $\tilde{U}_{\alpha}$ be an open set in $\mathbb{R}^{N}$ such that $U_{\alpha}=\tilde{U}_{\alpha} \cap X$. We define the collection of sets $\tilde{\mathcal{U}}=\left\{\tilde{U}_{\alpha}: \alpha \in I\right\}$. Let

$$
\begin{equation*}
\tilde{U}=\bigcup \tilde{U}_{\alpha} \tag{6.73}
\end{equation*}
$$

From our study of Euclidean space, we know that there exists a partition of unity $\tilde{\rho}_{i} \in \mathcal{C}_{0}^{\infty}(\tilde{U}), i=1,2,3, \ldots$, subordinate to $\tilde{\mathcal{U}}$. Let $\iota_{X}: X \rightarrow \tilde{\mathcal{U}}$ be the inclusion map. Then

$$
\begin{equation*}
\rho_{i}=\tilde{\rho}_{i} \circ \iota_{X}=\iota_{X}^{*} \tilde{\rho}_{i}, \tag{6.74}
\end{equation*}
$$

which you should check.
We review orientations in Euclidean space before generalizing to manifolds. For a more comprehensive review, read section 7 of the Multi-linear Algebra notes.

Suppose $\mathbb{L}$ is a one-dimensional vector space and that $v \in \mathbb{L}-\{0\}$. The set $\mathbb{L}-\{0\}$ has two components:

$$
\begin{equation*}
\{\lambda v: \lambda>0\} \text { and }\{\lambda v: \lambda<0\} . \tag{6.75}
\end{equation*}
$$

Definition 6.24. An orientation of $\mathbb{L}$ is a choice of one of these components.
Notation. We call the preferred component $\mathbb{L}_{+}$(the positive component). We call the other component $\mathbb{L}_{-}$(the negative component).

We define a vector $v$ to be positively oriented if $v \in \mathbb{L}_{+}$.
Now, let $V$ be an $n$-dimensional vector space.
Definition 6.25. An orientation of $V$ is an orientation of the one-dimensional vector space $\Lambda^{n}\left(V^{*}\right)$. That is, an orientation of $V$ is a choice of $\Lambda^{n}\left(V^{*}\right)_{+}$.

Suppose that $V_{1}, V_{2}$ are oriented $n$-dimensional vector spaces, and let $A: V_{1} \rightarrow V_{2}$ be a bijective linear map.

Definition 6.26. The map $A$ is orientation preserving if

$$
\begin{equation*}
\omega \in \Lambda^{n}\left(V_{2}\right)_{+} \Longrightarrow A^{*} \omega \in \Lambda^{n}\left(V_{1}\right)_{+} . \tag{6.76}
\end{equation*}
$$

Suppose that $V_{3}$ is also an oriented $n$-dimensional vector space, and let $B: V_{2} \rightarrow V_{3}$ be a bijective linear map. If $A$ and $B$ are orientation preserving, then $B A$ is also orientation preserving.

Finally, let us generalize the notion of orientation to orientations of manifolds. Let $X \subseteq \mathbb{R}^{N}$ be an $n$-dimensional manifold.

Definition 6.27. An orientation of $X$ is a function on $X$ which assigns to each point $p \in X$ an orientation of $T_{p} X$.

We give two examples of orientations of a manifold:
Example 1: Let $\omega \in \Lambda^{n}(X)$, and suppose that $\omega$ is nowhere vanishing. Orient $X$ by assigning to $p \in X$ the orientation of $T_{p} X$ for which $\omega_{p} \in \Lambda^{n}\left(T_{p}^{*} X\right)_{+}$.

Example 2: Take $X=U$, an open subset of $\mathbb{R}^{n}$, and let

$$
\begin{equation*}
\omega=d x_{1} \wedge \cdots \wedge d x_{n} \tag{6.77}
\end{equation*}
$$

Define an orientation as in the first example. This orientation is called the standard orientation of $U$.

Definition 6.28. An orientation of $X$ is a $\mathcal{C}^{\infty}$ orientation if for every point $p \in X$, there exists a neighborhood $U$ of $p$ in $X$ and an $n$-form $\omega \in \Omega^{n}(U)$ such that for all points $q \in U, \omega_{q} \in \Lambda^{n}\left(T_{q}^{*} X\right)_{+}$.

From now on, we will only consider $\mathcal{C}^{\infty}$ orientations.
Theorem 6.29. If $X$ is oriented, then there exists $\omega \in \Omega^{n}(X)$ such that for all $p \in X, \omega_{p} \in \Lambda^{n}\left(T_{p}^{*} X\right)_{+}$.

Proof. For every point $p \in X$, there exists a neighborhood $U_{p}$ of $p$ and an $n$ - form $\omega^{(p)} \in \Omega^{n}\left(U_{p}\right)$ such that for all $q \in U_{p},\left(\omega^{(p)}\right)_{q} \in \Lambda^{n}\left(T_{Q}^{*} X\right)_{+}$.

Take $\rho_{i}, i=1,2, \ldots$, a partition of unity subordinate to $\mathcal{U}=\left\{U_{p}: p \in X\right\}$. For every $i$, there exists a point $p$ such that $\rho_{i} \in \mathcal{C}_{0}^{\infty}\left(U_{p}\right)$. Let

$$
\omega_{i}= \begin{cases}\rho_{i} \omega^{(p)} & \text { on } U_{p}  \tag{6.78}\\ 0 & \text { on the } X-U_{p}\end{cases}
$$

Since the $\rho_{i}$ 's are compactly supported, $\omega_{i}$ is a $\mathcal{C}^{\infty}$ map. Let

$$
\begin{equation*}
\omega=\sum \omega_{i} . \tag{6.79}
\end{equation*}
$$

One can check that $\omega$ is positively oriented at every point.
Definition 6.30. An $n$-form $\omega \in \Omega^{n}(X)$ with the property hypothesized in the above theorem is called a volume form.

Remark. If $\omega_{1}, \omega_{2}$ are volume forms, then we can write $\omega_{2}=f \omega_{1}$, for some $f \in$ $\mathcal{C}^{\infty}(X)$ (where $f \neq 0$ everywhere). In general, $f(p)>0$ because $\left(\omega_{1}\right)_{p},\left(\omega_{2}\right)_{p} \in$ $\Lambda^{n}\left(T_{p}^{*} X\right)_{+}$. So, if $\omega_{1}, \omega_{2}$ are volume forms, then $\omega_{2}=f \omega_{1}$, for some $f \in \mathcal{C}^{\infty}(X)$ such that $f>0$.

Remark. Problem \#6 on the homework asks you to show that if $X$ is orientable and connected, then there are exactly two ways to orient it. This is easily proved using the above Remark.

Suppose that $X \subseteq \mathbb{R}^{n}$ is a one-dimensional manifold (a "curve"). Then $T_{p} X$ is one-dimensional. We can find vectors $v,-v \in T_{p} X$ such that $\|v\|=1$. An orientation of $X$ is just a choice of $v$ or $-v$.

