# NOTES ON MANIFOLDS 

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## 1. Introduction and Overview

Differentiable manifolds are objects upon which one may do calculus without coordinates. They abstract "differential structure" just as vector spaces abstract linear structure. In both of these settings, the fundamental example is ordinary Euclidean space. Manifolds are much more complicated, though, as one might expect. The only difference between $\mathbb{R}^{n}$ and an $n$-dimensional vector space (over $\mathbb{R}$ ) in terms of linear structure is a choice of basis. Therefore, the natural numbers effectively classify vector spaces. In contrast, the classification of manifolds is a rich subject with many open problems.

Whereas an $n$-dimensional vector space is isomorphic to $\mathbb{R}^{n}$, a manifold is by its very definition locally isomorphic to an open subset of $\mathbb{R}^{n}$. That property is in accordance with what you know about differentiation in $\mathbb{R}^{n}$. It is a local process: its resulting value at any point only depends on the behavior of the function in question in any small neighborhood of the point. The structure of a manifold includes instructions for gluing together this local information.

There is no way to make measurements of distances and angles in a vector space until we add an inner product. An $n$-dimensional manifold $M$ comes with a tangent bundle $T M$. It attaches a copy of $\mathbb{R}^{n}$ to each point (the tangent space at that point), and has the structure of a manifold, itself. In order to make measurements on a manifold, we need the additional structure of an inner product on the tangent space at each point, which varies smoothly with the point, resulting in a metric. Thus, our notion of distance will vary from point-to-point on a manifold. To see the utility of this notion, recall that mass distorts distances in our universe. A manifold with a metric is called a Riemannian manifold.

First example. We now describe the manifold structure of a two-dimensional sphere. While reading the following, please refer to Figure 1. Imagine the sphere as the surface of the earth. To find your way around on the earth, it suffices to have a world atlas. Each page $h$ of the atlas is a chart representing some portion of the earth. The page is essentially a mapping $h: U \rightarrow U^{\prime}$ from some region $U$ of the earth to a rectangle $U^{\prime}$ in $\mathbb{R}^{2}$. Using the features of the earth, one can infer the nature of $h$ by just looking at its image, i.e., the actual page in the atlas. From now on, we will identify these two things. Let $k: V \rightarrow V^{\prime}$ be another chart/page. Suppose the regions $U$ and $V$ on the earth meet with overlap $W:=U \cap V$. That means images of $W$ will appear on the two pages we are considering. Since drawing the earth on flat paper requires stretching, these two images will be distorted copies of each other. However, assuming your atlas has sufficient details, we can tell which points on the two images represent the same point on the earth.

Imagine now that the pages or your atlas are made of malleable putty, and your job is to reconstruct the earth. You cut out all the pages and are left with stretching and gluing together the overlaps. The likely result, given the nature of putty, will be a lop-sided lumpy version of a sphere.

In extracting the manifold structure from this example, it is important to remember that a manifold does not come with a metric. So we should try to forget


Figure 1. Charts $(U, h)$ and $(V, k)$ on the sphere and their corresponding transition mapping on the overlap.
that aspect of an ordinary atlas of the earth. Instead, as above, think of the underlying substrate of a manifold as putty. In mathematical terms, the pages of the atlas are open subsets of a topological space. We will assume the reader can quickly "review" the features of topology summarized in Appendix B.

One last observation: unlike our example, a manifold does not come with an embedding into Euclidean space - the embedding is separate information. Whitney's embedding theorem says that, in general, an $n$-dimensional manifold can be embedded in $\mathbb{R}^{2 n}$. So surfaces (two-dimensional manifolds) can be embedded in $\mathbb{R}^{4}$, but something special needs to occur to get an embedding in $\mathbb{R}^{3}$, as in the case of a sphere. The Klein bottle, which cannot be embedded in $\mathbb{R}^{3}$, is more typical.

Integration. While differentiation is a local process, integration is a global process. Thinking of our manifold as being created by gluing together stretchable pieces of putty, it is clear that integration will require adding some structure to the manifold. Having chosen coordinates near a point, we can integrate functions as usual, but what if we change coordinates? Of course, there is the usual change of coordinates formula from several-variable calculus, but how would we choose the initial set of coordinates (from which we could change)? And even more perplexing: what if we want to integrate over a portion of the manifold that is not contained in a single chart? These questions suggest that in order to perform integration on a manifold, we will need a structure that oversees these changes of coordinates and provides a standard of some sort, against which we can measure. That structure is called a non-vanishing $n$-form and will take us a while to define. Once we have it, though,
one of our goals will be to prove the ultimate version of Stokes' theorem:

$$
\int_{\partial M} \omega=\int_{M} d \omega .
$$

Here, $\omega$ is an $n$-form and $\partial M$ is the "boundary" of $M$ (so we will need to consider manifolds with boundaries). The operator $d$ is called the exterior derivative. It will have the property that applying $d$ twice gives $d^{2}=0$, which may remind you of some results from ordinary vector calculus in $\mathbb{R}^{n}$. Exploring this property leads to remarkable topological invariants in the form of cohomology groups. If time permits, we will consider these groups in the case of two classes of manifolds: toric varieties and Grassmannians. The cohomology of Grassmann manifolds has a ring structure (i.e., well-behaved addition and multiplication) known as the Schubert calculus.

## 2. Definition of a Manifold

Definition 2.1 (Charts). Let $X$ be a topological space. An $n$-dimensional chart on $X$ is a homeomorphism $h: U \xrightarrow{\cong} U^{\prime}$ from an open subset $U \subseteq X$, the chart domain, onto an open subset $U^{\prime} \subseteq \mathbb{R}^{n}$. We denote this chart by ( $U, h$ ) (see Figure 2). We say that $X$ is locally Euclidean if every point in $X$ belongs to some chart domain of $X$. If $X$ is locally Euclidean, choosing a chart containing some point $p \in X$ is called taking local coordinates at $p$.


Figure 2. A chart.

Definition 2.2. If $(U, h)$ and $(V, k)$ are two $n$-dimensional charts on $X$ such that $U \cap V \neq \emptyset$, then the homeomorphism $\left.\left(k \circ h^{-1}\right)\right|_{h_{(U \cap V)}}$ from $h(U \cap V)$ to $k(U \cap V)$ is called the change-of-charts map, change of coordinates, or transition map, from $h$ to $k$ (see Figure 3). If the transition map is furthermore a diffeomorphism (note that


Figure 3. Transition map.
these maps have domains and codomains in $\mathbb{R}^{n}$ ), then we say that the two charts are differentiably related. Recall that a function between subsets of Euclidean spaces is a diffeomorphism if it is bijective and both it and its inverse are differentiable. Throughout this text, we will take differentiable to mean smooth, i.e., having partial derivatives of all orders.

Definition 2.3. A set of $n$-dimensional charts on $X$ whose chart domains cover all of $X$ is an $n$-dimensional atlas on $X$. The atlas is differentiable if all its charts are differentiably related, and two differentiable atlases $\mathfrak{A}$ and $\mathfrak{B}$ are equivalent if $\mathfrak{A} \cup \mathfrak{B}$ is also differentiable.

Definition 2.4. An $n$-dimensional differentiable structure on a topological space $X$ is a maximal $n$-dimensional differentiable atlas with respect to inclusion.

Definition 2.5 (Differentiable manifolds). An $n$-dimensional differentiable manifold ${ }^{1}$ is a pair $(M, \mathcal{D})$ consisting of a second countable Hausdorff topological space $M$ with an $n$-dimensional differentiable structure $\mathcal{D}$.

Example 2.6. As a first (trivial) example of an $n$-manifold, take any open subset $U \subseteq \mathbb{R}^{n}$ with the atlas $\left\{\left(U, \mathrm{id}_{U}\right)\right\}$ containing a single chart.

Example 2.7 (The $n$-sphere). The $n$-sphere is the set

$$
S^{n}:=\left\{x \in \mathbb{R}^{n+1}| | x \mid=1\right\}
$$

Thus, for example, the one-sphere is the unit circle in the plane, and the twosphere is the usual sphere in three-space. If $p=\left(p_{1}, \ldots, p_{n+1}\right) \in S^{n}$, then there exists some $i$ such that $p_{i} \neq 0$. Let $U$ be any open neighborhood of $p$ consisting solely of points on the sphere whose $i$-coordinates are nonzero. Then the mapping at $h: U \rightarrow h(U) \subset \mathbb{R}^{n}$ defined by dropping the $i$-th coordinate of each point in $U$ serves as a chart containing $p$.

For another atlas, compatible with the one just given, one can use stereographic projection. Here the atlas as two open sets:

$$
U^{+}:=S^{n} \backslash\{(0,0, \ldots, 0,1)\} \quad \text { and } \quad U^{-}:=S^{n} \backslash\{(0,0, \ldots, 0,-1)\}
$$

each with chart defined by

$$
\left(x_{1}, \ldots, x_{n+1}\right) \mapsto\left(\frac{x_{1}}{1-x_{n+1}}, \ldots, \frac{x_{n}}{1-x_{n+1}}\right)
$$

2.1. Projective space. Projective $n$-space, denoted $\mathbb{P}^{n}$, is an $n$-manifold whose points are the lines in $\mathbb{R}^{n+1}$ passing through the origin, i.e., the collection of all one-dimensional linear subspaces of $\mathbb{R}^{n+1}$. To represent a line $\ell$ through the origin in $\mathbb{R}^{n+1}$, choose any nonzero point $p \in \ell$. (The point $p$ is a basis for $\ell$ as a onedimensional linear subspace.) For points $p, q \in \mathbb{R}^{n+1}$ write $p \sim q$ if $p=\lambda q$ for some nonzero constant $\lambda$. Then $p$ and $q$ represent the same line if and only if $p \sim q$. So we take our formal definition of projective $n$-space to be

$$
\mathbb{P}^{n}:=\mathbb{R}^{n+1} \backslash\{\overrightarrow{\boldsymbol{0}}\} / \sim
$$

with the quotient topology. (Thus, a set of points in $\mathbb{P}^{n}$ is open if and only if the union of the set of lines they represent is an open subset of $\mathbb{R}^{n+1} \backslash\{\overrightarrow{\boldsymbol{0}}\}$.)

If $\ell \in \mathbb{P}^{n}$ is represented by (the equivalence class of) $p=\left(a_{1}, \ldots, a_{n+1}\right) \in \ell$, then $\left(a_{1}, \ldots, a_{n+1}\right)$ are called the homogeneous coordinates for $\ell$, realizing that these "coordinates" are only defined up to scaling by a nonzero constant. One sometimes sees the notation $\left(a_{1}: a_{2}: \cdots: a_{n+1}\right)$ to represent the point $\ell \in \mathbb{P}^{n}$, but we will stick with $\left(a_{1}, \ldots, a_{n+1}\right)$.

We would now like to impose a manifold structure on $\mathbb{P}^{n}$. As a warm-up, we first treat the case $n=2$. Define the set

$$
\begin{aligned}
& U_{x}:=\left\{(x, y, z) \in \mathbb{P}^{2}: x \neq 0\right\} \\
& U_{y}:=\left\{(x, y, z) \in \mathbb{P}^{2}: y \neq 0\right\}
\end{aligned}
$$

[^1]$$
U_{z}:=\left\{(x, y, z) \in \mathbb{P}^{2}: z \neq 0\right\}
$$

Note the following:
(1) These sets are well-defined. Even though homogeneous coordinates are only defined up to a nonzero scalar constant. Nevertheless, replacing $(x, y, z)$ by $\lambda(x, y, z)$ with $\lambda \neq 0$, we have $x=0$ if and only if $\lambda x=0$, and similarly for $y$ and $z$.
(2) These are open sets. Since we are using the quotient topology, we need to consider $\pi^{-1}\left(U_{x}\right)$ which is the complement of the $y, z$ plane in $\mathbb{R}^{3} \backslash\{\overrightarrow{\boldsymbol{0}}\}$, which is open in $\mathbb{R}^{3} \backslash\{\overrightarrow{\mathbf{0}}\}$. A similar argument holds for $U_{y}$ and $U_{z}$.
(3) These sets cover $\mathbb{P}^{2}$. If $(x, y, z) \in \mathbb{P}^{2}$, then at least one of $x, y$, or $z$ must be nonzero.

Finally, we define the chart mappings:

$$
\begin{aligned}
& \phi_{x}: U_{x} \rightarrow \mathbb{R}^{2} \quad \phi_{y}: U_{y} \rightarrow \mathbb{R}^{2} \quad \phi_{z}: U_{z} \rightarrow \mathbb{R}^{2} \\
& (x, y, z) \mapsto(y / x, z / x) \quad(x, y, z) \mapsto(x / y, z / y) \quad(x, y, z) \mapsto(x / z, y / z)
\end{aligned}
$$

It is easy to check that each of these is a homeomorphism. For instance, the inverse of $\phi_{x}$ is given by $(u, v) \mapsto(1, u, v)$. The underlying motivation for these charts is as follows: Take $U_{x}$, for instance. Fix the plane $x=1$ in $\mathbb{R}^{3}$. Then a line through the origin in $\mathbb{R}^{3}$ meets this plane if and only if a representative nonzero point on the line has $x$-coordinate not equal to 0 . So $U_{x}$ consists of the lines meeting the plane $x=1$. If $\ell \in U_{x}$ has homogeneous coordinates $(x, y, z)$, we can scale $(x, y, z)$ by $\lambda=1 / x$ to get another representative $(1, y / x, z / x)$. This is exactly the point where $\ell$ meets the plane $x=1$. In this way, each point in $U_{x}$ has a set of homogeneous coordinates of the form $(1, u, v)$. Dropping the 1 , which is superfluous information, gives us the mapping $\phi_{x}$. The essential idea is that there is a one-to-one correspondence between points in the plane $x=1$ and points in $U_{x}$, and the plane $x=1$ is the same as $\mathbb{R}^{2}$ via $(1, u, v) \mapsto(u, v)$.

The collection

$$
\left\{\left(U_{x}, \phi_{x}\right),\left(U_{y}, \phi_{y}\right),\left(U_{z}, \phi_{z}\right)\right\}
$$

is the standard atlas for $\mathbb{P}^{2}$. What does a typical transition function look like? Consider the transition from $U_{x}$ to $U_{y}$. The overlap is $U_{x} \cap U_{y}=\left\{(x, y, z) \in \mathbb{P}^{2}: x \neq 0, y \neq 0\right\}$, and we have the commutative diagram:


In general, the standard charts of $\mathbb{P}^{n}$ are given by $\left(U_{i}, \phi_{i}\right)$ for $i=1, \ldots, n+1$, where $U_{i}=\left\{x=\left(x_{1}, \ldots, x_{n+1}\right) \in \mathbb{P}^{n} \mid x_{i} \neq 0\right\}$, and

$$
\begin{aligned}
\phi_{i}: U_{i} & \rightarrow \mathbb{R}^{n} \\
x & \mapsto\left(\frac{x_{1}}{x_{i}}, \ldots, \frac{\widehat{x_{i}}}{x_{i}}, \ldots, \frac{x_{n+1}}{x_{i}}\right) .
\end{aligned}
$$

The hat over $\frac{x_{i}}{x_{i}}$ means that this component of the vector should be omitted (just like we $\operatorname{did}$ for $\mathbb{P}^{2}$ ).

## 3. Differentiable Maps

Let $M$ be a manifold and $X$ some topological space. To study the behavior of a map $f: M \rightarrow X$ near some point $p \in M$, we choose a chart $(U, h)$ at $p$ and look at the "downstairs map" $f \circ h^{-1}: h(U) \rightarrow X$ instead (see Figure 4). If $f \circ h^{-1}$ has a


Figure 4. The downstairs map $f \circ h^{-1}$.
certain property locally at $h(p)$, then we say that $f$ has the property at $p$ relative to the chart $(U, h)$. If this property is independent of the choice of charts, then we just say that $f$ has this property at $p$. Our first example of such a property is the differentiability of a function.

Definition 3.1 (Real-valued functions on $M$ ). A function $f: M \rightarrow \mathbb{R}$ is differentiable at $p \in M$ if $f \circ h^{-1}$ is differentiable for some chart $(U, h)$ at $p$.
Exercise 3.2. Let $f: M \rightarrow \mathbb{R}$ be a function on a manifold $M$, and let $(U, h)$ and $(V, k)$ be two charts at $p \in M$. Show that if $f$ is differentiable at $p$ relative to $(U, h)$, then $f$ is differentiable at $p$ relative to $(V, k)$. (You can use the fact that a composition of differentiable functions on Euclidean space is differentiable.)

Definition 3.3 (Differentiable mappings of manifolds). A continuous map $f: M \rightarrow$ $N$ between manifolds is differentiable at $p \in M$ if it is differentiable with respect to charts at $p \in M$ and at $f(p) \in N$. Namely, if $(U, h)$ is a chart at $p$ and $(V, k)$ is a chart at $f(p)$ such that $f(U) \subseteq V$, we want the map $k \circ f \circ h^{-1}$ to be differentiable (recall that $h(U) \subseteq \mathbb{R}^{m}$ and $k(V) \subseteq \mathbb{R}^{n}$ for some $m, n$ ):


For a picture of this,see Figure 5. If $f$ is bijective with a differentiable inverse, then $f$ is called a diffeomorphism.