Now, suppose that $X$ is an $(n-1)$-dimensional manifold in $\mathbb{R}^{n}$. Define

$$
\begin{equation*}
N_{p} X=\left\{v \in T_{p} \mathbb{R}^{n}: v \perp w \text { for all } w \in T_{p} X\right\} . \tag{6.80}
\end{equation*}
$$

Then $\operatorname{dim} N_{p} X=1$, so you can find $v,-v \in N_{p} X$ such that $\|v\|=1$. By Exercise \#5 in section 4 of the Multi-linear Algebra Notes, an orientation of $T_{p} X$ is just a choice of $v$ or $-v$.

Suppose $X_{1}, X_{2}$ are oriented $n$-dimensional manifolds, and let $f: X_{1} \rightarrow X_{2}$ be a diffeomorphism.

Definition 6.31. The map $f$ is orientation preserving if for every $p \in X_{1}$,

$$
\begin{equation*}
d f_{p}: T_{p} X_{1} \rightarrow T_{q} X_{2} \tag{6.81}
\end{equation*}
$$

is orientation preserving, where $q=f(p)$.
Remark. Let $\omega_{2}$ be a volume form on $X_{2}$. Then $f$ is orientation preserving if and only if $f^{*} \omega_{2}=\omega_{1}$ is a volume form on $X_{1}$.

We look at an example of what it means for a map to be orientation preserving. Let $U, V$ be open sets on $\mathbb{R}^{n}$ with the standard orientation. Let $f: U \rightarrow V$ be a diffeomorphism. So, by definition, the form

$$
\begin{equation*}
d x_{1} \wedge \cdots \wedge d x_{n} \tag{6.82}
\end{equation*}
$$

is a volume form of $V$. The form

$$
\begin{equation*}
f^{*} d x_{1} \wedge \cdots \wedge d x_{n}=\operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}\right] d x_{1} \wedge \cdots \wedge d x_{n} \tag{6.83}
\end{equation*}
$$

is a volume form of $U$ if and only if

$$
\begin{equation*}
\operatorname{det}\left[\frac{\partial f_{i}}{\partial x_{j}}\right]>0 \tag{6.84}
\end{equation*}
$$

that is, if and only if $f$ is orientation preserving in our old sense.
Now that we have studied orientations of manifolds, we have all of the ingredients we need to study integration theory for manifolds.

## Lecture 35

Before moving on to integration, we make a few more remarks about orientations.
Let $X, Y$ be oriented manifolds. A diffeomorphism $f: X \rightarrow Y$ is orientation preserving if for every $p \in X$, the map

$$
\begin{equation*}
d f_{p}: T_{p} X \rightarrow T_{q} Y \tag{6.85}
\end{equation*}
$$

is orientation preserving, where $q=f(p)$.
Let $V$ be open in $X$, let $U$ be open in $\mathbb{R}^{n}$, and let $\phi: U \rightarrow V$ be a parameterization.
Definition 6.32. The map $\phi$ is an oriented parameterization if it is orientation preserving.

Suppose $\phi$ is orientation reversing. Let $A: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ be the linear map defined by

$$
\begin{equation*}
A\left(x_{1}, \ldots, x_{n}\right)=\left(-x_{1}, x_{2}, \ldots, x_{n}\right) \tag{6.86}
\end{equation*}
$$

The map $A$ is orientation reversing. Let $U^{\prime}=A^{-1}(U)$, and define $\phi^{\prime}=\phi \circ A: U^{\prime} \rightarrow V$. Both $\phi$ and $A$ are orientation reversing, so $\phi^{\prime}$ is orientation preserving.

Thus, for every point $p \in X$, there exists an oriented parameterization of $X$ at $p$.

### 6.7 Integration on Manifolds

Our goal for today is to take any $\omega \in \Omega_{c}^{n}(X)$ and define

$$
\begin{equation*}
\int_{X} \omega \tag{6.87}
\end{equation*}
$$

First, we consider a special case:
Let $\phi: U \rightarrow V$ be an oriented parameterization. Let $U$ be open in $\mathbb{R}^{n}$, and let $V$ be open in $X$. Take any $\omega \in \Omega_{c}^{n}(V)$. Then

$$
\begin{equation*}
\int_{V} \omega=\int_{U} \phi^{*} \omega \tag{6.88}
\end{equation*}
$$

where $\phi^{*} \omega=f(x) d x_{1} \wedge \cdots \wedge d x_{n}$, where $f \in \mathcal{C}_{0}^{\infty}(U)$ and

$$
\begin{equation*}
\int_{U} \phi^{*} \omega=\int_{U} f \tag{6.89}
\end{equation*}
$$

Claim. The above definition for $\int \omega$ does not depend on the choice of oriented parameterization $\phi$.

Proof. Let $\phi_{i}: U_{i} \rightarrow V, i=1,2$, be oriented parameterizations. Let $\omega \in \Omega_{c}^{n}\left(V_{1} \cap V_{2}\right)$. Define

$$
\begin{align*}
& U_{1,2}=\phi_{1}^{-1}\left(V_{1} \cap V_{2}\right),  \tag{6.90}\\
& U_{2,1}=\phi_{2}^{-1}\left(V_{1} \cap V_{2}\right), \tag{6.91}
\end{align*}
$$

which are open sets in $\mathbb{R}^{n}$.
Both $\phi_{1}$ and $\phi_{2}$ are diffeomorphisms, and we have the diagram


Therefore, $f=\phi_{2}^{-1} \circ \phi_{1}$ is a diffeomorphism, and $\phi_{1}=\phi_{2} \circ f$. Integrating,

$$
\begin{align*}
\int_{U_{1}} \phi_{1}^{*} \omega & =\int_{U_{1,2}} \phi_{1}^{*} \omega \\
& =\int_{U_{1,2}}\left(\phi_{2} \circ f\right)^{*} \omega  \tag{6.93}\\
& =\int_{U_{1,2}} f^{*}\left(\phi_{2}^{*} \omega\right)
\end{align*}
$$

Note that $f$ is orientation preserving, because $\phi_{1}$ and $\phi_{2}$ are orientation preserving.
Using the change of variables formula,

$$
\begin{align*}
\int_{U_{1,2}} f^{*} \phi_{2}^{*} \omega & =\int_{U_{2,1}} \phi_{2}^{*} \omega  \tag{6.94}\\
& =\int_{U_{2}} \phi_{2}^{*} \omega .
\end{align*}
$$

So, for all $\omega \in \Omega_{c}^{n}\left(V_{1} \cap V_{2}\right)$,

$$
\begin{equation*}
\int_{V_{1}} \omega=\int_{U_{1}} \phi_{1}^{*} \omega=\int_{U_{2}} \phi_{2}^{*} \omega=\int_{V_{2}} \omega . \tag{6.95}
\end{equation*}
$$

Above, we showed above how to take integrals over open sets, and now we generalize.

To define the integral, we need the following two inputs:

1. a set of oriented parameterizations $\phi_{i}: U_{i} \rightarrow V_{i}, i=1,2, \ldots$, such that $X=$ $\bigcup V_{i}$,
2. a partition of unity $\rho_{i} \in \mathcal{C}_{0}^{\infty}\left(V_{i}\right)$ subordinate to the cover $\left\{V_{i}\right\}$.

Definition 6.33. Let $\omega \in \Omega_{c}^{n}(X)$. We define the integral

$$
\begin{equation*}
\int_{X} \omega=\sum_{i=1}^{\infty} \int_{V_{i}} \rho_{i} \omega \tag{6.96}
\end{equation*}
$$

One can check various standard properties of integrals, such as linearity:

$$
\begin{equation*}
\int_{X} \omega_{1}+\omega_{2}=\int_{X} \omega_{1}+\int_{X} \omega_{2} \tag{6.97}
\end{equation*}
$$

We now show that this definition is independent of the choice of the two inputs (the parameterizations and the partition of unity).

Consider two different inputs:

1. oriented parameterizations $\phi_{j}^{\prime}: U_{j}^{\prime} \rightarrow V_{j}^{\prime}, j=1,2, \ldots$, such that $X=\bigcup V_{j}^{\prime}$,
2. a partition of unity $\rho_{i}^{\prime} \in \mathcal{C}_{0}^{\infty}\left(V_{j}^{\prime}\right)$ subordinate to the cover $\left\{V_{j}^{\prime}\right\}$.

Then,

$$
\begin{align*}
\int_{V_{i}} \rho_{i} \omega & =\int_{V_{i}}\left(\sum_{j=1}^{\infty} \rho_{j}^{\prime} \omega\right) \\
& =\sum_{j=1}^{\infty} \int_{V_{i}} \rho_{i} \rho_{j}^{\prime} \omega  \tag{6.98}\\
& =\sum_{j=1}^{\infty} \int_{V_{i} \cap V_{j}^{\prime}} \rho_{i} \rho_{j}^{\prime} \omega .
\end{align*}
$$

Summing over $i$,

$$
\begin{align*}
\sum_{i} \int_{V_{i}} \rho_{i} \omega & =\sum_{i, j=1}^{\infty} \int_{V_{i} \cap V_{j}^{\prime}} \rho_{i} \rho_{j}^{\prime} \omega \\
& =\sum_{j} \int_{V_{j}^{\prime}} \rho_{j}^{\prime} \omega, \tag{6.99}
\end{align*}
$$

where the first term equals the last term by symmetry. Therefore, the integral $\int \omega$ is independent of the choices of these two inputs.

Let $X \subseteq \mathbb{R}^{N}$ be an oriented connected $n$-dimensional manifold.
Theorem 6.34. For any $\omega \in \Omega_{c}^{n}(X)$, the following are equivalent:

1. $\int_{X} \omega=0$,
2. $\omega \in d \Omega_{c}^{n-1}(X)$.

Proof. This will be a five step proof:
Step 1: The following lemma is called the Connectivity Lemma.
Lemma 6.35. Given $p, q \in X$, there exists open sets $W_{j}, j=0, \ldots, N+1$, such that each $W_{j}$ is diffeomorphic to an open set in $\mathbb{R}^{n}$, and such that $p \in W_{0}, q \in W_{N+1}$, and $W_{i} \cap W_{i+1} \neq \phi$.

Proof Idea: Fix $p$. The points $q$ for which this is true form an open set. The points $q$ for which this isn't true also form an open set. Since $X$ is connected, only one of these sets is in $X$.

Step 2: Let $\omega_{1}, \omega_{2} \in \Omega_{c}^{n}(X)$. We say that $\omega_{1} \sim \omega_{2}$ if

$$
\begin{equation*}
\int_{X} \omega_{1}=\int_{X} \omega_{2} \tag{6.100}
\end{equation*}
$$

We can restate the theorem as

$$
\begin{equation*}
\omega_{1} \sim \omega_{2} \Longleftrightarrow \omega_{1}-\omega_{2} \in d \Omega_{c}^{n-1}(X) \tag{6.101}
\end{equation*}
$$

Step 3: It suffices to prove the statement (6.101) for $\omega_{1} \in \Omega_{c}^{n}(V)$ and $\omega_{2} \in \Omega_{c}^{n}\left(V^{\prime}\right)$, where $V, V^{\prime}$ are diffeomorphic to open sets in $\mathbb{R}^{n}$.

Step 4: We use a partition of unity
Lemma 6.36. The theorem is true if $V=V^{\prime}$.
Proof. Let $\phi: U \rightarrow V$ be an orientation preserving parameterization. If $\omega_{1} \sim \omega_{2}$, then

$$
\begin{equation*}
\int \phi^{*} \omega_{1}=\int \phi^{*} \omega_{2} \tag{6.102}
\end{equation*}
$$

which is the same as saying that

$$
\begin{equation*}
\phi^{*} \omega_{1}-\phi^{*} \omega_{2} \in d \Omega_{c}^{n-1}(U) \tag{6.103}
\end{equation*}
$$

which is the same as saying that

$$
\begin{equation*}
\omega_{1}-\omega_{2} \in d \Omega_{c}^{n-1}(V) \tag{6.104}
\end{equation*}
$$

Step 5: In general, by the Connectivity Lemma, there exists sets $W_{i}, i=0, \ldots, N+$ 1 , such that each $W_{i}$ is diffeomorphic to an open set in $\mathbb{R}^{n}$. We can choose $W_{0}=V$ and $W_{N+1}=V^{\prime}$ and $W_{i} \cap W_{i+1} \neq \phi$ (where $\phi$ here is the empty set).

We can choose $\mu_{i} \in \Omega_{c}^{n}\left(W_{i} \cap W_{i+1}\right)$ such that

$$
\begin{equation*}
c=\int_{V} \omega_{1}=\int \mu_{0}=\cdots=\int \mu_{N+1}=\int_{V^{\prime}} \omega_{2} . \tag{6.105}
\end{equation*}
$$

So,

$$
\begin{equation*}
\omega_{1} \sim \mu_{0} \sim \cdots \sim \mu_{N} \sim \omega_{2} \tag{6.106}
\end{equation*}
$$

We know that $\mu_{0}-\omega_{1} \in d \Omega_{c}^{n-1}$ and $\omega_{2}-\mu_{N+1} \in d \Omega_{C}^{n-1}$ Also, each difference $\omega_{i}-\omega_{i+1} \in$ $d \Omega_{c}^{n-1}$. Therefore, $\omega_{1}-\omega_{2} \in d \Omega_{c}^{n-1}$.

### 6.8 Degree on Manifolds

Suppose that $X_{1}, X_{2}$ are oriented $n$-dimensional manifolds, and let $f: X_{1} \rightarrow X_{2}$ be a proper map (that is, for every compact set $A \subseteq X$, the set pre-image $f^{-1}(A)$ is compact). It follows that if $\omega \in \Omega_{c}^{k}\left(X_{2}\right)$, then $f^{*} \omega \in \Omega_{c}^{k}\left(X_{1}\right)$.