Remark 3.4. The reader should check that differentiability at $p \in M$ is independent of the choice of charts.

Example 3.5 (The Veronese embedding). Define $\nu_{2}: \mathbb{P}^{2} \rightarrow \mathbb{P}^{5}$ by

$$
\begin{aligned}
\nu_{2}: \mathbb{P}^{2} & \rightarrow \mathbb{P}^{5} \\
(x, y, z) & \mapsto\left(x^{2}, x y, x z, y^{2}, y z, z^{2}\right)
\end{aligned}
$$



Figure 5. Using the charts to pull down a continuous map between manifolds.

This mapping is well-defined since

$$
\nu_{2}(\lambda(x, y, z))=\nu_{2}(\lambda x, \lambda y, \lambda z)=\lambda^{2}\left(x^{2}, x y, x z, y^{2}, y z, z^{2}\right)=\nu_{2}(x, y, z)
$$

We would like to show that $\nu_{2}$ is differentiable. Let $p=(a, b, c) \in \mathbb{P}^{2}$, and without loss of generality, suppose that $a \neq 0$. Let $\left(U_{x}, \phi_{x}\right)$ be the standard chart at $p$ whose domain consists of points with nonzero first coordinates. Then $\nu_{2}(p) \in \nu_{2}\left(U_{x}\right) \subset$ $U_{1} \subset \mathbb{P}^{5}$, and so we take the standard chart $\left(U_{1}, \phi_{1}\right)$ at $\nu_{2}(p)$. Using these charts, we calculate

$$
\phi_{1} \circ \nu_{2} \circ \phi_{x}^{-1}(u, v)=\phi_{1}\left(\nu_{2}(1, u, v)\right)=\phi_{1}\left(1, u, v, u^{2}, u v, v^{2}\right)=\left(u, v, u^{2}, u v, v^{2}\right)
$$

which is differentiable.
Example 3.6. The existence of a diffeomorphism between two manifolds means that as far as manifold structures are concerned, there is no difference between the manifolds (except for, possibly, their names). Thinking in those terms, the following theorem may be surprising:

Theorem 3.7 (Milnor). There exist differentiable structures $\mathcal{D}$ and $\mathcal{D}^{\prime}$ on the seven-sphere $S^{7}$ with no diffeomorphism $\left(S^{7}, \mathcal{D}\right) \rightarrow\left(S^{7}, \mathcal{D}^{\prime}\right)$.

## 4. Tangent Spaces

If we try to imagine a typical tangent space, we might think of a surface $S$ sitting inside $\mathbb{R}^{3}$ with a plane "kissing" the surface at some point. To compute the tangent plane, we could create a parametrization of the surface near that point. This would amount to finding an open set $U \subseteq \mathbb{R}^{2}$ and a "nice" ${ }^{2}$ differentiable mapping $f: U \rightarrow \mathbb{R}^{3}$ with a point $p \in U$ mapping to the point in question, $f(p)$. Recall that by definition,

$$
\lim _{|h| \rightarrow 0} \frac{\left|f(p+h)-f(p)-D f_{p}(h)\right|}{|h|}=0
$$

where $D f_{p}: \mathbb{R}^{2} \rightarrow \mathbb{R}^{3}$ is the derivative of $f$ at $p$. Hence, up to first order, we have $f(p+h)=f(p)+D f_{p}(h)$ where $D f_{p}$ is the derivative of $f$ at $p$. Letting $A f_{p}(h):=f(p+h)$, we get the best affine approximation to $f$ at $p$ :

$$
\begin{aligned}
A f_{p}: \mathbb{R}^{2} & \rightarrow \mathbb{R}^{3} \\
h & \mapsto f(p)+D f_{p}(h)
\end{aligned}
$$

Its image will be the tangent plane at $f(p)$.
An underlying assumption we made above in thinking about the tangent space is that our surface and its tangent plane are embedded in a larger space, $\mathbb{R}^{3}$. Our goal in this section is to take on the intriguing task of creating an intrinsic definition of the tangent space of a manifold at a given point, i.e., one that does not depend on embedding the manifold into another space. We will give three constructions of the tangent space which we will call the geometric, the algebraic, and the physical definitions of the tangent space, and we will see that all three are equivalent. ${ }^{3}$

At a couple of places in the following discussion, we will need at technical lemma. We'll get that out of the way now:

Lemma 4.1. Let $f: W \rightarrow \mathbb{R}$ be a smooth function on some open subset $W$ of $\mathbb{R}^{n}$, and let $w \in W$. Then there exist smooth functions $g_{i}: W \rightarrow \mathbb{R}$ for $i=1, \ldots, n$ satisfying:

$$
g_{i}(p)=\frac{\partial f}{\partial x_{i}}(p) \quad \text { and } \quad f(x)=f(p)+\sum_{i=1}^{n} g_{i}(x)\left(x_{i}-p_{i}\right)
$$

Proof. For $x \in W$, apply the fundamental theorem of calculus and then the chain rule to get

$$
\begin{aligned}
f(x)-f(p) & =\int_{0}^{1} \frac{d}{d t} f(t x+(1-t) p) d t \\
& =\int_{0}^{1} \sum_{i=1}^{n}\left[\frac{\partial f}{\partial x_{i}}(t x+(1-t) p)\right]\left(x_{i}-p_{i}\right) d t \\
& =\sum_{i=1}^{n}\left[\int_{0}^{1} \frac{\partial f}{\partial x_{i}}(t x+(1-t) p) d t\right]\left(x_{i}-p_{i}\right) .
\end{aligned}
$$

[^2]Define

$$
g_{i}(x)=\int_{0}^{1} \frac{\partial f}{\partial x_{i}}(t x+(1-t) p) d t
$$

For the remaining of this section, let $M$ be a manifold and take $p \in M$.

### 4.1. Three definitions of the tangent space.

Definition 4.2 (Geometrically-defined tangent space). Let $\mathcal{K}_{p}(M)$ denote the set of differentiable curves in $M$ that pass through $p$ at 0 . More precisely,

$$
\mathcal{K}_{p}(M)=\{\alpha:(-\varepsilon, \varepsilon) \rightarrow M \mid \alpha \text { is differentiable, } \varepsilon>0, \text { and } \alpha(0)=p\}
$$

Two such curves $\alpha, \beta \in \mathcal{K}_{p}(M)$ are called tangentially equivalent, denoted $\alpha \sim \beta$, if

$$
(h \circ \alpha)^{\prime}(0)=(h \circ \beta)^{\prime}(0) \in \mathbb{R}^{n}
$$

for some (hence any) chart $(U, h)$ around $p$. We call the equivalence classes $\mathcal{K}_{p}(M)$ the (geometrically-defined) tangent vectors of $M$ at $p$ and call the quotient

$$
T_{p}^{\text {geom }} M:=\mathcal{K}_{p}(M) / \sim
$$

the (geometrically-defined) tangent space to $M$ at $p$.
Exercise 4.3. Check that $\sim$ is independent of the choice of chart $(U, h)$.
Now let us put a linear structure on $T_{p}^{\mathrm{eeom}} M$. Fix a chart $(U, h)$ at $p$ and define

$$
\begin{aligned}
\phi: T_{p}^{\text {geom }} M & \rightarrow \mathbb{R}^{n} \\
{[\alpha] } & \mapsto(h \circ \alpha)^{\prime}(0) .
\end{aligned}
$$

Then $\phi$ is a bijection (check!). So we can use it to induce a linear structure on $T_{p}^{\text {geom }} M$. For curves $\alpha, \beta \in \mathcal{K}_{p}(M)$ and any real number $\lambda$, define

$$
\lambda[\alpha]+[\beta]:=\phi^{-1}(\lambda \phi(\alpha)+\phi(\beta))
$$

The result is independent of choice of chart: Let $(V, k)$ be another chart at $p$ and similarly define $\psi: T_{p}^{\text {geom }} M \rightarrow \mathbb{R}^{n}$ that sends $[\alpha]$ to $(k \circ \alpha)^{\prime}(0)$. We need to show that for $\alpha, \beta \in \mathcal{K}_{p}(M)$, we have

$$
\phi^{-1}(\lambda \phi(\alpha)+\phi(\beta))=\psi^{-1}(\lambda \psi(\alpha)+\psi(\beta)) .
$$

Pick $\gamma \in \mathcal{K}_{p}(M)$ such that $[\gamma]=\phi^{-1}(\lambda \phi(\alpha)+\phi(\beta))$. Equivalently, applying the bijection $\phi$ to both sides of this equation,

$$
(h \circ \gamma)^{\prime}(0)=\lambda \phi(\alpha)+\phi(\beta)
$$

Our goal is to show that $[\gamma]=\psi^{-1}(\lambda \psi(\alpha)+\psi(\beta))$, i.e., that

$$
(k \circ \gamma)^{\prime}(0)=\lambda \psi(\alpha)+\psi(\beta)
$$

This follows from the chain rule ${ }^{4}$ :

$$
(k \circ \gamma)^{\prime}(0)=J(k \circ \gamma)(0)
$$

[^3]\[

$$
\begin{aligned}
& =J\left(k \circ h^{-1} \circ h \circ \gamma\right)(0) \\
& =J\left(k \circ h^{-1}\right)(h(p)) J(h \circ \gamma)(0) \\
& =J\left(k \circ h^{-1}\right)(h(p))(h \circ \gamma)^{\prime}(0) \\
& =J\left(k \circ h^{-1}\right)(h(p))(\lambda \phi(\alpha)+\phi(\beta)) \\
& =\lambda J\left(k \circ h^{-1}\right)(h(p)) J(h \circ \alpha)(0)+J\left(k \circ h^{-1}\right)(h(p)) J(h \circ \beta)(0) \\
& =\lambda J\left(k \circ h^{-1} \circ h \circ \alpha\right)(0)+J\left(k \circ h^{-1} \circ h \circ \beta\right)(0) \\
& =\lambda(k \circ \alpha)^{\prime}(0)+(k \circ \beta)^{\prime}(0) \\
& =\lambda \psi(\alpha)+\psi(\beta) .
\end{aligned}
$$
\]

By definition of the linear structure on $T_{p}^{\text {geom }}(M)$, the mapping $\phi$ is an isomorphism of vector spaces. Thus, $\operatorname{dim} T_{p}^{\text {geom }}(M)=n$. The zero vector for $T_{p}^{\text {geom }}(M)$ is the class of the constant curve at $t \mapsto p$ for all $t \in(-1,1)$.

We now move on to the algebraic version of the tangent space. For a moment, let's think about ordinary vector calculus in $\mathbb{R}^{n}$, i.e., the case $M=\mathbb{R}^{n}$. Let $p \in \mathbb{R}^{n}$, and take a vector $v \in \mathbb{R}^{n}$. Think of $v$ as a tangent vector at $p$. Then given a function $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$, we could ask how fast $f$ is changing along $v$. In other words we could compute

$$
f_{v}(p):=\lim _{t \rightarrow 0} \frac{f(p+t v)-f(p)}{t}
$$

[Check that $f_{v}(p)=\nabla f(p) \cdot v$.] If $v$ is a unit vector, this would be the directional derivative of $f$ in the direction of $v$. What are the main properties of the derivative? For one, the derivative should be a linear function. In addition, its main property is the product rule (also know as the Leibniz rule).

We want to use this idea of a directional derivative to define tangent vectors on an arbitrary manifold. (It does not carry over verbatim since, for instance, $p+t v$ would not make sense when $p$ is a point in a general manifold.) The idea is that a tangent vector at a point $p$ should give us a kind of derivative of real-valued functions defined near $p$. (The function does not need to be defined on the whole manifold because, after all, the derivative at a point is a local property.) Thus, we are led to the idea that a tangent vector at $p$ should be a linear function on the space of functions defined near $p$, and that linear function should satisfy the product rule. We now make these notions precise.
Definition 4.4. Let two real-valued functions, each defined and differentiable in some neighborhood of a point $p$ of $M$, be called equivalent if they agree in a neighborhood of $p$. The equivalence classes are called the germs of differentiable functions on $M$ at $p$, and the set of these germs is denoted by $\mathcal{E}_{p}(M)$.

Remark 4.5. Note that equivalent functions need not be the same on the entire intersection of their domains, as illustrated in Figure 6.

Remark 4.6. The collection of germs $\mathcal{E}_{p}(M)$ is more than just a vector space-it is an $\mathbb{R}$-algebra! Namely, for $[f],[g] \in \mathcal{E}_{p}(M)$, apart from the usual operations of addition and scaling by $\mathbb{R}$, we can also multiply: $[f] \cdot[g]:=[f g]$. (Check that this multiplication is well-defined.)


Figure 6. Equivalent functions do not need to agree on the entirety of their common domains.

Definition 4.7 (Algebraically-defined tangent space). By an (algebraically-defined) tangent vector to $M$ at $p$, we mean a derivation of the ring $\mathcal{E}_{p}(M)$ of germs, that is, a linear map on the germs

$$
v: \mathcal{E}_{p}(M) \rightarrow \mathbb{R}
$$

that satisfies the product rule

$$
v(f \cdot g)=v(f) \cdot g(p)+f(p) \cdot v(g)
$$

for all $f, g \in \mathcal{E}_{p}(M)$. We call the vector space of these derivations the (algebraicallydefined) tangent space to $M$ at $p$ and denote it by $T_{p}^{\mathrm{alg}}(M)$.

Exercise 4.8. Let $f: M \rightarrow \mathbb{R}$ be a constant function. Let $v$ be a derivation of $\mathcal{E}_{p}(M)$. What is $v(f)$ ?

We would now like to describe a basis for $T_{p}^{\text {alg }}(M)$. Fix a chart $(U, h)$ at $p$ and define the derivations $\partial_{i}$ for $i=1, \ldots, n$ by

$$
\partial_{i}(f):=\frac{\partial}{\partial x_{i}}\left(f \circ h^{-1}\right)(h(p))
$$

for each $f \in \mathcal{E}_{p}(M)$. In other words, we use the chart $(U, h)$ to identify $f$ with an ordinary multivariable function, and then take its $i$-th partial derivative. We leave the straightforward check that each $\partial_{i}$ is a derivation to the reader. The reader may also check that

$$
\partial_{i}\left(h_{j}\right)= \begin{cases}1 & \text { if } i=j \\ 0 & \text { otherwise }\end{cases}
$$

We claim that these $\partial_{i}$ form a basis for $T^{\text {alg }}$. By the above displayed equation, they are linearly independent. To see they span is not so easy. So the reader may want to put off the following argument until well-rested! Take $v \in T^{\text {alg }}$, and let $f$ be a smooth real-valued function defined near $p$. Define $\ell: h(U) \rightarrow \mathbb{R}$ by $\ell:=f \circ h^{-1}$. By Lemma 4.1,

$$
\begin{aligned}
\ell(x) & =\ell(h(p))+\sum_{i=1}^{n} g_{i}(x)\left(x_{i}-h(p)_{i}\right) \\
& =f(p)+\sum_{i=1}^{n} g_{i}(x)\left(x_{i}-h(p)_{i}\right)
\end{aligned}
$$

for some smooth functions $g_{i}$ such that

$$
g_{i}(h(p))=\frac{\partial \ell}{\partial x_{i}}(h(p))=\frac{\partial}{\partial x_{i}}\left(f \circ h^{-1}\right)(h(p)) .
$$

Therefore,

$$
f(x)=\ell(h(x))=f(p)+\sum_{i=1}^{n} g_{i}(h(x))\left(h(x)_{i}-h(p)_{i}\right)
$$

Using the fact that $v$ is linear and satisfies the product rule, we get

$$
\begin{aligned}
v(f) & =v\left(f(p)+\sum_{i=1}^{n} g_{i}(h(x))\left(h(x)_{i}-h(p)_{i}\right)\right) \\
& =v(f(p))+\sum_{i=1}^{n} v\left(g_{i}(h(x))\left(h(x)_{i}-h(p)_{i}\right)\right) \\
& =\sum_{i=1}^{n} v\left(g_{i}(h(x))\left(h(x)_{i}-h(p)_{i}\right)\right) \\
& =\sum_{i=1}^{n} g_{i}(h(p)) v\left(h(x)_{i}-h(p)_{i}\right)+v\left(g_{i}(h(x))\right) \cdot 0 \\
& =\sum_{i=1}^{n} g_{i}(h(p)) v\left(h(x)_{i}\right)
\end{aligned}
$$

Letting $\alpha_{i}:=v\left(h(x)_{i}\right) \in \mathbb{R}$, we have

$$
v(f)=\sum_{i=1}^{n} \alpha_{i} g_{i}(h(p))=\sum_{i=1}^{n} \alpha_{i} \frac{\partial}{\partial x_{i}}\left(f \circ h^{-1}\right)(h(p))=\sum_{i=1}^{n} \alpha_{i} \partial_{i}(f)
$$

Thus,

$$
v=\sum_{i=1}^{n} \alpha_{i} \partial_{i}
$$

as required.
For our last formulation of tangent space, we take perhaps the most straightforward approach. We would like to define the tangent space at $p \in M$ by choosing a chart, thus identifying $M$ with $\mathbb{R}^{n}$ near $p$. We then take any vector $v \in \mathbb{R}^{n}$, and think of it as a tangent vector at $p$. The problem with this approach is that it would depend on a choice of charts, and the whole point of manifolds it to formulate calculus without coordinates. As a minimal fix then, let's repeat this process for every possible chart at $p$. Thus, we think of a tangent vector as being a collection of vectors in $\mathbb{R}^{n}$, one for each chart at $p$. However, these vectors should somehow reflect the way we glue charts together to construct the manifold, i.e., these vectors should satisfy some kind of compatibility condition as we change coordinates. From that point of view, it is perhaps natural to require the choice of vectors for each pair of charts to be related via the derivative of the transition function between the charts.

Definition 4.9 (Physically-defined tangent space). Let $M$ be an $n$-dimensional manifold, $p \in M$. Let $\mathcal{D}_{p}(M):=\{(U, h) \in \mathcal{D} \mid p \in U\}$ denote the set of charts around $p$. By a (physically-defined) tangent vector $v$ to $M$ at $p$, we mean a linear map

$$
v: \mathcal{D}_{p}(M) \rightarrow \mathbb{R}^{n}
$$

with the property that for any two charts at $p$, the associated vectors in $\mathbb{R}^{n}$ are mapped to each other by the derivative of the transition map; that is,

$$
v(V, k)=D_{h(p)}\left(k \circ h^{-1}\right)(v(U, h))
$$

for all $(U, h),(V, k) \in \mathcal{D}_{p}(M)$. We call the vector space of these maps $v$ the (physically-defined) tangent space to $M$ at $p$ and denote it by $T_{p}^{\mathrm{phy}} M$.
Remark 4.10. There is an explicit way to describe the property defining a physical tangent vector. Define $x^{i}=h_{i}$, the $i$-th component of $h$. Similarly, define $\widetilde{x}^{i}=k_{i}$, $v^{i}=v(U, h)_{i}$, and $\widetilde{v}^{i}(V, k)_{i}$. Referring to Figure 7, if we assume for convenience


Figure 7. Physically-defined tangent vector in coordinates.
that $h(p)=0$, in terms of the Jacobian matrix of $k \circ h^{-1}$, the equation $v(V, k)=$ $D_{h(p)}\left(k \circ h^{-1}\right)(v(U, h))$ becomes

$$
\left[J\left(k \circ h^{-1}\right)_{0}\right] v(U, h)=\left.\left[\frac{\partial \widetilde{x}^{i}}{\partial x^{j}}\right]\right|_{x=0}\left[\begin{array}{c}
v^{1} \\
\vdots \\
v^{n}
\end{array}\right]=\left[\begin{array}{c}
\widetilde{v}^{1} \\
\vdots \\
\widetilde{v}^{n}
\end{array}\right]=v(V, k)
$$

where $\left[\frac{\partial \widetilde{x}^{i}}{\partial x^{j}}\right]$ of partials of the components of $k \circ h^{-1}$.
To be explicit about the linear structure on $T_{p}^{\text {phy }}(M)$, if $v, w \in T_{p}^{\text {phy }}(M)$ and $\lambda \in$ $\mathbb{R}$, then, by definition,

$$
(\lambda v+w)(U, h):=\lambda v(U, h)+w(U, h)
$$

for each chart $(U, h)$ at $p$. Note that a physically-defined tangent vector $v: \mathcal{D}_{p}(M) \rightarrow$ $\mathbb{R}^{n}$ is determined by its value on any particular chart $(U, h)$. Its value on any other chart $(V, k)$ is then given by applying the derivative of the transition, as stated in Definition 4.9. In particular, fixing a chart $(U, h)$ determines an isomorphism of vector spaces.

$$
T_{p}^{\text {phy }}(M) \rightarrow \mathbb{R}^{n}
$$

$$
v \mapsto v(U, h)
$$

4.2. The three definitions are equivalent. The following proposition shows that our three definitions of tangent space are just three perspectives on the same thing.

Proposition 4.11. Let $M$ be an n-dimensional manifold and let $p \in M$. There are canonical (i.e., do not involve choosing bases) linear isomorphisms $\Phi_{1}, \Phi_{2}, \Phi_{3}$ that make the following diagram commute.


Proof. We start by describing the maps $\Phi_{1}, \Phi_{2}, \Phi_{3}$.

1. $T_{p}^{\text {geom }}(M) \longrightarrow T_{p}^{\mathrm{alg}}(M)$.

Define

$$
\begin{aligned}
\Phi_{1}: T_{p}^{\mathrm{geom}}(M) & \longrightarrow T_{p}^{\mathrm{alg}}(M) \\
{[\alpha] } & \longmapsto v_{[\alpha]}:[f] \mapsto(f \circ \alpha)^{\prime}(0)
\end{aligned}
$$

where $\alpha:(-\varepsilon, \varepsilon) \rightarrow M$ is a differentiable curve on $M$ with $\alpha(0)=p$ and $f \in \mathcal{E}_{p}(M)$ is a representative of a germ on $M$ at $p$.

First, we check that this map is well-defined. If $[f]=[g] \in \mathcal{E}_{p}(M)$ for some $g$, then since $f$ and $g$ agree on some smaller neighborhood $W$ of $p$, we have $f \circ \alpha=g \circ \alpha$ once we restrict to the appropriate domain. At the same time, if $\beta:(-\varepsilon, \varepsilon) \rightarrow M$ is a differentiable curve on $M$ such that $\alpha \sim \beta$ (we can always restrict the domains so that they are the same for both $\alpha$ and $\beta$ ), then the chain rule tells us that $v_{[\alpha]}=v_{[\beta]}$. In detail, suppose that the images of $\alpha$ and of $\beta$ are contained in some chart $(U, h)$ at $p$ (see Figure 8).


Figure 8. From $T_{p}^{\text {geom }}(M)$ to $T_{p}^{\text {alg }}(M)$.
Since $\alpha \sim \beta$, it follows that $(h \circ \alpha)^{\prime}(0)=(h \circ \beta)^{\prime}(0)$. So we have

$$
\begin{aligned}
(f \circ \alpha)^{\prime}(0) & =J\left(f \circ h^{-1} \circ h \circ \alpha\right)(0) \\
& =J\left(f \circ h^{-1}\right)(h(p)) J(h \circ \alpha)(0) \\
& =J\left(f \circ h^{-1}\right)^{\prime}(h(p)) J(h \circ \beta)(0)
\end{aligned}
$$

$$
\begin{aligned}
& =J\left(f \circ h^{-1} \circ h \circ \beta\right)(0) \\
& =(f \circ \beta)^{\prime}(0) .
\end{aligned}
$$

Next, we show that the image of $\Phi_{1}$ is in $T_{p}^{\text {alg }}(M)$. Namely, $v_{[\alpha]}$ is a derivation. Let $g \in \mathcal{E}_{p}(M)$. Using the product rule, we have

$$
\begin{aligned}
v_{[\alpha]}(f \cdot g) & =((f \cdot g) \circ \alpha)^{\prime}(0) \\
& =((f \circ \alpha) \cdot(g \circ \alpha))^{\prime}(0) \\
& =\left((f \circ \alpha)^{\prime} \cdot(g \circ \alpha)+(f \circ \alpha) \cdot(g \circ \alpha)^{\prime}\right)(0) \\
& =v_{[\alpha]}(f) \cdot g(p)+f(p) \cdot v_{[\alpha]}(g) .
\end{aligned}
$$

Finally, we show that $\Phi_{1}$ is linear. Take $[\alpha],[\beta] \in T_{p}^{\text {geom }} M$ and let $\lambda \in \mathbb{R}$. Suppose that $[\gamma]=\lambda[\alpha]+[\beta]$ for some $\gamma \in \mathcal{K}_{p}(M)$. So we have

$$
v_{\lambda[\alpha]+[\beta]}(f)=(f \circ \gamma)^{\prime}(0)
$$

Choosing a chart $(U, h)$ at $p$, recall that by definition of the linear structure on $T^{\text {geom }}$,

$$
(h \circ \gamma)^{\prime}(0)=\lambda(h \circ \alpha)^{\prime}(0)+(h \circ \beta)^{\prime}(0)
$$

A straightforward computation then gives us the linearity of $\Phi_{1}$ :

$$
\begin{aligned}
(f \circ \gamma)^{\prime}(0) & =J\left(f \circ h^{-1} \circ h \circ \gamma\right)(0) \\
& =J\left(f \circ h^{-1}\right)(h(p)) J(h \circ \gamma)(0) \\
& =J\left(f \circ h^{-1}\right)(h(p))(h \circ \gamma)^{\prime}(0) \\
& =J\left(f \circ h^{-1}\right)(h(p))\left(\lambda(h \circ \alpha)^{\prime}(0)+(h \circ \beta)^{\prime}(0)\right) \\
& =J\left(f \circ h^{-1}\right)(h(p))(\lambda J(h \circ \alpha)(0)+J(h \circ \beta)(0)) \\
& =\lambda J\left(f \circ h^{-1}\right)(h(p)) J(h \circ \alpha)(0)+J\left(f \circ h^{-1}\right)(h(p)) J(h \circ \beta)(0) \\
& =\lambda J\left(f \circ h^{-1} \circ h \circ \alpha\right)(0)+J\left(f \circ h^{-1} \circ h \circ \beta\right)(0) \\
& =\lambda(f \circ \alpha)^{\prime}(0)+(f \circ \beta)^{\prime}(0) \\
& =\lambda v_{[\alpha]}(f)+v_{[\beta]}(f) .
\end{aligned}
$$

2. $T_{p}^{\mathrm{alg}}(M) \longrightarrow T_{p}^{\mathrm{phy}}(M)$.

Define $\Phi_{2}: T_{p}^{\text {alg }}(M) \longrightarrow T_{p}^{\text {phy }}(M)$ as follows:

$$
\begin{aligned}
\Phi_{2}: T_{p}^{\mathrm{alg}}(M) & \longrightarrow T_{p}^{\mathrm{phy}}(M) \\
v & \longmapsto \bar{v}:(U, h) \mapsto\left(v\left(h_{1}\right), \ldots, v\left(h_{n}\right)\right)
\end{aligned}
$$

where $v: \mathcal{E}_{p}(M) \rightarrow \mathbb{R}$ is a linear derivation (defined on the germs of differentiable functions on $M$ at $p$ ), and $h_{i}$ is the $i$-th component of $h$.

Linearity of $\Phi_{2}$ is straightforward. We need to show that $\bar{v}$ behaves well under a change of coordinates. Let $(V, k)$ be another chart at $p$. We want $\bar{v}(V, k)=$ $D_{h(p)}\left(k \circ h^{-1}\right)(\bar{v}(U, h))$. That is,

$$
J\left(k \circ h^{-1}\right)(h(p))\left(\begin{array}{c}
v\left(h_{1}\right) \\
\vdots \\
v\left(h_{n}\right)
\end{array}\right)=\left(\begin{array}{c}
v\left(k_{1}\right) \\
\vdots \\
v\left(k_{n}\right)
\end{array}\right)
$$

Use $w=\left(w_{1}, \ldots, w_{n}\right)$ to denote $k \circ h^{-1}$, and, for convenience, assume that $h(p)=k(p)=0$. By Lemma 4.1, we can write

$$
w_{i}=\sum_{j=1}^{n} x_{j} w_{i, j}
$$

where $w_{i, j}(x)=\int_{0}^{1} \frac{\partial w_{i}}{\partial x_{j}}(t x) d t$. Taking partial derivative on both sides of the above displayed equation with respect to some $x_{\ell}$ and evaluating at zero:

$$
\begin{equation*}
\frac{\partial w_{i}}{\partial x_{\ell}}(0)=\left.\sum_{j=1}^{n} \frac{\partial x_{j}}{\partial x_{\ell}}\right|_{x=0} w_{i, j}(0)+\left.\sum_{j=1}^{n} x_{j}\right|_{x=0} \frac{\partial w_{i, j}}{\partial x_{\ell}}(0)=w_{i, \ell}(0) \tag{1}
\end{equation*}
$$

Note that

$$
k(x)=\left(k \circ h^{-1} \circ h\right)(x)=(w \circ h)(x)=\left(w_{1}(h(x)), \ldots, w_{n}(h(x))\right),
$$

and, therefore,

$$
k_{i}(x)=w_{i}(h(x))=\sum_{j=1}^{n} h_{j}(x) w_{i, j}(h(x)) .
$$

Since $v$ is a derivation (and recall that by assumption $h(p)=k(p)=0$ ),

$$
v\left(k_{i}\right)=\sum_{j=1}^{n}\left(v\left(h_{j}\right) w_{i, j}(h(p))+h_{j}(p) v\left(w_{i, j} \circ h\right)\right)=\sum_{j=1}^{n} v\left(h_{j}\right) w_{i, j}(0) .
$$

Therefore, by Equation 1, we have

$$
v\left(k_{i}\right)=\sum_{j=1}^{n} v\left(h_{j}\right) \frac{\partial w_{i}}{\partial x_{j}}(0)
$$

Write this out in matrix form:

$$
\left[\begin{array}{ccc} 
& \vdots & \\
\frac{\partial w_{i}}{\partial x_{1}}(0) & \cdots & \frac{\partial w_{i}}{\partial x_{n}}(0) \\
& \vdots &
\end{array}\right]\left[\begin{array}{c}
v\left(h_{1}\right) \\
\vdots \\
v\left(h_{n}\right)
\end{array}\right]=\left[\begin{array}{c}
v\left(k_{1}\right) \\
\vdots \\
v\left(k_{n}\right)
\end{array}\right]
$$

to see that this is exactly what we want.
3. $T_{p}^{\mathrm{phy}}(M) \longrightarrow T_{p}^{\text {geom }}(M)$.

Given $v \in T_{p}^{\text {phy }}(M)$ and a chart $(U, h)$ at $p$, define a curve $\alpha_{v}:(-\varepsilon, \varepsilon) \rightarrow M$ as follows: Pick a curve $\gamma:(-\varepsilon, \varepsilon) \rightarrow h(U) \subseteq \mathbb{R}^{n}$ such that $\gamma(0)=h(p)$ and $\gamma^{\prime}(0)=v(U, h)$. To be specific, say $\gamma(t)=h(p)+t v(U, h)$ for small enough $\varepsilon$. Then let $\alpha_{v}:=h^{-1} \circ \gamma:(-\varepsilon, \varepsilon) \rightarrow M$. Define $\Phi_{3}: T_{p}^{\text {phy }}(M) \longrightarrow T_{p}^{\text {geom }}(M)$ to be the map that assigns $v \in T_{p}^{\text {phy }}(M)$ to $\left[\alpha_{v}\right]$.

The main thing to check is that the equivalence class of $\alpha_{v}$ is independent of the choice of chart. Let $(V, k)$ be another chart at $p$ and define another curve

$$
\beta(t)=k^{-1}(k(p)+t v(V, k))
$$

Then $(k \circ \beta)^{\prime}(0)=v(V, k)$. To see that $\left[\alpha_{v}\right]=[\beta]$, we calculate

$$
\begin{aligned}
(k \circ \alpha)^{\prime}(0) & =J(k \circ \alpha)(0) \\
& =J\left(k \circ h^{-1} \circ h \circ \alpha\right)(0)
\end{aligned}
$$

$$
\begin{aligned}
& =J\left(k \circ h^{-1}\right)(h(p)) J(h \circ \alpha)(0) \\
& =J\left(k \circ h^{-1}\right)(h(p)) v(U, h) \\
& =v(V, k) \\
& =(k \circ \beta)^{\prime}(0)
\end{aligned}
$$

where the penultimate step follows since $v$ is a physically-defined tangent vector.
We summarize our work in Figure 9.


Figure 9. The linear isomorphisms $\Phi_{1}, \Phi_{2}$, and $\Phi_{3}$.

The last thing that needs to be shown is that $\Phi_{3} \circ \Phi_{2} \circ \Phi_{1}$ is the identity on $T_{p}^{\text {geom }}(M)$. Beside commutativity of the diagram in Proposition 4.11, this result will imply that all of the $\Phi_{i}$ are linear isomophisms since we have already established that all three versions of tangent space are $n$-dimensional.

Given a curve $[\alpha] \in T_{p}^{\text {geom }} M$, the resulting curve after applying $\Phi_{3} \circ \Phi_{2} \circ \Phi_{1}$ is the equivalence class of the curve $\beta$, where

$$
\beta(t)=h^{-1}\left(h(p)+t(h \circ \alpha)^{\prime}(0)\right)
$$

(Note that the $i$-th component of $\left(\Phi_{2} \circ \Phi_{1}\right)([\alpha])$ is exactly $\left.\left(h_{i} \circ \alpha\right)^{\prime}(0)\right)$. Then, since $h$ is a homeomorphism,

$$
(h \circ \beta)(t)=h(p)+t(h \circ \alpha)^{\prime}(0) .
$$

We can therefore conclude that $(h \circ \beta)^{\prime}(0)=(h \circ \alpha)^{\prime}(0)$, and $\alpha \sim \beta$ as desired.
4.3. Standard bases. We have now shown in precisely what sense the three spaces $T_{p}^{\text {geom }} M, T_{p}^{\text {alg }} M$, and $T_{p}^{\text {phy }} M$ are actually the same object. Thus, we are safe to talk about the tangent space to $M$ at $p$, denoted $T_{p} M$, and use any of the three definitions to denote a tangent vector at $p$.