Theorem 6.37. If $X_{1}, X_{2}$ are connected and $f: X_{1} \rightarrow X_{2}^{\prime}$ is a proper $\mathcal{C}^{\infty}$ map, then there exists a topological invariant of $f$ (called the degree of $f$ ) written $\operatorname{deg}(f)$ such that for every $\omega \in \Omega_{c}^{n}\left(X_{2}\right)$,

$$
\begin{equation*}
\int_{X_{1}} f^{*} \omega=\operatorname{deg}(f) \int_{X_{2}} \omega . \tag{6.107}
\end{equation*}
$$

Proof. The proof is pretty much verbatim of the proof in Euclidean space.
Let us look at a special case. Let $\phi_{1}: U \rightarrow V$ be an oriented parameterization, and let $V_{1}$ be open in $X_{1}$. Let $f: X_{1} \rightarrow X_{2}$ be an oriented diffeomorphism. Define $\phi_{2}=f \circ \phi_{1}$, which is of the form $\phi_{2}: U \rightarrow V_{2}$, where $V_{2}=f\left(V_{1}\right)$. Notice that $\phi_{2}$ is an oriented parameterization of $V_{2}$.

Take $\omega \in \Omega_{c}^{n}\left(V_{2}\right)$ and compute the integral

$$
\begin{align*}
\int_{V_{1}} f^{*} \omega & =\int_{U} \phi_{1}^{*} f^{*} \omega \\
& =\int_{U}\left(f \circ \phi_{1}\right)^{*} \omega  \tag{6.108}\\
& =\int_{U} \phi_{2}^{*} \omega
\end{align*}
$$

The $n$-form $\omega$ is compactly supported on $V_{2}$, so

$$
\begin{align*}
\int_{V_{1}} f^{*} \omega & =\int_{U} \phi_{2}^{*} \omega  \tag{6.109}\\
& =\int_{X_{2}} \omega
\end{align*}
$$

On the other hand,

$$
\begin{equation*}
\int_{X_{1}} f^{*} \omega=\int_{V_{1}} f^{*} \omega \tag{6.110}
\end{equation*}
$$

Combining these results,

$$
\begin{equation*}
\int_{X_{1}} f^{*} \omega=\int_{X_{2}} \omega \tag{6.111}
\end{equation*}
$$

Therefore,

$$
\begin{equation*}
\operatorname{deg}(f)=1 \tag{6.112}
\end{equation*}
$$

So, we have proved the following theorem, which is the Change of Variables theorem for manifolds:

Theorem 6.38. Let $X_{1}, X_{2}$ be connected oriented $n$-dimensional manifolds, and let $f: X_{1} \rightarrow X_{2}$ be an orientation preserving diffeomorphism. Then, for all $\omega \in \Omega_{c}^{n}\left(X_{2}\right)$,

$$
\begin{equation*}
\int_{X_{1}} f^{*} \omega=\int_{X_{2}} \omega \tag{6.113}
\end{equation*}
$$

## Lecture 36

The first problem on today's homework will be to prove the inverse function theorem for manifolds. Here we state the theorem and provide a sketch of the proof.

Let $X, Y$ be $n$-dimensional manifolds, and let $f: X \rightarrow Y$ be a $\mathcal{C}^{\infty}$ map with $f(p)=p_{1}$.

Theorem 6.39. If $d f_{p}: T_{p} X \rightarrow T_{p_{1}} Y$ is bijective, then $f$ maps a neighborhood $V$ of $p$ diffeomorphically onto a neighborhood $V_{1}$ of $p_{1}$.

Sketch of proof: Let $\phi: U \rightarrow V$ be a parameterization of $X$ at $p$, with $\phi(q)=p$. Similarly, let $\phi_{1}: U_{1} \rightarrow V_{1}$ be a parameterization of $Y$ at $p_{1}$, with $\phi_{1}\left(q_{1}\right)=p_{1}$.

Show that we can assume that $f: V \rightarrow V_{1}$ (Hint: if not, replace $V$ by $V \cap f^{-1}\left(V_{1}\right)$ ).
Show that we have a diagram

$$
\begin{array}{cc}
V \xrightarrow{f} V_{1} \\
\phi \uparrow &  \tag{6.114}\\
& \phi_{1} \uparrow \\
U \xrightarrow{g} & U_{1},
\end{array}
$$

which defines $g$,

$$
\begin{gather*}
g=\phi_{1}^{-1} \circ f \circ \phi,  \tag{6.115}\\
g(q)=q_{1} . \tag{6.116}
\end{gather*}
$$

So,

$$
\begin{equation*}
(d g)_{q}=\left(d \phi_{1}\right)_{q_{1}}^{-1} \circ d f_{p} \circ(d \phi)_{q} . \tag{6.117}
\end{equation*}
$$

Note that all three of the linear maps on the r.h.s. are bijective, so $(d g)_{q}$ is a bijection. Use the Inverse Function Theorem for open sets in $\mathbb{R}^{n}$.

This ends our explanation of the first homework problem.
Last time we showed the following. Let $X, Y$ be $n$-dimensional manifolds, and let $f: X \rightarrow Y$ be a proper $\mathcal{C}^{\infty}$ map. We can define a topological invariant $\operatorname{deg}(f)$ such that for every $\omega \in \Omega_{c}^{n}(Y)$,

$$
\begin{equation*}
\int_{X} f^{*} \omega=\operatorname{deg}(f) \int_{Y} \omega \tag{6.118}
\end{equation*}
$$

There is a recipe for calculating the degree, which we state in the following theorem. We lead into the theorem with the following lemma.

First, remember that we defined the set $C_{f}$ of critical points of $f$ by

$$
\begin{equation*}
p \in C_{f} \Longleftrightarrow d f_{p}: T_{p} X \rightarrow T_{q} Y \text { is not surjective, } \tag{6.119}
\end{equation*}
$$

where $q=f(p)$.

Lemma 6.40. Suppose that $q \in Y-f\left(C_{f}\right)$. Then $f^{-1}(q)$ is a finite set.
Proof. Take $p \in f^{-1}(q)$. Since $p \notin C_{f}$, the map $d f_{p}$ is bijective. The Inverse Function Theorem tells us that $f$ maps a neighborhood $U_{p}$ of $p$ diffeomorphically onto an open neighborhood of $q$. So, $U_{p} \cap f^{-1}(q)=p$.

Next, note that $\left\{U_{p}: p \in f^{-1}(q)\right\}$ is an open covering of $f^{-1}(q)$. Since $f$ is proper, $f^{-1}(q)$ is compact, so there exists a finite subcover $U_{p_{1}}, \ldots, U_{p_{N}}$. Therefore, $f^{-1}(q)=\left\{p_{1}, \ldots, p_{N}\right\}$.

The following theorem gives a recipe for computing the degree.
Theorem 6.41.

$$
\begin{equation*}
\operatorname{deg}(f)=\sum_{i=1}^{N} \sigma_{p_{i}} \tag{6.120}
\end{equation*}
$$

where

$$
\sigma_{p_{i}}= \begin{cases}+1 & \text { if } d f_{p_{i}}: T_{p_{i}} X \rightarrow T_{q} Y \text { is orientation preserving },  \tag{6.121}\\ -1 & \text { if } d f_{p_{i}}: T_{p_{i}} X \rightarrow T_{q} Y \text { is orientation reversing }\end{cases}
$$

Proof. The proof is basically the same as the proof in Euclidean space.
We say that $q \in Y$ is a regular value of $f$ if $q \notin f\left(C_{f}\right)$. Do regular values exist? We showed that in the Euclidean case, the set of non-regular values is of measure zero (Sard's Theorem). The following theorem is the analogous theorem for manifolds.

Theorem 6.42. If $q_{0} \in Y$ and $W$ is a neighborhood of $q_{0}$ in $Y$, then $W-f\left(C_{f}\right)$ is non-empty. That is, every neighborhood of $q_{0}$ contains a regular value (this is known as the Volume Theorem).

Proof. We reduce to Sard's Theorem.
The set $f^{-1}\left(q_{0}\right)$ is a compact set, so we can cover $f^{-1}\left(q_{0}\right)$ by open sets $V_{i} \subset X, i=$ $1, \ldots, N$, such that each $V_{i}$ is diffeomorphic to an open set in $\mathbb{R}^{n}$.

Let $W$ be a neighborhood of $q_{0}$ in $Y$. We can assume the following:

1. $W$ is diffeomorphic to an open set in $\mathbb{R}^{n}$,
2. $f^{-1}(W) \subset \bigcup V_{i}$ (which is Theorem 4.3 in the Supp. Notes),
3. $f\left(V_{i}\right) \subseteq W$ (for, if not, we can replace $V_{i}$ with $V_{i} \cap f^{-1}(W)$ ).

Let $U$ and the sets $U_{i}, i=1, \ldots, N$, be open sets in $\mathbb{R}^{n}$. Let $\phi: U \rightarrow W$ and the maps $\phi_{i}: U_{i} \rightarrow V_{i}$ be diffeomorphisms. We have the following diagram:

$$
\begin{align*}
& V_{i} \xrightarrow{f} W \\
& \phi_{i}, \cong \uparrow \quad \phi, \cong \uparrow  \tag{6.122}\\
& U_{i} \xrightarrow{g_{i}} U,
\end{align*}
$$

which define the maps $g_{i}$,

$$
\begin{equation*}
g_{i}=\phi^{-1} \circ f \circ \phi_{i} . \tag{6.123}
\end{equation*}
$$

By the chain rule, $x \in C_{g_{i}} \Longrightarrow \phi_{i}(x) \in C_{f}$, so

$$
\begin{equation*}
\phi_{i}\left(C_{g_{i}}=C_{f} \cap V_{i} .\right. \tag{6.124}
\end{equation*}
$$

So,

$$
\begin{equation*}
\phi\left(g_{i}\left(C_{g_{i}}\right)\right)=f\left(C_{f} \cap V_{i}\right) . \tag{6.125}
\end{equation*}
$$

Then,

$$
\begin{equation*}
f\left(C_{f}\right) \cap W=\bigcup_{i} \phi\left(g_{i}\left(C_{g_{i}}\right)\right) . \tag{6.126}
\end{equation*}
$$

Sard's Theorem tells us that $g_{i}\left(C_{g_{i}}\right)$ is a set of measure zero in $U$, so

$$
\begin{gather*}
U-\bigcup g_{i}\left(C_{g_{i}}\right) \text { is non-empty, so }  \tag{6.127}\\
W-f\left(C_{f}\right) \text { is also non-empty. } \tag{6.128}
\end{gather*}
$$

In fact, this set is not only non-empty, but is a very, very "full" set.
Let $f_{0}, f_{1}: X \rightarrow Y$ be proper $\mathcal{C}^{\infty}$ maps. Suppose there exists a proper $\mathcal{C}^{\infty}$ map $F: X \times[0,1] \rightarrow Y$ such that $F(x, 0)=f_{0}(x)$ and $F(x, 1)=f_{1}(x)$. Then

$$
\begin{equation*}
\operatorname{deg}\left(f_{0}\right)=\operatorname{deg}\left(f_{1}\right) \tag{6.129}
\end{equation*}
$$

In other words, the degree is a homotopy. The proof of this is essential the same as before.

### 6.9 Hopf Theorem

The Hopf Theorem is a nice application of the homotopy invariance of the degree.
Define the $n$-sphere

$$
\begin{equation*}
S^{n}=\left\{v \in \mathbb{R}^{n+1}:\|v\|=1\right\} . \tag{6.130}
\end{equation*}
$$

Hopf Theorem. Let $n$ be even. Let $f: S^{n} \rightarrow \mathbb{R}^{n+1}$ be a $\mathcal{C}^{\infty}$ map. Then, for some $v \in S^{n}$,

$$
\begin{equation*}
f(v)=\lambda v, \tag{6.131}
\end{equation*}
$$

for some scalar $\lambda \in \mathbb{R}$.
Proof. We prove the contrapositive. Assume that no such $v$ exists, and take $w=f(v)$. Consider $w-\langle v, w\rangle v \equiv w-w_{1}$. It follows that $w-w_{1} \neq 0$.

Define a new map $\tilde{f}: S^{n} \rightarrow S^{n}$ by

$$
\begin{equation*}
\tilde{f}(v)=\frac{f(v)-\langle v, f(x)\rangle}{\|f(v)-\langle v, f(x)\rangle\|} \tag{6.132}
\end{equation*}
$$

Note that $\left(w-w_{1}\right) \perp v$, so $\tilde{f}(v) \perp v$.
Define a family of functions

$$
\begin{gather*}
f_{t}: S^{n} \rightarrow S^{n}  \tag{6.133}\\
f_{t}(v)=(\cos t) v+(\sin t) \tilde{w} \tag{6.134}
\end{gather*}
$$

where $\tilde{w}=\tilde{f}(v)$ has the properties $\|\tilde{w}\|=1$ and $\tilde{w} \perp v$.
We compute the degree of $f_{t}$. When $t=0, f_{t}=\mathrm{id}$, so

$$
\begin{equation*}
\operatorname{deg}\left(f_{t}\right)=\operatorname{deg}\left(f_{0}\right)=1 \tag{6.135}
\end{equation*}
$$

When $t=\pi, f_{t}(v)=-v$. But, if $n$ is even, a map from $S^{n} \rightarrow S^{n}$ mapping $v \rightarrow(-v)$ has degree -1 . We have arrived at a contradiction.