Here we define the standard basis for $T_{p} M$ with respect to chart $(U, h)$ at $p$, denoted

$$
\left(\frac{\partial}{\partial x_{1}}\right)_{p}, \ldots,\left(\frac{\partial}{\partial x_{n}}\right)_{p} .
$$

As an element of $T_{p}^{\text {geom }}(M)$, we define $\left(\frac{\partial}{\partial x_{i}}\right)_{p}$ to be the equivalence class of curves represented by

$$
t \mapsto h^{-1}\left(h(p)+t e_{i}\right)
$$

where $e_{i}$ is the $i$-th standard basis vector of $\mathbb{R}^{n}$. As an element of $T_{p}^{\text {alg }} M$, i.e., as a derivation, for $f$ a germ at $p$, we define

$$
\left(\frac{\partial}{\partial x_{i}}\right)_{p} f:=\frac{\partial}{\partial x_{i}}\left(f \circ h^{-1}\right)(h(p)) .
$$

And finally, as an element of $T_{p}^{\text {phy }} M$, define $\left(\frac{\partial}{\partial x_{i}}\right)_{p}:=e_{i}$, the $i$-th standard basis vector of $\mathbb{R}^{n}$.

Our previous discuss of the three versions of tangent space have shown that these are bases, and one can check that they are compatible with our isomorphisms $\Phi_{i}$.
4.4. The differential of a mapping of manifolds. Let $f: M \rightarrow N$ be a differentiable mapping of manifolds and let $p \in M$. Finally, we are ready to define the differential of $f$ at $p$. It is a linear mapping $d f_{p}: T_{p} M \rightarrow T_{f}(p) N$ with the following three descriptions compatible with the maps $\Phi_{1}, \Phi_{2}, \Phi_{3}$ :

## - Geometric.

$$
\begin{aligned}
d^{\text {geom }} f_{p}: T_{p}^{\text {geom }} M & \longrightarrow T_{f(p)}^{\text {geom }} N \\
{[\alpha] } & \longmapsto[f \circ \alpha]
\end{aligned}
$$

## - Algebraic.

Precomposing by $f$ assigns germs at $p$ to germs at $f(p)$ (see Figure 10). So $f$ induces an algebra homomorphism, $f^{*}: \mathcal{E}_{f(p)}(N) \longrightarrow \mathcal{E}_{p}(M)$ that sends $\phi \in \mathcal{E}_{f(p)}(N)$ to $\phi \circ f \in \mathcal{E}_{p}(M)$. Therefore, precomposing by $f$ turns a derivation at $p$ to a derivation at $f(p)$. We define the differential to be

$$
\begin{aligned}
d^{\mathrm{alg}} f_{p}: T_{p}^{\mathrm{alg}} M & \longrightarrow T_{f(p)}^{\mathrm{alg}} N \\
v & \longmapsto v \circ f^{*}
\end{aligned}
$$



Figure 10. The germ of $\left.\phi \circ f\right|_{f^{-1}(U)}$ at $p$ is assigned to the germ of $\phi: U \rightarrow \mathbb{R}$ at $f(p)$.

Where $v \circ f^{*}$ is the derivation that sends a germ $\phi$ at $f(p)$ to the germ $v(\phi \circ f)$ at $p$ (again, see Figure 10).

## - Physical.

Choose a chart $(U, h)$ at $p$ and let $(V, k)$ be a chart at $f(p)$ such that $f(U) \subseteq V$. For $v \in T_{p}^{\text {phy }} M$, define

$$
\left(d^{\text {phy }} f_{p}(v)\right)(V, k):=D_{h(p)}\left(k \circ f \circ h^{-1}\right) v(U, h) .
$$

Thus, once local coordinates are taken, the differential is the ordinary derivative mapping given by the Jacobian matrix. See Figure 11.


Figure 11. The differential in terms of physically-defined tangent spaces.

Example 4.12. Consider

$$
\begin{aligned}
f: \mathbb{P}^{2} & \rightarrow \mathbb{P}^{3} \\
(x, y, z) & \mapsto\left(x^{3}, y^{3}, z^{3}, x y z\right)
\end{aligned}
$$

Let $p=(1, s, t) \in U_{x}$. Then $f(p)=\left(1, s^{3}, t^{3}, s t\right)$. Consider the standard open set $V_{a}=\left\{(a, b, c, d) \in \mathbb{P}^{3} \mid a \neq 0\right\}$ with coordinate mapping $\phi_{a}(a, b, c, d)=(b / a, c / a, d / a)$. With respect to these charts, we have

$$
\widetilde{f}(u, v)=\left(\phi_{a} \circ f \circ \phi_{x}^{-1}\right)(u, v)=\left(u^{3}, v^{3}, u v\right)
$$

Its Jacobian matrix at $\phi_{x}(p)=(s, t)$ is

$$
J \widetilde{f}(s, t)=\left[\begin{array}{cc}
3 s^{2} & 0 \\
0 & 3 t^{2} \\
t & s
\end{array}\right]
$$

So given $v \in T_{p}^{\text {phy }} M$ that assigns the vector $\left(v_{1}, v_{2}\right)$ to $\left(U_{x}, \phi_{x}\right)$, we have that $d f_{p}(v)$ assigns

$$
J \widetilde{f}(s, t)\left[\begin{array}{l}
v_{1} \\
v_{2}
\end{array}\right]
$$

to $\left(V_{a}, \phi_{a}\right)$.
Remark 4.13. The differential is functorial! The differential of the identity of $M$ is the identity of $T_{p} M$ :

$$
d \operatorname{id}_{p}=\operatorname{id}_{T_{p} M}
$$

The differential also respects composition, i.e., the chain rule holds. For a composition $M_{1} \xrightarrow{f} M_{2} \xrightarrow{g} M_{3}$ of differentiable maps, we have

$$
d(g \circ f)_{p}=d g_{f(p)} \circ d f_{p}
$$

## 5. Linear Algebra

Here we will summarize what we need to know about tensors. The presentation is most extracted from Frank Warner's, Foundations of Differential Manifolds and Lie Groups. In the following, all vector spaces are finite-dimensional and define over an arbitrary field $\mathbb{k}$ unless otherwise specified.
5.1. Products and coproducts. Let $V$ and $W$ be vector spaces. The vector space product of $V$ and $W$, is the Cartesian product $V \times W$ with linear structure

$$
\lambda(v, w)+\left(v^{\prime}, w^{\prime}\right):=\left(\lambda v+v^{\prime}, w+w^{\prime}\right)
$$

for all $\lambda \in \mathbb{k}$ and $(v, w),\left(v^{\prime}, w^{\prime}\right) \in V \times W$. It is the unique vector space (up to isomorphism) having the following universal property: Given a vector space $X$ and linear mappings $f: X \rightarrow V$ and $g: X \rightarrow W$, there exists a linear mapping $h: X \rightarrow$ $V \times W$ making the following diagram commute:


Here $\pi_{1}(v, w)=v$ and $\pi_{2}(v, w)=w$ are the first and second projection mappings, respectively. The mapping $h$ is given by $h(x)=(f(x), g(x))$.

Similarly, define the coproduct of $V$ and $W$, denoted $V \oplus W$, by $V \oplus W=V \times W$ with the same vector space structure. It has the universal property that given linear mappings $f: V \rightarrow X$ and $g: W \rightarrow X$ to a vector space $X$, there exists a unique linear mapping $h: V \oplus W \rightarrow X$ making the following diagram commute:


Here $\iota_{1}(v)=(v, 0)$ and $\iota_{2}(w)=(0, w)$ are the first and second inclusion mappings, respectively. The mapping $h$ is determined by $h(v, w)=f(v)+g(w)$. Notice how the commutative diagram for coproducts is obtained from that for products by flipping the direction of the arrows.

We could also define the product $V_{1} \times \cdots \times V_{n}$ and the coproduct $V_{1} \oplus \cdots \oplus V_{n}$ for vector spaces $V_{1}, \ldots, V_{n}$ by slightly extending the definition given above for the case $n=2$. We will leave that to the reader along with the statement and verification of the corresponding universal properties.

It may seem peculiar to make a distinction between the product and coproduct here given that they are exactly the same vector spaces. The difference comes when we consider an infinite family $\left\{V_{\alpha}\right\}_{\alpha \in A}$ of vector spaces. In that case, we can define their product $\prod_{\alpha \in A} V_{\alpha}$ using the Cartesian product, just as above. However,
their coproduct $\amalg_{\alpha \in A} V_{\alpha}$ is the vector subspace of $\prod_{\alpha \in A} V_{\alpha}$ consisting of vectors for which all but a finite number of components are zero. In that way, the universal properties are satisfied. (To see why the definition of the coproduct must change in this case, note that for the coproduct of two vector spaces, the mapping $h$ was given by $h(x)=f(x)+g(x)$. For an infinite family, the corresponding mapping would involve an infinite sum, and infinite sums of vectors are not defined in a general vector space.)
5.2. Tensor Products. Tensor products of vector spaces will allow us to think of multilinear objects (such as scalar products or determinants, and their generalizations) in terms of linear objects. We start with an informal description of the tensor product $U \otimes V \otimes W$ of three vector spaces $U, V$, and $W$. Its elements are linear combinations of expressions of the form $u \otimes v \otimes w$ where $u, v$, and $w$ are elements in $U, V$, and $W$, respectively. We are not allowed to swap the vectors, i.e., $v \otimes u \otimes w \neq u \otimes v \otimes w$, in general. The defining property of the tensor is that it is, roughly speaking, linear with respect to each entry. Thus, for example, if $\alpha \in \mathbb{k}, u^{\prime} \in U$ and $v^{\prime} \in V$,

$$
\left(\alpha u+u^{\prime}\right) \otimes v \otimes w=\alpha(u \otimes v \otimes w)+u^{\prime} \otimes v \otimes w
$$

and

$$
u \otimes\left(\alpha v+v^{\prime}\right) \otimes w=\alpha(u \otimes v \otimes w)+u \otimes v^{\prime} \otimes w
$$

and similarly for the last component. As a last example, let $w^{\prime} \in W$ a compute, using multilinearity:

$$
\begin{aligned}
\left(2 u+u^{\prime}\right) \otimes v \otimes\left(4 w-3 w^{\prime}\right)= & (2 u) \otimes v \otimes\left(4 w-3 w^{\prime}\right)+u^{\prime} \otimes v \otimes\left(4 w-3 w^{\prime}\right) \\
= & (2 u) \otimes v \otimes(4 w)+(2 u) \otimes v \otimes\left(-3 w^{\prime}\right) \\
& \quad+u^{\prime} \otimes v \otimes(4 w)+u^{\prime} \otimes v \otimes\left(-3 w^{\prime}\right) \\
= & 8 u \otimes v \otimes w-6 u \otimes v \otimes w^{\prime} \\
& \quad+4 u^{\prime} \otimes v \otimes w-3 u^{\prime} \otimes v \otimes w^{\prime} .
\end{aligned}
$$

Example 5.1. Let $e_{1}, e_{2}$ be the standard basis vectors for $\mathbb{R}^{2}$, and let $f_{1}, f_{2}, f_{3}$ be the standard basis vectors for $\mathbb{R}^{3}$. Takev $=(2,3) \in \mathbb{R}^{2}$ and $w=(3,2,1) \in \mathbb{R}^{2}$. Then we can write $v \otimes w \in \mathbb{R}^{2} \otimes \mathbb{R}^{3}$ in terms of the $e_{i} \otimes f_{j}$ :

$$
\begin{aligned}
v \otimes w & =(2,3) \otimes(3,2,1) \\
& =\left(2 e_{1}+3 e_{2}\right) \otimes\left(3 f_{1}+2 f_{2}+f_{3}\right) \\
& =6 e_{1} \otimes f_{1}+4 e_{1} \otimes f_{2}+2 e_{1} \otimes f_{3}+9 e_{2} \otimes f_{1}+6 e_{2} \otimes f_{2}+3 e_{2} \otimes f_{3}
\end{aligned}
$$

If you follow the above calculations, then you understand exactly the type of gadget we are looking for. We pause now for the formal construction (which is not as important as understanding the above calculation). We then describe the purpose of the tensor product be exhibiting its universal property.
5.2.1. Construction of the tensor product. To construct the tensor product $V \otimes W$ of vector spaces $V$ and $W$, let $F(V, W)$ be the free vector space on the set of symbols $\{[v, w] \mid v \in V, w \in W\}$. Thus, these symbols form a basis for $F(V, W)$ : an arbitrary element of $F(V, W)$ is a linear combination of these symbols and there is no relation among them. For instance, $[v, w],\left[v^{\prime}, w\right]$ and $\left[v+v^{\prime}, w\right]$ are linearly independent if $v, v^{\prime}$, and $v+v^{\prime}$ are distinct.

We now mod out by a subspace of $F(V, W)$ in order to force the resulting equivalence classes of the $[v, w]$ to be "multilinear". To that end, define $T$ to be the subspace of $F(V, W)$ generated by the following vectors:

$$
\begin{aligned}
& {\left[v_{1}+v_{2}, w\right]-\left[v_{1}, w\right]-\left[v_{2}, w\right]} \\
& {\left[v, w_{1}+w_{2}\right]-\left[v, w_{1}\right]-\left[v, w_{2}\right]} \\
& {[\alpha v, w]-\alpha[v, w]} \\
& {[v, \alpha w]-\alpha[v, w]}
\end{aligned}
$$

for all $\alpha \in \mathbb{k}, v, v_{1}, v_{2} \in V$, and $w, w_{1}, w_{2} \in W$. Finally, define

$$
V \otimes W:=F(V, W) / T
$$

and $v \otimes w:=[v, w] \bmod T$ for each $v \in V$ and $w \in W$. (Modding out by $T$ forces (the equivalence class of) each of the generators listed above to be 0 , which gives just the multilinearity we want.) The tensor product of vector spaces $V_{1}, \ldots, V_{n}$ is defined similarly.

Remark 5.2. Note that scalars can "float around" in tensors: for $\alpha \in \mathbb{k}$ and $u \otimes v \otimes$ $w \in U \otimes V \otimes W$,

$$
\alpha(u \otimes v \otimes w)=(\alpha u) \otimes v \otimes w=u \otimes(\alpha v) \otimes w=u \otimes v \otimes(\alpha w)
$$

5.2.2. Universal property of the tensor product. Define

$$
\begin{aligned}
\iota: V \times W & \rightarrow V \otimes W \\
(v, w) & \mapsto v \otimes w
\end{aligned}
$$

and note that $\iota$ is bilinear $\iota\left(\alpha v+v^{\prime}, w\right)=\alpha \iota(v, w)+\iota\left(v^{\prime}, w\right)$, and similarly for the second component). The tensor product $V \otimes W$ is characterized (up to isomorphism of vector spaces) by the following universal property: Given any bilinear mapping $f: V \times W \rightarrow U$ to a vector space $U$, there exists a unique linear mapping $h: V \otimes W \rightarrow U$ such that the following diagram commutes:


Thus, the tensor product allows us to represent a bilinear mapping with a linear mapping-each contains the same information. The proof of the universal property is left as an exercise.

More generally, there is a similar commutative diagram that relates a multilinear mapping $V_{1} \times \cdots \times V_{n} \rightarrow U$ with a linear mapping $V_{1} \otimes \cdots \otimes V_{n} \rightarrow U$ :

5.2.3. Identities. A typical use of the universal property of tensor is to define a linear mapping $V \otimes W \rightarrow U$ for some vector space. One could imagine defining a function by describing the image of an element $v \otimes w$ in terms of some rule involving $v$ and $w$. The problem is that each $v \otimes w$ is really an equivalence class in $F(V, W)$. So there is the question of whether the mapping is well-defined. To get around that problem, one instead defines a bilinear mapping $f: V \times W \rightarrow U$. By the universal property of tensor products, there is then an induced linear mapping $\tilde{f}: V \otimes W \rightarrow U$ with the property that

$$
\tilde{f}(v \otimes w)=f(v, w) .
$$

Also note that any linear mapping $f: V \otimes W \rightarrow U$ is determined by its values on tensors of the form $v \otimes w$. Not every element of $V \otimes W$ has the form $v \otimes w$ for some choices of $v$ and $w$. However, elements of that form span $V \otimes W$.

The proof of the following proposition illustrates the principle of using the universal property to define mappings involving tensors.

Proposition 5.3. Let $U, V$ and $W$ be vector spaces over $\mathbb{k}$.
(1) $V \otimes \mathbb{k} \approx V$.
(2) $V \otimes W \approx W \otimes V$.
(3) $(V \otimes W) \otimes U \approx V \otimes(W \otimes U) \approx V \oplus W \oplus U$.
(4) $V \otimes(W \oplus U) \approx(V \otimes W) \oplus(V \otimes U)$.