## Lecture 37

### 6.10 Integration on Smooth Domains

Let $X$ be an oriented $n$-dimensional manifold, and let $\omega \in \Omega_{c}^{n}(X)$. We defined the integral

$$
\begin{equation*}
\int_{X} \omega \tag{6.136}
\end{equation*}
$$

but we can generalize the integral

$$
\begin{equation*}
\int_{D} \omega \tag{6.137}
\end{equation*}
$$

for some subsets $D \subseteq X$. We generalize, but only to very simple subsets called smooth domains (essentially manifolds-with-boundary). The prototypical smooth domain is the half plane:

$$
\begin{equation*}
\mathbb{H}^{n}=\left\{\left(x_{1}, \ldots, x_{n}\right) \in \mathbb{R}^{n}: x_{1} \leq 0\right\} \tag{6.138}
\end{equation*}
$$

Note that the boundary of the half plane is

$$
\begin{equation*}
\operatorname{Bd}\left(\mathbb{H}^{n}\right)=\left\{\left(x_{1}, \ldots, x_{n}\right) \in \mathbb{R}^{n}: x_{1}=0\right\} \tag{6.139}
\end{equation*}
$$

Definition 6.43. A closed subset $D \subseteq X$ is a smooth domain if for every point $p \in \operatorname{Bd}(D)$, there exists a parameterization $\phi: U \rightarrow V$ of $X$ at $p$ such that $\phi(U \cap$ $\left.\mathbb{H}^{n}\right)=V \cap D$.

Definition 6.44. The map $\phi$ is a parameterization of $D$ at $p$.
Note that $\phi: U \cap \mathbb{H}^{n} \rightarrow V \cap D$ is a homeomorphism, so it maps boundary points to boundary points. So, it maps $U^{b}=U \cap \operatorname{Bd}\left(\mathbb{H}^{n}\right)$ onto $V^{b}=V \cap \operatorname{Bd}(D)$.

Let $\psi=\phi \mid U^{b}$. Then $\psi: U^{b} \rightarrow V^{b}$ is a diffeomorphism. The set $U^{b}$ is an open set in $\mathbb{R}^{n-1}$, and $\psi$ is a parameterization of the $\operatorname{Bd}(D)$ at $p$. We conclude that

$$
\begin{equation*}
\mathrm{Bd}(D) \text { is an }(n-1) \text {-dimensional manifold. } \tag{6.140}
\end{equation*}
$$

Here are some examples of how smooth domains appear in nature:
Let $f: X \rightarrow \mathbb{R}$ be a $\mathcal{C}^{\infty}$ map, and assume that $f^{-1}(0) \cap C_{f}=\phi$ (the empty set). That is, for all $p \in f^{-1}(0), d f_{p} \neq 0$.

Claim. The set $D=\{x \in X: f(x) \leq 0\}$ is a smooth domain.
Proof. Take $p \in \operatorname{Bd}(D)$, so $p=f^{-1}(0)$. Let $\phi: U \rightarrow V$ be a parameterization of $X$ at $p$. Consider the map $g=f \circ \phi: U \rightarrow \mathbb{R}$. Let $q \in U$ and $p=\phi(q)$. Then

$$
\begin{equation*}
\left(d g_{q}\right)=d f_{p} \circ(d \phi)_{q} . \tag{6.141}
\end{equation*}
$$

We conclude that $d g_{q} \neq 0$.

By the canonical submersion theorem, there exists a diffeomorphism $\psi$ such that $g \circ \psi=\pi$, where $\pi$ is the canonical submersion mapping $\left(x, \ldots, x_{n}\right) \rightarrow x_{1}$. We can write simply $g \circ \psi=x_{1}$. Replacing $\phi=\phi_{\text {old }}$ by $\phi=\phi_{\text {new }}=\phi_{\text {old }} \circ \psi$, we get the new map $\phi: U \rightarrow V$ which is a parameterization of $X$ at $p$ with the property that $f \circ \phi\left(x_{1}, \ldots, x_{n}\right)=x_{1}$. Thus, $\phi$ maps $\mathbb{H}^{n} \cap U$ onto $D \cap V$.

We give an example of using the above claim to construct a smooth domain. Let $X=\mathbb{R}^{n}$, and define

$$
\begin{equation*}
f(x)=1-\left(x_{1}^{2}+\cdots+x_{n}^{2}\right) . \tag{6.142}
\end{equation*}
$$

By definition,

$$
\begin{equation*}
f(x) \leq 0 \Longleftrightarrow x \in B^{n} \tag{6.143}
\end{equation*}
$$

where $B^{n}=\left\{x \in \mathbb{R}^{n}:\|x\| \leq 1\right\}$ is the "unit ball." So, the unit ball $B^{n}$ is a smooth domain.

We now define orientations of smooth domains. Assume that $X$ is oriented, and let $D$ be a smooth domain. Let $\phi: U \rightarrow V$ be a parameterization of $D$ at $p$.

Definition 6.45. The map $\phi$ is an oriented parameterization of $D$ if it is an oriented parameterization of $X$.

Assume that $\operatorname{dim} X=n>1$. We show that you can always find an oriented parameterization.

Let $\phi: U \rightarrow V$ be a parameterization of $D$ at $p$. Suppose that $\phi$ is not oriented. That is, as a diffeomorphism $\phi$ is orientation reversing. Let $A: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ be the map

$$
\begin{equation*}
A\left(x_{1}, \ldots, x_{n}\right)=\left(x_{1}, \ldots, x_{n-1},-x_{n}\right) \tag{6.144}
\end{equation*}
$$

Then $A$ maps $\mathbb{H}^{n} \rightarrow \mathbb{H}^{n}$, and $\phi \circ A$ is orientation preserving. So, $\phi \circ A$ is an oriented parameterization of $D$ at $p$.

Now, let $\phi: U \rightarrow V$ be an oriented parameterization of $D$ at $p$. We define

$$
\begin{align*}
U^{b} & =U \cap \operatorname{Bd}\left(\mathbb{H}^{n}\right),  \tag{6.145}\\
V^{b} & =V \cap \operatorname{Bd}(D),  \tag{6.146}\\
\psi & =\phi \mid U^{b}, \tag{6.147}
\end{align*}
$$

where $\psi$ is a parameterization of $\operatorname{Bd}(D)$ at $p$.
We oriented $\operatorname{Bd}(D)$ at $p$ by requiring $\psi$ to be an oriented parameterization. We need to check the following claim.

Claim. The definition of oriented does not depend on the choice of parameterization. Proof. Let $\phi_{i}: U_{i} \rightarrow V_{i}, i=1,2$, be oriented parameterizations of $D$ at $p$. Define

$$
\begin{align*}
& U_{1,2}=\phi_{1}^{-1}\left(V_{1} \cap V_{2}\right),  \tag{6.148}\\
& U_{2,1}=\phi_{2}^{-1}\left(V_{1} \cap V_{2}\right), \tag{6.149}
\end{align*}
$$

from which we obtain the following diagram:

which defines a map $g$. By the properties of the other maps $\phi_{1}, \phi_{2}$, the map $g$ is an orientation preserving diffeomorphism of $U_{1,2}$ onto $U_{2,1}$. Moreover, $g$ maps

$$
\begin{equation*}
U_{1,2}^{b}=\operatorname{Bd}\left(\mathbb{H}^{n}\right) \cap U_{1,2} \tag{6.151}
\end{equation*}
$$

onto

$$
\begin{equation*}
U_{2,1}^{b}=\operatorname{Bd}\left(\mathbb{H}^{n}\right) \cap U_{2,1} . \tag{6.152}
\end{equation*}
$$

Let $h=g \mid U_{1,2}^{b}$, so $h: U_{1,2}^{b} \rightarrow U_{2,1}^{b}$. We want to show that $h$ is orientation preserving. To show this, we write $g$ and $h$ in terms of coordinates.

$$
\begin{equation*}
g=\left(g_{1}, \ldots, g_{n}\right), \quad \text { where } g_{i}=g_{i}\left(x_{1}, \ldots, x_{n}\right) \tag{6.153}
\end{equation*}
$$

So,

$$
g \text { maps } \mathbb{H}^{n} \text { to } \mathbb{H}^{n} \Longleftrightarrow \begin{cases}g_{1}\left(x_{1}, \ldots, x_{n}\right)<0 & \text { if } x_{1}<0  \tag{6.154}\\ g_{1}\left(x_{1}, \ldots, x_{n}\right)>0 & \text { if } x_{1}>0 \\ g_{1}\left(0, x_{2}, \ldots, x_{n}\right)=0 & \end{cases}
$$

These conditions imply that

$$
\left\{\begin{array}{l}
\frac{\partial}{\partial x_{1}} g_{1}\left(0, x_{2}, \ldots, x_{n}\right) \geq 0  \tag{6.155}\\
\frac{\partial}{\partial x_{i}} g_{1}\left(0, x_{2}, \ldots, x_{n}\right)=0, \text { for } i \neq 1
\end{array}\right.
$$

The map $h$ in coordinates is then

$$
\begin{align*}
h & =h\left(x_{2}, \ldots, x_{n}\right)  \tag{6.156}\\
& =\left(g\left(0, x_{2}, \ldots, x_{n}\right), \ldots, g_{n-1}\left(0, x_{2}, \ldots, x_{n}\right)\right),
\end{align*}
$$

which is the statement that $h=g \mid \operatorname{Bd}\left(\mathbb{H}^{n}\right)$.
At the point $\left(0, x_{2}, \ldots, x_{n}\right) \in U_{1,2}^{b}$,

$$
D g=\left[\begin{array}{cccc}
\frac{\partial g_{1}}{\partial x_{1}} & 0 & \cdots & 0  \tag{6.157}\\
* & & & \\
\vdots & & D h & \\
* & & &
\end{array}\right]
$$

The matrix $D g$ is an $n \times n$ block matrix containing the $(n-1) \times(n-1)$ matrix $D h$, because

$$
\begin{equation*}
\frac{\partial h_{i}}{\partial x_{j}}=\frac{\partial g_{i}}{\partial x_{j}}\left(0, x_{2}, \ldots, x_{n}\right), i, j>1 \tag{6.158}
\end{equation*}
$$

Note that

$$
\begin{equation*}
\operatorname{det}(D g)=\frac{\partial g_{1}}{\partial x_{1}} \operatorname{det}(D h) \tag{6.159}
\end{equation*}
$$

We know that the l.h.s $>0$ and that $\frac{\partial g_{1}}{\partial x_{1}}>0$, so $\operatorname{det}(D h)>0$. Thus, the map $h: U_{1,2}^{b} \rightarrow U_{2,1}^{b}$ is orientation preserving.

To repeat, we showed that in the following diagram, the map $h$ is orientation preserving:


We conclude that $\psi_{1}$ is orientation preserving if and only if $\psi_{2}$ is orientation preserving.

## Lecture 38

We begin with a review from last time.
Let $X$ be an oriented manifold, and let $D \subseteq X$ be a smooth domain. Then $\operatorname{Bd}(D)=Y$ is an oriented $(n-1)$-dimensional manifold.

We defined integration over $D$ as follows. For $\omega \in \Omega_{c}^{n}(X)$ we want to make sense of the integral

$$
\begin{equation*}
\int_{D} \omega . \tag{6.161}
\end{equation*}
$$

We look at some special cases:
Case 1: Let $p \in \operatorname{Int} D$, and let $\phi: U \rightarrow V$ be an oriented parameterization of $X$ at $p$, where $V \subseteq \operatorname{Int} D$. For $\omega \in \Omega_{c}^{n}(X)$, we define

$$
\begin{equation*}
\int_{D} \omega=\int_{V} \omega=\int_{U} \phi^{*} \omega=\int_{\mathbb{R}^{n}} \phi^{*} \omega . \tag{6.162}
\end{equation*}
$$

This is just our old definition for

$$
\begin{equation*}
\int_{V} \omega . \tag{6.163}
\end{equation*}
$$

Case 2: Let $p \in \operatorname{Bd}(D)$, and let $\phi: U \rightarrow V$ be an oriented parameterization of $D$ at $p$. That is, $\phi$ maps $U \cap \mathbb{H}^{n}$ onto $V \cap D$. For $\omega \in \Omega_{c}^{n}(V)$, we define

$$
\begin{equation*}
\int_{D} \omega=\int_{\mathbb{H}^{n}} \phi^{*} \omega . \tag{6.164}
\end{equation*}
$$

We showed last time that this definition does not depend on the choice of parameterization.

General case: For each $p \in \operatorname{Int} D$, let $\phi: U_{p} \rightarrow V_{p}$ be an oriented parameterization of $X$ at $p$ with $V_{p} \subseteq \operatorname{Int} D$. For each $p \in \operatorname{Bd}(D)$, let $\phi: U_{p} \rightarrow V_{p}$ be and oriented parameterization of $D$ at $p$. Let

$$
\begin{equation*}
U=\sum_{p \in D} U_{p}, \tag{6.165}
\end{equation*}
$$

where the set $\mathcal{U}=\left\{U_{p}: p \in D\right\}$ be an an open cover of $U$. Let $\rho_{i}, i=1,2, \ldots$, be a partition of unity subordinate to this cover.

Definition 6.46. For $\omega \in \Omega_{c}^{n}(X)$ we define the integral

$$
\begin{equation*}
\int_{D} \omega=\sum_{i} \int_{D} \rho_{i} \omega . \tag{6.166}
\end{equation*}
$$

Claim. The r.h.s. of this definition is well-defined.

Proof. Since the $\rho_{i}$ 's are a partition of unity, there exists an $N$ such that

$$
\begin{equation*}
\operatorname{supp} \omega \cap \operatorname{supp} \rho_{i}=\phi, \tag{6.167}
\end{equation*}
$$

for all $i>N$.
Hence, there are only a finite number of non-zero terms in the summand. Moreover, each summand is an integral of one of the two types above (cases 1 and 2), and is therefore well-defined.