Proof. We will prove the first two parts, the others being similar.
(1) Define $f: V \rightarrow V \otimes \mathbb{k}$ by $f(v)=v \otimes 1$. It is straightforward to check that $f$ is linear. To define the inverse, note that

$$
\begin{aligned}
V \times \mathbb{k} & \rightarrow V \\
(v, \alpha) & \mapsto \alpha v
\end{aligned}
$$

is bilinear. It, thus, induces a linear mapping $g: V \otimes \mathbb{k} \rightarrow V$ determined by $g(v \otimes \alpha)=\alpha v$. We then have

$$
(f \circ g)(v \otimes \alpha)=f(\alpha v)=(\alpha v) \otimes 1=v \otimes \alpha .
$$

and

$$
(g \circ f)(v)=g(f(v))=g(v \otimes 1)=1 \cdot v=v .
$$

(2) The bilinear mapping

$$
\begin{aligned}
V \times W & \rightarrow W \otimes V \\
(v, w) & \mapsto w \otimes v
\end{aligned}
$$

induces a linear mapping $f: V \otimes W \rightarrow W \otimes V$ with the property that $f(v \otimes$ $w)=w \otimes v$. By a similar argument, there is a linear mapping $g: W \otimes V \rightarrow$ $V \otimes W$ with the property that $g(w \otimes v)=v \otimes w$. It is then easy to check that $f$ and $g$ are inverses.

Proposition 5.4. Let $V$ and $W$ be vector spaces with bases $v_{1}, \ldots, v_{m}$ and $w_{1}, \ldots, w_{n}$, respectively. Then $V \otimes W$ has basis $\left\{v_{i} \otimes w_{j}\right\}_{i . j}$. In particular,

$$
\operatorname{dim}(V \otimes W)=\operatorname{dim}(V) \operatorname{dim}(W)
$$

Proof. Using the ideas presented above, we can define a sequence of isomorphisms

$$
V \otimes W \approx V \otimes\left(\oplus_{j=1}^{n} \mathbb{k}\right) \approx \oplus_{j=1}^{n}(V \otimes \mathbb{k}) \approx \oplus_{j=1}^{n} V \approx \oplus_{j=1}^{n} \oplus_{i=1}^{m} \mathbb{k} \approx \mathbb{k}^{m n}
$$

sending the $v_{i} \otimes w_{j}$ to the standard basis vectors for $\mathbb{k}^{m n}$.
5.3. Symmetric and exterior products. If $u \otimes v \in V \otimes V$ and $u \neq v$, then it is typically the case that $u \otimes v \neq v \otimes u$. (For an exception, we have $u \otimes(\alpha u)=(\alpha u) \otimes u$ for any scalar $\alpha$.) However, there are many situations in which such a commutativity property would be desirable. Thus, we are led to the notion of symmetric tensors. For $\ell \in \mathbb{N}$, define

$$
V^{\otimes \ell}:=\underbrace{V \otimes \cdots \otimes V}_{\ell} .
$$

We take $V^{\otimes 0}:=\mathbb{k}$. Consider the subspace $T$ of $V^{\otimes \ell}$ generated by elements of the form

$$
v_{1} \otimes \cdots \otimes v_{i} \otimes \cdots \otimes v_{j} \otimes \cdots \otimes v_{\ell}-v_{1} \otimes \cdots \otimes v_{j} \otimes \cdots \otimes v_{i} \otimes \cdots \otimes v_{\ell}
$$

formed by swapping two components of the tensor. Then the $\ell$-th symmetric product of $V$ is quotient vector space

$$
\operatorname{Sym}^{\ell} V=V^{\otimes \ell} / T
$$

The element $v_{1} \otimes \cdots \otimes v_{\ell} \in \operatorname{Sym}^{\ell} V$ is denoted $v_{1} \cdots v_{n}$, and we are allowed to commute the $v_{i}$ without changing the element. Elements of $\mathrm{Sym}^{\ell} V$ are called symmetric tensors. The elements of $\mathrm{Sym}^{\ell} V$ behave just like usual tensors, except we are now allowed to swap components.

A multilinear function $f: V^{\times \ell} \rightarrow W$ is symmetric if $f\left(v_{1}, \ldots, v_{i}, \ldots, v_{j}, \ldots, v_{\ell}\right)=$ $f\left(v_{1}, \ldots, v_{j}, \ldots, v_{i}, \ldots, v_{\ell}\right)$ for all $i, j$. The universal property of symmetric tensors is that given a multilinear symmetric mapping $V^{\times \ell} \rightarrow W$ from the $\ell$-fold Cartesian product of $V$ with itself to $W$, there exists a unique linear mapping $\operatorname{Sym}^{\ell} V \rightarrow W$ making the following diagram commute:


The mapping $V^{\times \ell} \rightarrow \operatorname{Sym}^{\ell} V$ is the natural one determined by $\left(v_{1}, \ldots, v_{\ell}\right) \mapsto$ $v_{1} \cdots v_{n}$.

Example 5.5. The ordinary dot product:

$$
\mathbb{R}^{n} \times \mathbb{R}^{n} \rightarrow \mathbb{R}
$$

$$
(v, w) \mapsto\langle v, w\rangle=\sum_{i=1}^{n} v_{i} w_{i}
$$

is bilinear. Hence, there is a unique induced mapping $\operatorname{Sym}^{2} \mathbb{R}^{n} \rightarrow \mathbb{R}$ making the following diagram commute.


It is determined by $(v, w) \mapsto\langle v, w\rangle$ for all $v, w \in \mathbb{R}^{n}$.
Proposition 5.6. If $V$ has dimension $n$, then

$$
\operatorname{dim} \operatorname{Sym}^{\ell} V=\binom{n+\ell-1}{\ell}
$$

Proof. Exercise. If $V$ has basis $x_{1}, \ldots, x_{n}$, show that $\operatorname{Sym}^{\ell} V$ has basis

$$
\left\{x_{1}^{a_{1}} \cdots x_{n}^{a_{n}} \mid a_{i} \geq 0 \text { and } \sum_{i} a_{i}=\ell\right\}
$$

i.e., the basis consists of all monomials of degree $\ell$ in $x_{1}, \ldots, x_{n}$. The result then follows from the usual stars-and-bars argument.

We also define

$$
\operatorname{Sym} V:=\oplus_{\ell \geq 0} \operatorname{Sym}^{\ell} V
$$

where $\operatorname{Sym}^{0} V:=\mathbb{k}$. One may use the universal property for symmetric products to show there is a well-defined multiplication mapping

$$
\begin{aligned}
\operatorname{Sym}^{a} V \times \operatorname{Sym}^{b} V & \rightarrow \operatorname{Sym}^{a+b} V \\
(\alpha, \beta) & \mapsto \alpha \beta .
\end{aligned}
$$

This multiplication turns $\operatorname{Sym} V$ into a $\mathbb{k}$-algebra (i.e., a vector space over $\mathbb{k}$ which is also a ring (i.e., has a nice multiplication operation)). In fact, if $V$ has basis $x_{1}, \ldots, x_{n}$, then $\operatorname{Sym} V$ is isomorphic to the polynomial ring in $x_{1}, \ldots, x_{n}$ with coefficients in $\mathbb{k}$ (but see the section on duality, below, to get a better formulation).

In the same way we defined symmetric products by forcing tensors to commute, we define exterior products by forcing tensors to anti-commute.

The $\ell$-th exterior product of $V$, denoted $\Lambda^{\ell} V$, to be the quotient vector space

$$
\Lambda^{\ell} V:=V^{\otimes \ell} / T
$$

where $T=\operatorname{Span}_{\mathbb{k}}\left\{v_{1} \otimes \cdots \otimes v_{\ell} \mid v_{i}=v_{j}\right.$ for some $\left.i \neq j\right\}$. The equivalence class of $v_{1} \otimes \cdots \otimes v_{\ell}$ is denoted $v_{1} \wedge \cdots \wedge v_{\ell}$. Elements of $\Lambda^{\ell} V$ are called alternating tensors.

Proposition 5.7. In $\Lambda^{\ell} V$, we have, swapping $v_{i}$ and $v_{j}$ for any $i \neq j$,

$$
v_{1} \wedge \cdots \wedge v_{i} \wedge \cdots \wedge v_{j} \wedge \cdots \wedge v_{\ell}=-v_{1} \wedge \cdots \wedge v_{j} \wedge \cdots \wedge v_{i} \wedge \cdots \wedge v_{\ell}
$$

Proof. Through direct computation:

$$
\begin{aligned}
0 & =v_{1} \wedge \cdots \wedge\left(v_{i}+v_{j}\right) \wedge \cdots \wedge\left(v_{i}+v_{j}\right) \wedge \cdots \wedge v_{\ell} \\
& =0+v_{1} \wedge \cdots \wedge v_{i} \wedge \cdots \wedge v_{j} \wedge \cdots \wedge v_{\ell}+v_{1} \wedge \cdots \wedge v_{j} \wedge \cdots \wedge v_{i} \wedge \cdots \wedge v_{\ell}+0
\end{aligned}
$$

Remark 5.8. If char $\mathbb{k} \neq 2$, i.e., if $1+1 \neq 0$ in $\mathbb{k}$, then we can use the equation displayed in the above proposition to define $T$.

A multilinear map $f: V^{\times \ell} \rightarrow W$ is alternating if $f\left(v_{1}, \ldots, v_{\ell}\right)=0$ if $v_{i}=v_{j}$ for some $i \neq j$. There is a canonical multilinear alternating mapping

$$
\begin{aligned}
\iota: V^{\times \ell} & \rightarrow \Lambda^{\ell} V \\
\left(v_{1}, \ldots, v_{\ell}\right) & \mapsto v_{1} \wedge \cdots \wedge v_{\ell}
\end{aligned}
$$

The $\ell$-th exterior product is characterized by the following universal property: given any vector space $W$ and multilinear alternating mapping $V^{\times \ell} \rightarrow W$, there is a unique linear map $\Lambda^{\ell} V \rightarrow W$ making the following diagram commute:


Example 5.9. Let $\mathbb{k}=\mathbb{R}$. The cross product $\times: \mathbb{R}^{n} \times \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is bilinear and alternating. So there is a unique map $\times: \Lambda^{2} \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ that sends $u \wedge v$ to $u \times v$ for any $u, v \in \mathbb{R}^{n}$.

Proposition 5.10. If $V$ has dimension $n$, then

$$
\operatorname{dim} \Lambda^{\ell} V=\binom{n}{\ell}
$$

Proof. Exercise. Show that if $e_{1}, \ldots, e_{n}$ is a basis for $V$, then $\left\{e_{i_{1}} \wedge \cdots \wedge e_{i_{\ell}} \mid i_{1}<\cdots<i_{\ell}\right\}$ is a basis for $\Lambda^{\ell} V$.

Exercise 5.11. Recall that the determinant det of a square matrix over $\mathbb{k}$ is the unique multilinear alternating function of its rows that sends the identity matrix to 1 . Let $e_{1}, \ldots, e_{n}$ be the standard basis of $\mathbb{K}^{n}$ and let $v_{1}, \ldots, v_{n} \in \mathbb{K}^{n}$. Show that

$$
v_{1} \wedge \cdots \wedge v_{n}=\operatorname{det}\left(v_{1}, \ldots, v_{n}\right) e_{1} \wedge \cdots \wedge e_{n}
$$

Exercise 5.12. Show that $v_{1}, \ldots, v_{\ell} \in V$ are linearly dependent if and only if $v_{1} \wedge \cdots \wedge v_{\ell}=0$.

Remark 5.13. If $v_{1}, \ldots, v_{\ell}$ are linearly independent, it is tempting to identify the one-dimensional space spanned by $v_{1} \wedge \cdots \wedge v_{\ell} \in \Lambda^{\ell} V$ with the linear space spanned by $v_{1}, \ldots, v_{\ell}$ in $V$. We will talk about this more when we talk about Grassmannians later in these notes.

Define

$$
\Lambda^{\bullet} V:=\oplus_{\ell \geq 0} \Lambda^{\ell} V
$$

where $\Lambda^{0} V:=\mathbb{k}$. Using the universal property of exterior products, we can define a multiplication on alternating tensors:

$$
\begin{aligned}
\Lambda^{r} V \times \Lambda^{s} V & \rightarrow \Lambda^{r+s} V \\
(\lambda, \mu) & \mapsto \lambda \wedge \mu
\end{aligned}
$$

The vector space $\Lambda^{\bullet} V$ with this multiplication is called the Grassmann algebra on $V$. Note that for $\lambda \in \Lambda^{r} V$ and $\mu \in \Lambda^{s} V$ we have

$$
\lambda \wedge \mu=(-1)^{r s} \mu \wedge \lambda
$$

5.4. Dual spaces. Let hom $(V, W)$ denote the linear space of all linear mappings $V \rightarrow$ $W$. If $f, g \in \operatorname{hom}(V, W)$ and $\lambda \in \mathbb{k}$, then $\lambda f+g$ is defined by

$$
(\lambda f+g)(v):=\lambda f(v)+g(v)
$$

The dual of the vector space $V$ is

$$
V^{*}:=\operatorname{hom}(V, \mathbb{k})
$$

Exercise 5.14. If $v_{1}, \ldots, v_{n}$ is a basis for $V$, define $v_{i}^{*} \in V^{*}$ for each $i$ by

$$
v_{i}^{*}\left(v_{j}\right):=\delta(i, j)= \begin{cases}1 & \text { if } i=j \\ 0 & \text { otherwise }\end{cases}
$$

Show that $\left\{v_{1}^{*}, \ldots, v_{n}^{*}\right\}$ is a basis for $V^{*}$. It is called the dual basis to $\left\{v_{1}, \ldots, v_{n}\right\}$. Note that this exercise shows that $V$ and $V^{*}$ are isomorphic (if $V$ is finite-dimensional). However, the isomorphism depends on a choice of basis.

A linear mapping $f: V \rightarrow W$ induces a dual linear mapping

$$
\begin{aligned}
f^{*}: W^{*} & \rightarrow V^{*} \\
\phi & \mapsto \phi \circ f
\end{aligned}
$$

as pictured below;


The read should check that dualization if functorial: $\mathrm{id}_{V}: V \rightarrow V{\text { induces } \mathrm{id}_{V^{*}}: V^{*} \rightarrow}$ $V^{*}$, and commutative diagrams are preserved:

...........>


Proposition 5.15. We have two isomorphisms:
(1)

$$
\begin{aligned}
\Lambda^{\ell} V^{*} & \rightarrow\left(\Lambda^{\ell} V\right)^{*} \\
\phi_{1} \wedge \cdots \wedge \phi_{\ell} & \mapsto\left[v_{1} \wedge \cdots \wedge v_{\ell} \mapsto \operatorname{det}\left(\phi_{i}\left(v_{j}\right)\right)\right]
\end{aligned}
$$

(2) and

$$
\begin{aligned}
& \operatorname{Sym}^{\ell} V^{*} \rightarrow\left(\operatorname{Sym}^{\ell} V\right)^{*} \\
&\left.\phi_{1} \cdots \phi_{\ell} \mapsto\left[v_{1} \cdots v_{\ell} \mapsto \sum_{\sigma \in \mathfrak{S}_{\ell}} \prod_{i=1}^{\ell} \phi_{\sigma(i)}\left(v_{i}\right)\right)\right]
\end{aligned}
$$

where $\mathfrak{S}_{\ell}$ is the symmetric group of permutations of $[\ell]:=\{1, \ldots, \ell\}$.
Sketch of proof of (1). The mapping in (1) exists from the universal property of alternating tensors and the fact that the determinant is alternating and multilinear as a function of its rows.

We define the inverse mapping $\left(\Lambda^{\ell} V\right)^{*} \rightarrow \Lambda^{\ell} V^{*}$ in terms of a bases. Fix a basis $e_{1}, \ldots, e_{n}$ for $V$, and for each integer vector $\mu$ such that

$$
1 \leq \mu_{1}<\cdots<\mu_{\ell} \leq n
$$

define

$$
e_{\mu}:=e_{\mu_{1}} \wedge \cdots \wedge e_{\mu_{\ell}}
$$

These $\binom{n}{\ell}$ vectors form a basis for $\Lambda^{\ell} V$. Letting $e_{1}^{*}, \ldots, e_{\ell}^{*}$ be the corresponding dual basis for $V^{*}$, we get the corresponding basis of vectors

$$
e_{\mu}^{*}:=e_{\mu_{1}}^{*} \wedge \cdots \wedge e_{\mu_{\ell}}^{*}
$$

for $\Lambda^{\ell}\left(V^{*}\right)$.
For each $\omega \in\left(\Lambda^{\ell} V\right)^{*}$ and basis vector $e_{\mu}$, define

$$
\omega_{\mu}:=\omega\left(e_{\mu_{1}}\right) \wedge \cdots \wedge \omega\left(e_{\mu_{\ell}}\right)
$$

Finally, define

$$
\begin{aligned}
g:\left(\Lambda^{\ell} V\right)^{*} & \longrightarrow \Lambda^{\ell} V^{*} \\
w & \sum_{\mu: 1 \leq \mu_{1}<\cdots<\mu_{l} \leq n} w_{\mu} e_{\mu}^{*}
\end{aligned}
$$

where $e_{\mu}^{*}=e_{\mu_{1}}^{*} \wedge \cdots \wedge e_{\mu_{l}}^{*}$. The reader may check that $g$ is well-defined and inverse to the mapping in (1).
5.4.1. Forms. A multilinear $\ell$-form on $V$ is a multilinear mapping $V^{\times \ell} \rightarrow \mathbb{k}$. By the universal property of tensor products, each corresponds to a mapping $V^{\otimes \ell} \rightarrow \mathbb{k}$, and hence to an element of $\left(V^{\otimes \ell}\right)^{*}$. Thus, there is a linear isomorphism between the vector space of multilinear $\ell$-forms on $V$ and the vector space $\left(V^{\otimes \ell}\right)^{*} \approx\left(V^{*}\right)^{\otimes \ell}$. A symmetric $\ell$-form on $V$ is a multilinear symmetric mapping $V^{\times \ell} \rightarrow \mathbb{k}$. By the universal property of symmetric tensors this mapping is identified with a mapping $\operatorname{Sym}^{\ell} V \rightarrow \mathbb{k}$, i.e., to an element of $\left(\operatorname{Sym}^{\ell} V\right)^{*}$. Thus, we get an isomorphism between the linear space of symmetric $\ell$-forms and $\left(\mathrm{Sym}^{\ell}(V)\right)^{*} \approx \mathrm{Sym}^{\ell} V^{*}$. An alternating $\ell$-form on $V$ is a multilinear alternating mapping $V^{\times \ell} \rightarrow \mathbb{k}$. Since it
will be useful later, we will let $\mathrm{Alt}^{\ell} V$ denote the linear space of alternating $\ell$-forms on $V$. Arguing as above, the universal property of alternating tensors gives an isomorphism

$$
\operatorname{Alt}^{\ell} V \approx\left(\Lambda^{\ell} V\right)^{*} \approx \Lambda^{\ell} V^{*}
$$

Example 5.16. The space $\mathrm{Sym}^{2} V^{*}$ is the linear space of symmetric bilinear forms on $V$. For instance, any inner product $\langle$,$\rangle on V$ can be thought of as an element of $\mathrm{Sym}^{2} V^{*}$. (Recall that an inner product is a nondegenate bilinear form. Nondegenerate means that if $v \in V$ and $\langle v, w\rangle=0$ for all $w \in V$, then $w=0$.)

Recall that the wedge product defined for the exterior algebra $\Lambda^{\bullet} V^{*}$ is given by

$$
\begin{aligned}
\wedge: \Lambda^{r} V^{*} \times \Lambda^{s} V^{*} & \longrightarrow \Lambda^{r+s} V^{*} \\
(\omega, \eta) & \mapsto \omega \wedge \eta
\end{aligned}
$$

The product is bilinear, associative, and anti-commutative.
Exercise 5.17. Show that taking the exterior algebra of the dual space defines a contravariant functor $\Lambda^{\bullet}: V \mapsto \Lambda^{\bullet} V^{*}$ from the category of finite-dimensional $\mathbb{k}$ vector spaces to the category of graded anti-commutative $\mathbb{k}$-algebras. That is, show that (1) for $f: V \rightarrow W$ and $\omega, \eta \in \Lambda^{\bullet} W^{*}, f^{*}(\omega \wedge \eta)=\left(f^{*} \omega\right) \wedge\left(f^{*} \eta\right)$, (2) for the composition $U \xrightarrow{f} V \xrightarrow{g} W$, we have $(g \circ f)^{*}=f^{*} \circ g^{*}: \Lambda^{\bullet} W^{*} \rightarrow \Lambda^{\bullet} V^{*} \rightarrow \Lambda^{\bullet} U^{*}$, and (3) $\mathrm{id}_{V}^{*}$ is the identity map on $\Lambda^{\bullet} V^{*}$.
5.4.2. Pullbacks. A linear map $L: V \rightarrow W$ induces, for each $\ell \geq 0$, a pullback mapping

$$
\begin{aligned}
L^{*}:\left(\Lambda^{\ell} W\right)^{*} & \rightarrow\left(\Lambda^{\ell} V\right)^{*} \\
\omega & \mapsto\left[v_{1} \wedge \cdots \wedge v_{\ell} \mapsto \omega\left(L v_{1} \wedge \cdots \wedge L v_{l}\right)\right]
\end{aligned}
$$

or equivalently (using the isomorphisms established above,

$$
\begin{array}{ccccc}
L^{*}: \Lambda^{\ell} W^{*} & \rightarrow & \Lambda^{\ell} V^{*} & \rightarrow & \left(\Lambda^{\ell} V\right)^{*} \\
\phi_{1} \wedge \cdots \wedge \phi_{\ell} & \mapsto & L^{*} \phi_{1} \wedge \cdots \wedge L^{*} \phi_{\ell} & \mapsto & {\left[v_{1} \wedge \cdots \wedge v_{\ell} \mapsto \operatorname{det}\left(\left(\phi_{i} \circ L\right)\left(v_{j}\right)\right)\right]}
\end{array}
$$

## 6. Vector Bundles

Recall that in section 4, we learned three equivalent ways to define the tangent space $T_{p} M$ of a manifold $M$ at a point $p \in M$. We are also interested in its dual space $T_{p}^{*} M$, called the cotangent space, as well as other related vector spaces $\Lambda^{k} T_{p} M, \Lambda^{k} T_{p}^{*} M, \operatorname{Sym}^{k} T_{p} M, T_{p}^{*} M^{\otimes k}$, etc.

For each $p$, we can choose a chart $U$ at $p$ and look at all the tangent spaces at points in $U$. Properties of these tangent spaces give local properties of $M$. But to study the global properties, we need to look at all points in $M$ and form a vector bundle. That is, to each point $p \in M$, we attach a vector space such that these vector spaces behave well under change of charts. Before studying specific bundles, we briefly introduce the general theory of vector bundles.

Definition 6.1. Let $M$ be an $n$-dimensional manifold, $E$ an $(n+r)$-dimensional manifold, and $\pi: E \rightarrow M$ a smooth surjection. We say $\pi: E \rightarrow M$ is a vector bundle of rank $r$ over $M$ if
(1) For every $p \in M$, the fiber over $p, E_{p}:=\pi^{-1}(p)$ is a real vector space of dimension $r$;
(2) Every point $p \in M$ has an open neighborhood $U$ such that there is a fiberpreserving diffeomorphism $\phi_{U}: \pi^{-1}(U) \rightarrow U \times \mathbb{R}^{r}$, meaning it makes the following diagram commute:

where $\pi_{1}$ is the usual projection map onto the first coordinate. Furthermore, $\phi_{U}$ restricts to a linear isomorphism $E_{p} \rightarrow\{p\} \times \mathbb{R}^{r}$ on each fiber.

We call $E$ the total space, $M$ the base space, and the map $\phi_{U}$ a trivialization of $\pi^{-1}(U)$. A line bundle is a vector bundle of rank 1.

Example 6.2 (Product Bundle). Let $V$ be a vector space of dimension $r$. Then the projection $\pi: M \times V \rightarrow M$ is a vector bundle of rank $r$. For example, the cylinder $S^{1} \times \mathbb{R}$ together with the projection $\pi: S^{1} \times \mathbb{R} \rightarrow S^{1}$ is a product bundle over $S^{1}$.

Definition 6.3. Let $\pi_{E}: E \rightarrow M$ and $\pi_{F}: F \rightarrow N$ be vector bundles. A bundle map from $E$ to $F$ is a pair of differentiable maps $(\phi: E \rightarrow F, f: M \rightarrow N)$ such that
(1) The following diagram commutes:

(2) $\phi$ restricts to a linear map $\phi_{p}: E_{p} \rightarrow F_{f(p)}$ for each $p \in M$.

Abusing language, we usually call $\phi: E \rightarrow F$ alone the bundle map.

If $M=N$, then we call the pair $\left(\phi: E \rightarrow F, \mathrm{id}_{M}\right)$ a bundle map over $M$. Also, $\phi$ is a bundle isomorphism over $M$ if there is another bundle map $\psi: F \rightarrow E$ over $M$ such that $\phi \circ \psi=\mathrm{id}_{F}$ and $\psi \circ \phi=\mathrm{id}_{E}$.

Definition 6.4. A vector bundle $\pi: E \rightarrow M$ of rank $r$ is trivial if it is isomorphic to the product bundle $M \times \mathbb{R}^{r} \rightarrow M$.

Example 6.5 (Möbius Strip). The open Möbius strip is the quotient of $[0,1] \times \mathbb{R}$ by the identification $(0, t) \sim(1,-t)$. It gives a line bundle over $S^{1}$ that is not isomorphic to the cylinder. See Exercise 6.9.


Figure 12. A Möbius strip.

Definition 6.6. Let $\pi: E \rightarrow M$ be a vector bundle and $U \subseteq M$ be open. A section of $E$ over $U$ is a smooth function $s: U \rightarrow E$ such that $\pi \circ s=\mathrm{id}_{U}$. That is, for each $p \in U$, the section $s$ picks out one element of the fiber $E_{p}$. We use $\Gamma(U, E)$ to denote the collection of all sections of $E$ over $U$. If $U=M$, we also write $\Gamma(E)$ instead of $\Gamma(M, E)$ and call elements of $\Gamma(E)$ the global sections.

Remark 6.7. First note that $\Gamma(U, E)$ is a vector space over $\mathbb{R}$. Also note that there is an action on $\Gamma(U, E)$ by smooth functions on $U$ : for a section $s \in \Gamma(U, E)$, a smooth function $f: U \rightarrow \mathbb{R}$, and a point $p \in U$, we define $(f s)(p):=f(p) s(p) \in E_{p}$. Thus, $\Gamma(U, E)$ is a $\mathcal{C}^{\infty}(U)$-module.

Example 6.8 (Sections of a product line bundle). A section $s$ of the product line bundle $M \times \mathbb{R}^{k} \rightarrow M$ is a map $s(p)=(p, f(p))$ for each $p \in M$ (see Figure 13). So


Figure 13. A section of the product bundle $M \times \mathbb{R}^{k}$
there is a bijection
$\left\{\right.$ sections of $\left.M \times \mathbb{R}^{k} \rightarrow M\right\} \leftrightarrow\left\{\right.$ smooth functions $\left.f: M \rightarrow \mathbb{R}^{k}\right\}$.

Exercise 6.9. Show that the open Möbius strip in Example 6.5 as a line bundle over $S^{1}$ is not trivial. (Hint: if it were, it would have a non-vanishing global section, e.g., $s(p)=1$ for all $p \in S^{1}$. What is wrong with that?)

Remark 6.10. Given a global section $s \in \Gamma(E)$ and an open set $U \subseteq M$, we can always get a section $\left.s\right|_{U}$ over $U$ by restricting the domain of $s$. On the other hand, if $s \in \Gamma(U, E)$ is a section over $U$, then for every $p \in U$, we can find a global section $\bar{s}$ that agrees with $s$ over some neighborhood of $p$ (presumably contained in $U$ ) by multiplying $s$ with a bump function.

Definition 6.11. A frame for a vector bundle $E$ of rank $r$ over an open set $U \subseteq M$ is a collection of sections $e_{1}, \ldots, e_{r}$ of $E$ over $U$ such that at each point $p \in U$, the elements $e_{1}(p), \ldots, e_{r}(p)$ form a basis for the fiber $E_{p}$.

Proposition 6.12. A vector bundle $\pi: E \rightarrow M$ of rank $r$ is trivial if and only if it has a frame over $M$.

Proof. $(\Rightarrow)$ Suppose that $\pi: E \rightarrow M$ is trivial with a bundle isomorphism $\phi: E \rightarrow$ $M \times \mathbb{R}^{r}$. Let $u_{1}, \ldots, u_{r}$ be the standard basis for $\mathbb{R}^{r}$. Then for every $p \in M$, the elements $\left(p, u_{1}\right), \ldots,\left(p, u_{r}\right)$ form a basis for $\{p\} \times \mathbb{R}^{r}$. So the corresponding sections $e_{i}$ over $M$ defined by $e_{i}(p):=\phi^{-1}\left(p, u_{i}\right)$ form a basis for $E_{p}$.
$(\Leftarrow)$ Suppose that $e_{1}, \ldots, e_{r} \in \Gamma(E)$ is a frame over $M$. Then for any $p \in M$ and $e \in E_{p}$, we have $e=\sum_{i=1}^{r} a_{i} e_{i}(p)$ for some $a_{i} \in \mathbb{R}$. Now define

$$
\begin{aligned}
\phi: E & \longrightarrow M \times \mathbb{R}^{r} \\
e & \longmapsto\left(p, a_{1}, \ldots, a_{r}\right) .
\end{aligned}
$$

This is a bundle map with inverse

$$
\begin{aligned}
\psi: M \times \mathbb{R}^{r} & \longrightarrow E \\
\left(p, a_{1}, \ldots, a_{r}\right) & \longmapsto \sum_{i=1}^{r} a_{i} e_{i}(p) .
\end{aligned}
$$

## 7. Tangent Bundle

Definition 7.1. The tangent bundle of $M$, denoted $T M$, is the disjoint union of tangent spaces:

$$
T M:=\bigsqcup_{p \in M} T_{p} M
$$

together with a projection $\pi: T M \rightarrow M$ defined by $\pi(v)=p$ if $v \in T_{p} M$. We often denote an element of $T M$ as $(p, v)$, meaning $v \in T_{p} M$.

The manifold structure on $T M$ induced by the structure on $M$. For a chart $(U, h)$ on $M$, we can form a chart $\left(\pi^{-1}(U), \widetilde{h}\right)$ on $T M$ with $\widetilde{h}$ defined by

$$
\begin{aligned}
& \widetilde{h}: \pi^{-1}(U) \longrightarrow h(U) \times \mathbb{R}^{n} \subseteq \mathbb{R}^{n} \times \mathbb{R}^{n} \\
& v \in T_{p} M \longmapsto(h(p), v(U, h)) .
\end{aligned}
$$

Let $(U, h)$ and $(V, k)$ be charts at $p \in M$. The transition function of the corresponding charts in $T M$ is given by $\left(k \circ h^{-1}, D\left(k \circ h^{-1}\right)\right)$.


The differentiable structure on $T M$ is chosen to be the maximal atlas containing these charts. To put a topology on $T M$, we say that $A \subseteq T M$ is open if $\widetilde{h}\left(\pi^{-1}(U) \cap\right.$ $A$ ) is open for each chart $(U, h)$. Then $T M$ being Hausdorff and second countable follows. Note that with this structure, the mapping $\pi$ is differentiable.

Exercise 7.2. Check that $T M$ is Hausdorff and second countable. Conclude that $T M$ is a differentiable manifold and that $\pi: T M \rightarrow M$ is a vector bundle over $M$ of rank $n$ (recall the vector space structure of $T_{p} M$ ).

Example 7.3. Let $M=\mathbb{R}^{n}$. Fix the chart ( $\mathbb{R}^{n}$, id) for $M$. This induces the map $T_{p} M \cong \mathbb{R}^{n}$ for all $p \in M$. Then $T M=\pi^{-1}\left(\mathbb{R}^{n}\right) \cong \mathbb{R}^{n} \times \mathbb{R}^{n}$. We can think of $T \mathbb{R}^{n}$ as the result of attaching a copy of $\mathbb{R}^{n}$ to each point in $\mathbb{R}^{n}$.

Exercise 7.4. Consider the circle $S^{1}:=\left\{(x, y) \in \mathbb{R}^{2} \mid x^{2}+y^{2}=1\right\}$. The following picture leads one to think that $T S^{1}$ is trivial:


Prove this fact. [Hint: see Proposition 6.12, parametrize $S^{1}$, and use this parametrization to define a derivation $v(p)$ smoothly varying with $p$ and never equal to the zero derivation.]

Definition 7.5. A global section $s: M \rightarrow T M$ of the tangent bundle $T M$ is called a vector field.

Remark 7.6. It is not true that the tangent bundle is always trivial. For example, consider $M=S^{2}$. The "hairy ball" theorem says that there cannot be nonvanishing continuous vector fields on $M$. That is, if $s \in \Gamma\left(T S^{2}\right)$ is a global section, then there must be some $p \in M$ such that $s(p)=0$. It follows that $T S^{2}$, unlike $T S^{1}$, is non-trivial.

Remark 7.7. Let $s: M \rightarrow T M$ be a vector field. The zero locus of $s$ is the set $\{p \in M \mid s(p)=0\}$. Note that this is well-defined since $s(p)=0$ is not dependent on the chart.

Let $f: M \rightarrow N$ be a differentiable map of manifolds. Recall again in section 4 that we defined $d f_{p}: T_{p} M \rightarrow T_{p} N$, the differential of $f$ at a point $p \in M$. Now we use this information to induce a bundle map $f_{*}: T M \rightarrow T N$ of tangent bundles and define $d f$, the differential of $f$ to be this induced map.