Claim. The l.h.s. of the definition does not depend on the choice of the partition of unity $\rho_{i}$.

Proof. We proved an analogous assertion about the definition of $\int_{X} \omega$ a few lectures ago, and the proof of the present claim is exactly the same.

### 6.11 Stokes' Theorem

Stokes' Theorem. For all $\omega \in \Omega_{c}^{n-1}(X)$,

$$
\begin{equation*}
\int_{D} d \omega=\int_{\operatorname{Bd}(D)} \omega . \tag{6.168}
\end{equation*}
$$

Proof. Let $\rho_{i}, i=1,2 \ldots$, be a partition of unity as defined above. Replacing $\omega$ with $\sum \rho_{i} \omega$, it suffices to prove this for the two special cases below:

Case 1: Let $p \in \operatorname{Int} D$, and let $\phi: U \rightarrow V$ be an oriented parameterization of $X$ at $p$ with $V \subseteq \operatorname{Int} D$. If $\omega \in \Omega_{c}^{n-1}(V)$, then

$$
\begin{equation*}
\int_{D} d \omega=\int_{\mathbb{R}^{n}} \phi^{*} d \omega=\int_{\mathbb{R}^{n}} d \phi^{*} \omega=0 \tag{6.169}
\end{equation*}
$$

Case 2: Let $p \in \operatorname{Bd}(D)$, and let $\phi: U \rightarrow V$ be an oriented parameterization of $D$ at $p$. Let $U^{b}=U \cap \operatorname{Bd}\left(\mathbb{H}^{n}\right)$, and let $V^{b}=V \cap \operatorname{Bd}(D)$. Define $\psi: \phi \mid U^{b}$, so $\psi: U^{b} \rightarrow V^{b}$ is an oriented parameterization of $\operatorname{Bd}(D)$ at $p$. If $\omega \in \Omega_{c}^{n-1}(V)$, then

$$
\begin{equation*}
\phi^{*} \omega=\sum f_{i}\left(x_{1}, \ldots, x_{n}\right) d x_{1} \wedge \cdots \wedge \widehat{d x_{i}} \wedge \cdots \wedge d x_{n} \tag{6.170}
\end{equation*}
$$

What is $\psi^{*} \omega$ ? Let $\iota: \mathbb{R}^{n-1} \rightarrow \mathbb{R}^{n}$ be the inclusion map mapping $\operatorname{Bd}\left(\mathbb{H}^{n}\right) \rightarrow \mathbb{R}^{n}$. The inclusion map $\iota$ maps $\left(x_{2}, \ldots, x_{n}\right) \rightarrow\left(0, x_{2}, \ldots, x_{n}\right)$. Then $\phi \circ \iota=\psi$, so

$$
\begin{align*}
\psi^{*} \omega & =\iota^{*} \phi^{*} \omega \\
& =\iota^{*}\left(\sum_{i=1}^{n} f_{i} d x_{1} \wedge \cdots \wedge \widehat{d x_{i}} \wedge \cdots \wedge d x_{n}\right) . \tag{6.171}
\end{align*}
$$

But,

$$
\begin{equation*}
\iota^{*} d x_{1}=d \iota^{*} x_{1}=0, \quad \text { since } \iota^{*} x_{1}=0 \tag{6.172}
\end{equation*}
$$

So,

$$
\begin{align*}
\psi^{*} \omega & =\iota^{*} f_{1} d x_{2} \wedge \cdots \wedge d x_{n}  \tag{6.173}\\
& =f_{1}\left(0, x_{2}, \ldots, x_{n}\right) d x_{2} \wedge \cdots \wedge d x_{n}
\end{align*}
$$

Thus,

$$
\begin{equation*}
\int_{\operatorname{Bd}(D)} \omega=\int_{\mathbb{R}^{n-1}} \psi^{*} \omega=\int_{\mathbb{R}^{n-1}} f_{1}\left(0, x_{2}, \ldots, x_{n}\right) d x_{2} \ldots d x_{n} \tag{6.174}
\end{equation*}
$$

On the other hand,

$$
\begin{equation*}
\int_{D} d \omega=\int_{\mathbb{H}^{n}} \phi^{*} d \omega=\int_{\mathbb{H}^{n}} d \phi^{*} \omega . \tag{6.175}
\end{equation*}
$$

One should check that

$$
\begin{align*}
d \phi^{*} \omega & =d\left(\sum f_{i} d x_{1} \wedge \cdots \wedge \widehat{x x_{i}} \wedge \cdots \wedge d x_{n}\right) \\
& =\left(\sum(-1)^{i-1} \frac{\partial f_{i}}{\partial x_{i}}\right) d x_{1} \wedge \cdots \wedge d x_{n} . \tag{6.176}
\end{align*}
$$

So, each summand

$$
\begin{equation*}
\int \frac{\partial f_{i}}{\partial x_{i}} d x_{1} \ldots d x_{n} \tag{6.177}
\end{equation*}
$$

can be integrated by parts, integrating first w.r.t. the $i$ th variable. For $i>1$, this is the integral

$$
\begin{align*}
\int_{-\infty}^{\infty} \frac{\partial f_{i}}{\partial x_{i}} d x_{i} & =\left.f_{i}\left(x_{1}, \ldots, x_{n}\right)\right|_{\substack{x_{i}=\infty \\
x_{i}=-\infty}} ^{\substack{2}}  \tag{6.178}\\
& =0
\end{align*}
$$

For $i=1$, this is the integral

$$
\begin{equation*}
\int_{-\infty}^{\infty} \frac{\partial f_{1}}{\partial x_{1}}\left(x_{1}, \ldots, x_{n}\right) d x_{1}=f_{1}\left(0, x_{2}, \ldots, x_{n}\right) \tag{6.179}
\end{equation*}
$$

Thus, the total integral of $\phi^{*} d \omega$ over $\mathbb{H}^{n}$ is

$$
\begin{equation*}
\int f_{1}\left(0, x_{2}, \ldots, x_{n}\right) d x_{2} \ldots d x_{n} \tag{6.180}
\end{equation*}
$$

We conclude that

$$
\begin{equation*}
\int_{D} d \omega=\int_{\operatorname{Bd}(D)} \omega \tag{6.181}
\end{equation*}
$$

We look at some applications of Stokes' Theorem.
Let $D$ be a smooth domain. Assume that $D$ is compact and oriented, and let $Y=\operatorname{Bd}(D)$. Let $Z$ be an oriented $n$-manifold, and let $f: Y \rightarrow Z$ be a $\mathcal{C}^{\infty}$ map.
Theorem 6.47. If $f$ extends to $a \mathcal{C}^{\infty}$ map $F: D \rightarrow Z$, then

$$
\begin{equation*}
\operatorname{deg}(f)=0 \tag{6.182}
\end{equation*}
$$

Corollary 9. The Brouwer fixed point theorem follows from the above theorem.