Take $p \in M$ and a chart $(U, h)$ at $p$. Let $(V, k)$ be a chart at $f(p)$ such that $f(U) \subseteq V$. Let $\left(\frac{\partial}{\partial x_{1}}\right)_{p}, \ldots,\left(\frac{\partial}{\partial x_{m}}\right)_{p}$ and $\left(\frac{\partial}{\partial y_{1}}\right)_{f(p)}, \ldots,\left(\frac{\partial}{\partial y_{n}}\right)_{f(p)}$ denote the bases for $T_{p} M$ and $T_{f(p)} N$ correspondingly. Recall that for $d f_{p}$, we have the following diagram:

$$
\begin{aligned}
& T_{p} M \xrightarrow{d f_{p}} T_{f(p)} N \\
& \underset{\mathbb{R}_{J\left(k \circ f \circ h^{-1}\right)(h(p))}^{\downarrow} \stackrel{\text { ® } \frac{\partial f_{i}}{\partial x_{j}}(p) "}{\mathbb{R}^{n}}}{\downarrow} .
\end{aligned}
$$

To define $f_{*}$ on $T M$, we first define $f_{*}$ locally on the chart $\pi^{-1}(U)$ :


Then we can "glue" the pieces together since $f_{*}$ commutes with the transition maps: Take two pairs of charts $\left(U_{i}, h_{i}\right)$ and $\left(V_{i}, k_{i}\right)$ with $f\left(U_{i}\right) \subseteq V_{i}$ for $i=1,2$. Let $\widetilde{f}_{i}$ denote the composition $k_{i} \circ f \circ h_{i}^{-1}$ and observe that the following diagram commutes.

$$
\begin{aligned}
& h_{1}\left(U_{1} \cap U_{2}\right) \times \mathbb{R}^{m} \xrightarrow{\left(\widetilde{f}_{1}, J \widetilde{f}_{1,(h(p))}\right)} k_{1}\left(V_{1} \cap V_{2}\right) \times \mathbb{R}^{n} \\
& \quad h_{2} \circ h_{1}^{-1} \downarrow \\
& h_{2}\left(U_{1} \cap U_{2}\right) \times \mathbb{R}^{m} \xrightarrow\left[\left(\widetilde{\left.f_{2}, J \widetilde{f_{2,(h(p))}}\right)}\right]{ } k_{2}\left(V_{1} \cap V_{1}^{-1}\right) \times \mathbb{R}^{n} .\right.
\end{aligned}
$$

Thus, these local mappings glue to define the differential:


Example 7.8. Let $M=\mathbb{R}^{2}$ and $N=\mathbb{R}^{3}$, and consider the function

$$
\begin{aligned}
f: M & \rightarrow N \\
(x, y) & \mapsto\left(x, y, x^{2}-x^{2}\right)
\end{aligned}
$$

Choosing the charts $\left(\mathbb{R}^{2}, \mathrm{id}_{\mathbb{R}^{2}}\right)$ and $\left(\mathbb{R}^{3}, \mathrm{id}_{\mathbb{R}^{3}}\right)$, the commutative diagram (2) becomes

with

$$
\begin{aligned}
d f: \mathbb{R}^{2} \times \mathbb{R}^{2} & \rightarrow \mathbb{R}^{3} \times \mathbb{R}^{3} \\
((x, y),(u, v)) & \mapsto\left(\left(x, y, x^{2}-y^{2}\right),(u, v, 2 x u-2 y v)\right)
\end{aligned}
$$

since

$$
J f(x, y)(u, v)=\left(\begin{array}{cc}
1 & 0 \\
0 & 1 \\
2 x & -2 y
\end{array}\right)\binom{u}{v}=\left(\begin{array}{c}
u \\
v \\
2 x u-2 y v
\end{array}\right)
$$

## 8. The Algebra of Differential Forms

Let $M$ be a smooth manifold of dimension $n$ and let $p \in M$. The cotangent space to $M$ at $p$ is the dual space $T_{p}^{*} M=\left\{f: T_{p} M \rightarrow \mathbb{R} \mid f\right.$ linear $\}$. We define the cotangent bundle for $M$ using the same procedure we used for the tangent bundle. Start by defining $T^{*} M$ as the disjoint union:

$$
T^{*} M:=\bigsqcup_{p \in M} T_{p}^{*} M
$$

To define the manifold/vector bundle structure on $T^{*} M$, for each chart $(U, h)$ on $M$ we chose the standard basis $\left(\partial / \partial x_{i}\right)_{p}$ for $i=1, \ldots, n$ for tangent space with respect to that chart at each point $p \in U$. The gluing instructions were provided by the linear mappings $D_{p}\left(k \circ h^{-1}\right): \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$, where $k \circ h^{-1}$ is a transition function for $M$. For the cotangent bundle, we choose the dual basis $d x_{i, p}:=\left(\partial / \partial x_{i}\right)_{p}^{*}$. Thus,

$$
d x_{i, p}\left(\left(\frac{\partial}{\partial x_{j}}\right)_{p}\right)=\delta(i, j)= \begin{cases}1 & \text { if } i=j \\ 0 & \text { if } i \neq j\end{cases}
$$

The gluing instructions for the bundle are provided by the dual mapping $D_{p}(k \circ$ $\left.h^{-1}\right)^{*}$. We can take the charts for $T^{*} M$ so be the same as those for $T M$ by identifying $T_{p} M$ and $T_{p}^{*} M$ in local coordinates via the mapping $\left(\partial / \partial x_{i}\right)_{p} \mapsto d x_{i, p}$. Then, if $(U, h))$ and $(V, k)$ are overlapping charts on $M$, the corresponding transition function for $T^{*} M$ is given by


Similarly, we can define vector bundles of the form $\operatorname{Sym}^{\ell} T M, \Lambda^{\ell} T M, \Lambda^{\ell} T^{*} M$, $T M \otimes T^{*} M$, etc. For the special case of $\Lambda^{n} T^{*} M$, where $n=\operatorname{dim} M$, the transition function for charts $(U, h)$ and $(V, k)$ is as displayed above for $T^{*} M$ except that $\mathbb{R}^{n}$ is replaced by $\mathbb{R} \approx \Lambda^{n} \mathbb{R}^{n}$ and $J\left(k \circ h^{-1}\right)^{T}(p)$ is replaced by $\operatorname{det}\left(J\left(k \circ h^{-1}\right)^{T}(p)\right)=$ $\operatorname{det}\left(J\left(k \circ h^{-1}\right)(p)\right)$, i.e., multiplication by the determinant of the transition function at each point $p$.

Definition 8.1. A $k$-form on $M$ is a global section of $\Lambda^{k} T^{*} M$, the vector bundle over $M$ where the fiber over $p \in M$ is $\Lambda^{k} T_{p}^{*} M$, the $k$-th exterior power of the cotangent space at $p$. We use $\Omega^{k} M$ to denote the vector space of $k$-forms.

Fixing a chart $(U, h)$ and bases as above, a basis for $\Lambda^{k} T_{p}^{*} M$ is provided by

$$
\left\{d x_{\mu, p}:=d x_{\mu_{1}, p} \wedge \cdots \wedge d x_{\mu_{k}, p} \mid 1 \leq \mu_{1}<\cdots<\mu_{k} \leq n\right\}
$$

where the subscript $p$ is sometimes dropped for convenience. Now consider a $k$-form $\omega: M \rightarrow \Lambda^{k} T * M$. In local coordinates we get

$$
\omega(p)=\sum_{\mu} \omega_{\mu}(p) d x_{\mu}=\sum_{\mu} \omega_{\mu}\left(p_{1}, \ldots, p_{n}\right) d x_{\mu_{1}} \wedge \cdots \wedge d x_{\mu_{k}} \in \Lambda^{k} T_{p}^{*} M
$$

where each function $\omega_{\mu}: h(U) \rightarrow \mathbb{R}$ is differentiable. (In the above displayed equation, we are abusing notation slightly by using $p$ to denote what really should be $h(p)$.)

We note that in the special case where $k=n=\operatorname{dim} M$, then changing basis affects the local form of $\omega$ via the determinant of the transition function.
8.1. The pullback of a differential form by a smooth map. A smooth map $f: M \rightarrow N$ of manifolds induces a map $\Omega^{k} f: \Omega^{k} N \rightarrow \Omega^{k} M$ of $k$-forms. To describe this map in coordinates, let us fix some $p \in M$ and pick charts $(U, h)$ at $p$ and $(V, k)$ at $f(p)$ such that $f(U) \subseteq V$.

First note that $f: M \rightarrow N$ induces a map of tangent spaces $d f_{p}: T_{p} M \rightarrow T_{p} N$. Taking its dual gives us a map of cotangent spaces (in the opposite direction) $d f_{p}^{*}: T_{p}^{*} N \rightarrow T_{p}^{*} M$.

To be more specific, let us start with the case $N=\mathbb{R}$ and suppose that $g: M \rightarrow \mathbb{R}$ is differentiable. We can decompose the induced map $d g_{p}: T_{p} M \rightarrow T_{g(p)} \mathbb{R} \cong \mathbb{R}$ as a linear combination of the standard basis elements of $T_{p}^{*} M$ using the following push-forward mapping:


A basis vector $\left(\frac{\partial}{\partial x_{i}}\right)_{p}$ of $T_{p} M$ is sent to $e_{i}$, the $i$-th standard basis vector of $\mathbb{R}^{n}$, which is then sent to $\frac{\partial g}{\partial x_{i}}(p)=\frac{\partial \mathrm{id}_{\mathbb{R}} \circ g \circ h^{-1}}{\partial x_{i}}(h(p))$. Thus $d g_{p} \in T_{p}^{*} M$ and in coordinates, we can simply write

$$
d g_{p}=\sum_{i=1}^{n}\left(\frac{\partial g}{\partial x_{i}}\right)(p) d x_{i, p}
$$

Example 8.2. For example, if we have $g: \mathbb{R}^{2} \rightarrow \mathbb{R}$ with $g(x, y)=x y^{2}+y$, then at $p=(1,1), d g=y^{2} d x+(2 x y+1) d y$ and $d g_{p}=d x+3 d y$.

Now let us return to the function we started with, a smooth map $f: M \rightarrow N$ of manifolds. Suppose that $M$ has dimension $m$ and $N$ has dimension $n$. What does the previous case say about the induced map $d f_{p}^{*}$ ? Again recall the diagram for $d f_{p}$ :


Taking its dual gives us the following diagram:


This means that $d f_{p}^{*}\left(d y_{i, p}\right)$ is the $i$-th column of the matrix $J f_{p}^{\top}$, which is the $i$-th row of the Jacobian $J f_{p}$. In this way we have

$$
d f_{p}^{*}\left(d y_{i, p}\right)=\nabla f_{i}(p)=\sum_{j=1}^{m} \frac{\partial f_{i}}{\partial x_{j}}(p) d x_{j}=d\left(f_{i}\right)_{p} \stackrel{\text { call }}{=} d f_{i, p}
$$

It follows that $d f_{p}^{*}$ induces a map

$$
\begin{aligned}
\Lambda^{k} d f_{p}^{*}: \Lambda^{k} T_{f(p)}^{*} N & \longrightarrow \Lambda^{k} T_{p}^{*} M \\
d y_{\mu, p} & \longmapsto d f_{\mu_{1}, p} \wedge \cdots \wedge d f_{\mu_{k}, p} \stackrel{\text { call }}{=} d f_{\mu, p}
\end{aligned}
$$

Now take a $k$-form $\omega: N \rightarrow \Lambda^{k} T^{*} N$ on $N$. Define

$$
\begin{aligned}
\Omega^{k} f_{p} \omega \stackrel{\text { call }}{=} f_{p}^{*} \omega:=\Lambda^{k} d f_{p}^{*} \circ \omega \circ f: M & \longrightarrow \Lambda^{k} T^{*} M \\
p & \longmapsto \sum_{\mu} \omega_{\mu}(f(p)) d f_{\mu, p}
\end{aligned}
$$

where $\omega_{\mu}(f(p))$ is the coefficient of $d y_{\mu, p}$ in $\omega$. Finally, we can glue everything together and obtain a corresponding $k$-form $f^{*} \omega: M \rightarrow \Lambda^{k} T^{*} M$ on $M$ as shown in the following diagram.


And locally, we have

$$
f_{p}^{*}\left(\sum_{\mu} w_{\mu}(f(p)) d y_{\mu, f(p)}\right)=\sum_{\mu} w_{\mu}(f(p)) d f_{\mu, p}
$$

Example 8.3. Define

$$
\begin{aligned}
f: \mathbb{R}^{2} & \longrightarrow \mathbb{R}^{3} \\
(u, v) & \longmapsto\left(u^{2}-v, u+2 v, v^{2}\right)
\end{aligned}
$$

and consider $\omega=x^{2} d x \wedge d y+(x+z) d y \wedge d z \in \Omega^{2} \mathbb{R}^{3}$. We can compute $f^{*} \omega$ :

$$
\begin{aligned}
\left(f^{*} \omega\right)(d u \wedge d v)= & \omega_{x, y}(f(u, v)) d f_{x, y}+\omega_{y, z}(f(u, v)) d f_{y, z} \\
= & \left(u^{2}-v\right)^{2} d\left(u^{2}-v\right) \wedge d(u+2 v)+\left(u^{2}-v+v^{2}\right) d(u+2 v) \wedge d\left(v^{2}\right) \\
= & \left(u^{4}-2 u^{2} v+v^{2}\right)(2 u d u-d v) \wedge(d u+2 d v) \\
& +\left(u^{2}-v+v^{2}\right)(d u+2 d v) \wedge(2 v d v) \\
= & \cdots
\end{aligned}
$$

$$
=\left(4 u^{5} v+u^{4}-8 u^{3} v^{2}+4 u v^{3}+2 v^{3}-v^{2}\right) d u \wedge d v
$$

Exercise 8.4. Note that there is an isomorphism $\Lambda^{1} T^{*} M \cong T^{*} M$. So for $\omega$ a 1-form on $M$ and any point $p \in M, \omega(p)$ is the same as a smooth, real-valued function on the tangent space $T_{p} M$. Consider a differentiable map $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$. Let $\omega=d x \in \Omega^{1} \mathbb{R}$. What is the pullback $f^{*} \omega$ ?
8.2. The exterior derivative. The next goal is to associate to each manifold $M$ a cochain complex

$$
0 \rightarrow \Omega^{0} M \xrightarrow{d} \Omega^{1} M \xrightarrow{d} \Omega^{2} M \xrightarrow{d} \Omega^{3} M \xrightarrow{d} \cdots
$$

called the de Rham complex. For each $k \geq 0$, we will construct a linear operator $d: \Omega^{k} M \rightarrow \Omega^{k+1} M$, called the exterior derivative, such that $d^{2}=0$ (and $d$ behaves well under change of coordinates). The exterior derivative and the product on exterior products discussed earlier (via concatentation) makes $\Omega^{\bullet} M:=\bigoplus_{k \geq 0} \Omega^{k} M$ into what is called a graded, anti-commutative, differential algebra.

Theorem 8.5. Let $M$ be a manifold of dimension $n$. There exists a unique sequence of linear maps

$$
\begin{equation*}
0 \rightarrow \Omega^{0} M \xrightarrow{d} \Omega^{1} M \xrightarrow{d} \Omega^{2} M \xrightarrow{d} \Omega^{3} M \xrightarrow{d} \cdots \tag{3}
\end{equation*}
$$

such that:
(i) If $f \in \Omega^{0} M$, i.e., $f: M \rightarrow \mathbb{R}$, then $d f$ is the normal differential;
(ii) Equation (3) is a complex, i.e., $d^{2}=d \circ d=0$;
(iii) $d$ satisfies the product rule: $d(\omega \wedge \eta)=d \omega \wedge \eta+(-1)^{r} \omega \wedge d \eta$ for $\omega \in \Omega^{r} M$. In local coordinates, $d$ is given by

$$
\begin{equation*}
d\left(\sum_{\mu} a_{\mu} d x_{\mu}\right)=\sum_{\mu} \sum_{i=1}^{n} \frac{\partial a_{\mu}}{\partial x_{i}} d x_{i} \wedge d x_{\mu} \tag{4}
\end{equation*}
$$

Proof. Suppose that $d: \Omega^{k} M \rightarrow \Omega^{k+1} M$ is a linear map satisfying the above three conditions. We first prove uniqueness of $d$ by showing that it has to be given by equation (4) in any local coordinates.

Let $\omega$ be a $k$-form and suppose that $\omega=\sum_{\mu} \omega_{\mu} d x_{\mu}$ using local coordinates on $U$. Using all three properties of $d$, we have

$$
\begin{aligned}
d \omega & =d\left(\sum_{\mu} \omega_{\mu} d x_{\mu}\right) \\
& =\sum_{\mu}\left(d\left(\omega_{\mu} d x_{\mu}\right)\right) \\
& =\sum_{\mu}\left(d\left(\omega_{\mu}\right) \wedge d x_{\mu}+(-1)^{0} \omega_{\mu}\left(d\left(d x_{\mu}\right)\right)\right) \\
& =\sum_{\mu}\left(\left(\sum_{i=1}^{n} \frac{\partial \omega_{\mu}}{\partial x_{i}} d x_{i} \wedge d x_{\mu}\right)+0\right) \\
& =\sum_{\mu} \sum_{i=1}^{n} \frac{\partial \omega_{\mu}}{\partial x_{i}} d x_{i} \wedge d x_{\mu}
\end{aligned}
$$

Now we show that the definition in equation (4) satisfies the three properties, proving existence of such an operator. Note that $d f$ being the normal differential follows immediately from the definition. To see that $d f^{2}=0$, without loss of generality, consider the form $a d x_{\mu}$ and compute:

$$
d^{2}\left(a d x_{\mu}\right)=d\left(\sum_{j=1}^{n} \frac{\partial a}{\partial x_{j}} d x_{j} \wedge d x_{\mu}\right)=\sum_{i, j=1}^{n} \frac{\partial^{2} a}{\partial x_{i} \partial x_{j}} d x_{i} \wedge d x_{j} \wedge d x_{\mu}
$$

If $i=j$, then $d x_{i} \wedge d x_{j}=0$. If not, then $\frac{\partial^{2} a}{\partial x_{i} \partial x_{j}} d x_{i} \wedge d x_{j}=-\frac{\partial^{2} a}{\partial x_{j} \partial x_{i}} d x_{j} \wedge d x_{i}$.
To show that $d f$ satisfies the product rule, first note that for any zero-form $f \in \Omega^{0} M$, by definition we have

$$
d\left(f d x_{\mu}\right)=(d f) \wedge d x_{\mu}
$$

Now consider a $k$-form $\omega=f d x_{\mu}$ and any other differential form $\eta=g d x_{\nu}$. Using the usual product rule, we compute:

$$
\begin{aligned}
d(\omega \wedge \eta) & =d\left(f d x_{\mu} \wedge g d x_{\nu}\right) \\
& =d\left(f g d x_{\mu} \wedge d x_{\nu}\right) \\
& =d(f g) \wedge d x_{\mu} \wedge d x_{\nu} \\
& =(g d f+f d g) \wedge d x_{\mu} \wedge d x_{\nu} \\
& =\left(d f \wedge d x_{\mu}\right) \wedge\left(g d x_{\nu}\right)+(-1)^{k}\left(f d x_{\mu}\right) \wedge\left(d g \wedge d x_{\nu}\right)
\end{aligned}
$$

where the $(-1)^{k}$ comes from $d g \wedge d x_{\mu}=(-1)^{k} d x_{\mu} \wedge d g$.
We have now shown that for each chart $(U, h)$ in an atlas for $M$, we have an operator $d_{U}$ satisfying the required properties. On overlaps, these operators must agree by uniqueness, and hence they glue together to define a global operator $d$.

Definition 8.6. The operator $d f$ is called exterior differentiation, and $d \omega$ is called the exterior derivative of $\omega$.

Lemma 8.7. Let $f: M \rightarrow N$ be a smooth map of manifolds. Then $f^{*}$, the pullback by $f$, commutes with $d$. That is, for any $\omega \in \Omega^{k} N$, we have

$$
f^{*}(d \omega)=d\left(f^{*} \omega\right)
$$

Exercise 8.8. Prove the above lemma.
The fact that

$$
0 \rightarrow \Omega^{0} M \xrightarrow{d} \Omega^{1} M \xrightarrow{d} \Omega^{2} M \xrightarrow{d} \Omega^{3} M \xrightarrow{d} \cdots
$$

is a sequence of mappings of abelian groups (even of vector spaces) and that $d^{2}=0$ says that the sequence forms a cochain complex. ${ }^{5}$ A mapping of cochain complexes is a sequence of homomorphisms between spaces with the same indices and commuting with the corresponding $d$ mappings. For instance, Lemma 8.7 tells us that the pullback is a cochain map. In detail, given $f: M \rightarrow N$ we get the following commutative diagram:

[^4]

We might abbreviate the above by $f^{*}: \Omega^{\bullet} N \rightarrow \Omega^{\bullet} M$.
Lemma 8.9. Pullbacks respect compositions: applying pullbacks to a commutative diagram of mappings of manifolds

induces, for each $k \geq 0$, the commutative diagram


Together with the fact that the pullback by the identity function on a manifold is the identity on $k$-forms, we have a just seen that pullback of differential forms gives a contravariant functor from the category of smooth manifolds to the category of cochain complexes.

Definition 8.10. A $k$-form $\omega$ is exact if there is a $(k-1)$-form $\eta$ such that $d \eta=\omega$. It is closed if $d \omega=0$.

Note that an exact form is closed since $d^{2}=0$. We will have much more to say about exactness when we discuss DeRham cohomology, later.

## 9. Oriented Manifolds

In ordinary integration of a one-variable function, we have

$$
\int_{0}^{1} f d x=-\int_{1}^{0} f d x
$$

This formula is a first hint at the role played by an orientation of the underlying manifold. In ordinary integration in several variables, reordering the standard basis vectors gives a change of basis whose Jacobian is the corresponding permutation matrix. Thus, via the change of variables formula for integration, the sign of the integral will change by the sign of this permutation. Thus, before we can talk about (coordinate-free) integration on a manifold, we need to add the additional structure of an orientation.

Definition 9.1. Let $V$ be a real vector space. Two ordered bases $\left\langle v_{1}, \ldots, v_{n}\right\rangle$ and $\left\langle w_{1}, \ldots, w_{n}\right\rangle$ have the same orientation if the mapping $V \rightarrow V$ that sends $v_{i}$ to $w_{i}$ has positive determinant.

The property of having the same orientation defines an equivalence relation on the set of ordered bases of $V$ with two equivalence classes, each of which is called an orientation on $V$. Having chosen an orientation $\mathcal{O}$, we get an orientated vector space $(V, \mathcal{O})$.

Ordered bases in $\mathcal{O}$ are said to be positively oriented; otherwise they are negatively oriented.

Example 9.2. Let $V=\mathbb{R}^{3}$ and let $e_{1}, e_{2}, e_{3}$ denote the standard basis of $\mathbb{R}^{3}$. The six possible ordered bases formed from these vectors fall into two equivalence classes:

$$
\begin{aligned}
& \left\langle e_{1}, e_{2}, e_{3}\right\rangle \sim\left\langle e_{2}, e_{3}, e_{1}\right\rangle \sim\left\langle e_{3}, e_{1}, e_{2}\right\rangle \\
& \left\langle e_{2}, e_{1}, e_{3}\right\rangle \sim\left\langle e_{1}, e_{3}, e_{2}\right\rangle \sim\left\langle e_{3}, e_{2}, e_{1}\right\rangle
\end{aligned}
$$

Exercise 9.3. Draw the six frames of the form

for the vectors given in the previous example. How is orientation reflected in the geometry of these frames.

Note that if we have $\mathbb{R}^{n}$ with the standard basis $e_{1}, \ldots, e_{n}$, then an ordered basis formed using these vectors can be identified with a permutation $\sigma \in S_{n}$ on $n$ letters. In this way, the ordered basis $\left\langle e_{\sigma(1)}, \ldots, e_{\sigma(n)}\right\rangle$ has the same orientation as $\left\langle e_{1}, \ldots, e_{n}\right\rangle$ if $\operatorname{sign}(\sigma)=1$.

Also, for example, $\left\langle 2 e_{1}+e_{2}, e_{2}, e_{3}\right\rangle \sim\left\langle e_{1}, e_{2}, e_{3}\right\rangle$ since

$$
\operatorname{det}\left[\begin{array}{lll}
2 & 0 & 0 \\
1 & 1 & 0 \\
0 & 0 & 1
\end{array}\right]=2>0
$$

In order to integrate on manifolds, we want the tangent spaces to be oriented in a coherent way.

Definition 9.4. Let $M$ be a manifold of dimension $n$. A family $\left\{\mathcal{O}_{p}\right\}_{p \in M}$ of orientations $\mathcal{O}_{p}$ of the tangent space $T_{p} M$ is locally coherent if around every point of $M$ there is an orientation-preserving chart $(U, h)$, meaning that for every $u \in U$, the differential

$$
d h_{u}: T_{u} M \xrightarrow{\cong} \mathbb{R}^{n}
$$

takes the orientation $\mathcal{O}_{u}$ to the usual orientation of $\mathbb{R}^{n}$ (the orientation of $\left\langle e_{1}, \ldots, e_{n}\right\rangle$ ).
Definition 9.5. An orientation of a manifold $M$ is a locally coherent family $\mathcal{O}=$ $\left\{\mathcal{O}_{p}\right\}_{p \in M}$ of orientations of its tangent spaces. An oriented manifold is a pair $(M, \mathcal{O})$ consisting of a manifold $M$ and an orientation $\mathcal{O}$ of $M$.

Definition 9.6. A diffeomorphism $f: M \rightarrow N$ is orientation-preserving if $d f_{p}: T_{p} M \rightarrow$ $T_{f(p)} N$ is orientation-preserving for all $p \in M$.

Definition 9.7. An atlas $\mathfrak{A}$ of $M$ is called an orienting atlas if all its transition maps $w$ are orientation-preserving, that is, if $\operatorname{det}(J w(x))>0$ for all appropriate $x$.

Remark 9.8. If $M$ is already oriented, then the orientation of $M$ gives a maximal orienting atlas. Conversely, if $\mathfrak{A}$ is an orienting atlas, then there is a unique orientation of $M$ such that $\mathfrak{A}$ consists of orientation-preserving charts. (Note that "orientation-preserving" means different things for charts and for diffeomorphisms.)

Exercise 9.9. Convince yourself that the above remark is true.
Example 9.10. The $n$-sphere is orientable. To see this, we use charts given by stereographic projection. For example, we look at the circle $S^{1}$ centered at the origin $(0,0)$ as shown in Figure 14, with the arrow indicating the chosen orientation (counterclockwise). Let $x, y$ denote the coordinates of $\mathbb{R}^{2}$. We can either project $S^{1}$


Figure 14. $S^{1}$ with the counterclockwise orientation.
onto the line $y=-1$ from the north pole $(0,1)$ or project $S^{1}$ onto the line $y=1$ from the south pole $(0,-1)$ as shown in Figure 15. In this way we obtain two charts covering $S^{1}$. Let $U^{+}=S^{1} \backslash\{(0,1)\}, U^{-}=S^{1} \backslash\{(0,-1)\}$, and let $\phi^{+}$and $\phi^{-}$ denote the corresponding projection maps. Note that both $\phi^{+}\left(U^{+}\right)$and $\phi^{-}\left(U^{-}\right)$ are isomorphic to $\mathbb{R}$ and the transition map from $U^{+}$to $U^{-}$is smooth (check this!). However, the transition map is not orientation-preserving. Notice that the orientation of the image $\phi^{-}\left(U^{-}\right)$is different from that of $\phi^{+}\left(U^{+}\right)$. Hence the charts


Figure 15. Stereographic projection of $S^{1}$ from the north pole and from the south pole.
$\left(U^{+}, \phi^{+}\right),\left(U^{-}, \phi^{-}\right)$do not form an orienting atlas. To resolve the problem, we can change $\phi^{-}$to $-\phi^{-}$and use the charts $\left(U^{+}, \phi^{+}\right),\left(U^{-},-\phi^{-}\right)$instead.

Example 9.11. The open Möbius strip in Example 6.5 is non-orientable. The projective space $\mathbb{P}^{n}$ is orientable if and only if $n$ is odd, e.g. $\mathbb{P}^{1} \approx S^{1}$. We will later develop better tools for proving non-orientability.

## 10. Integration of Forms

For a quick introduction to the Lebesgue integral, see Appendix C.
Let $M$ be an oriented manifold of dimension $n$ and let $\omega \in \Omega^{n} M$ be an $n$-form. In order to integrate $\omega$ on $M$ and compute the integral $\int_{\omega} M$, we want to break $M$ into nice "cells" and integrate $\omega$ on these cells with respect to local coordinates.

### 10.1. Definition and properties of the integral.

Definition 10.1. A subset $X \subseteq M$ is measurable if $h(X \cap U) \subseteq \mathbb{R}^{n}$ is measurable for all charts $(U, h)$. We say that $X$ has measure zero if $h(U \cap X) \subseteq \mathbb{R}^{n}$ has measure zero for all charts $(U, h)$.

To define the integral, we cover $M$ by a countable collection of disjoint measurable sets $\left\{A_{i}\right\}_{i}$ and orientation-preserving charts $\left\{\left(U_{i}, h_{i}\right)\right\}_{i}$ with $A_{i} \subseteq U_{i}$. (Recall that $M$ is second countable.) Locally, with respect to each $\left(U_{i}, h_{i}\right)$, we have

$$
\omega(p)=\widetilde{a}_{i}(p) d x_{1, p} \wedge \cdots \wedge d x_{n, p}
$$

for some $\widetilde{a}_{i}: U_{i} \rightarrow \mathbb{R}$. Define $a_{i}: h_{i}\left(U_{i}\right) \rightarrow \mathbb{R}$ by $a_{i}=\widetilde{a}_{i} \circ h_{i}^{-1}$.
Remark 10.2. To see that the sets $\left\{A_{i}\right\}$ and $\left\{U_{i}\right\}$ with the claimed properties always exist, start with any orienting atlas $\mathfrak{V}=\left\{\left(V_{\alpha}, k_{\alpha}\right)\right\}$ (this is possible since we are assuming $M$ is orientable). Since $M$ is second-countable, it has a countable basis $\mathcal{B}$ for its topology. For each $p \in M$, take a chart $\left(V_{\alpha}, k_{\alpha}\right)$ at $p$, and then, since $\mathcal{B}$ is a basis, we can find $U \in \mathcal{B}$ such that $p \in U \subseteq V_{\alpha}$. Save the chart $(U, h)$ where $h:=\left.k_{\alpha}\right|_{U}$. In the end, since $\mathcal{B}$ is countable, so is the set of charts we saved. Thus, we have new countable orienting atlas $\mathfrak{U}=\left\{\left(U_{i}, h_{i}\right)\right\}$. Now let $A_{1}:=U_{1}$, and let $A_{i+1}:=U_{i+1} \backslash \cup_{j=1}^{i} A_{j}$ for $i \geq 1$.
Definition 10.3. Using the notation above, we say that $\omega$ is integrable if each $a_{i}: h_{i}\left(U_{i}\right) \rightarrow \mathbb{R}$ is integrable on $h\left(A_{i}\right)$ and if $\sum_{i} \int_{h\left(A_{i}\right)}\left|a_{i}\right|<\infty$. In this case, we define the integral to be the sum:

$$
\int_{M} \omega:=\sum_{i} \int_{h_{i}\left(A_{i}\right)} a_{i}
$$

Theorem 10.4. Let $M$ be a manifold of dimension $n$ and let $\omega \in \Omega^{n} M$ be integrable. Then $\int_{M} \omega$ is independent of choice of charts.
Proof. Suppose that there is another countable collection $\left\{B_{j}\right\}_{j}$ of disjoint measurable sets that cover $M$ and that there are orientation-preserving charts $\left\{\left(V_{j}, k_{j}\right)\right\}$ such that $B_{j} \subseteq V_{j}$. Locally, with respect to each ( $V_{j}, k_{j}$ ), write

$$
\omega=\widetilde{b}_{j} d y_{1} \wedge \cdots \wedge d y_{n}
$$

for some $\widetilde{b}_{j}: V_{j} \rightarrow \mathbb{R}$ and define $b_{j}: k_{j}\left(V_{j}\right) \rightarrow \mathbb{R}$ by $b_{j}=\widetilde{b}_{j} \circ k_{j}^{-1}$. We want to show that: (1) $b_{j}$ is integrable on $k_{j}\left(B_{j}\right) ;(2) \sum_{j} \int_{k_{j}\left(B_{j}\right)}\left|b_{j}\right|<\infty$; and (3) $\sum_{i} \int_{h_{i}\left(A_{i}\right)} a_{i}=\sum_{j} \int_{k_{j}\left(B_{j}\right)} b_{j}$.

First we look at the intersection $A_{i} \cap B_{j}$ and explicitly apply the change-ofvariable theorem. Let $\phi_{i, j}=h_{i} \circ k_{j}^{-1}$ be the transition function, and recall that we have the following diagrams:


Now take the $n$-th exterior power, and the transition map becomes multiplication by $\operatorname{det}\left(J \phi_{i, j}\left(k_{j}(p)\right)\right)$. Since $\Lambda^{n} \mathbb{R}^{n} \cong \mathbb{R}$, we have the following diagram:


This means that

$$
d x_{1, p} \wedge \cdots \wedge d x_{n, p}=\operatorname{det}\left(\left(J \phi_{i, j}\right)_{k_{j}(p)}\right) d y_{1, p} \wedge \cdots \wedge d y_{n, p}
$$

Also, that for any $p \in A_{i} \cap B_{j}$, we have

$$
\omega(p)=\widetilde{a}_{i}(p) d x_{1, p} \wedge \cdots \wedge d x_{n, p}=\widetilde{b}_{j}(p) d y_{1, p} \wedge \cdots \wedge d y_{n, p}
$$

which gives us

$$
\widetilde{a}_{i}(p) \operatorname{det}\left(\left(J \phi_{i, j}\right)_{k_{j}(p)}\right)=\widetilde{b}_{j}(p)
$$

Therefore,

$$
\begin{aligned}
b_{j}\left(k_{j}(p)\right) & =\widetilde{b}_{j}(p) \\
& =\widetilde{a}_{i}(p) \operatorname{det}\left(\left(J \phi_{i, j}\right)_{k_{j}(p)}\right) \\
& =a_{i}\left(h_{i}(p)\right) \operatorname{det}\left(\left(J \phi_{i, j}\right)_{k_{j}(p)}\right) \\
& =\left(a_{i} \circ \phi_{i, j}\right)\left(k_{j}(p)\right) \operatorname{det}\left(\left(J \phi_{i, j}\right)_{k_{j}(p)}\right)
\end{aligned}
$$

i.e.,

$$
b_{j}=\left(a_{i} \circ \phi_{i, j}\right) \cdot \operatorname{det}\left(J \phi_{i, j}\right)
$$

Now compute, applying the change of variables theorem:

$$
\begin{aligned}
\int_{M} \omega=\sum_{i} \int_{h_{i}\left(A_{i}\right)} a_{i} & =\sum_{i} \sum_{j} \int_{h_{i}\left(A_{i} \cap B_{j}\right)} a_{i} \\
& =\sum_{i} \sum_{j} \int_{\left(\phi_{i, j} \circ k_{j}\right)\left(A_{i} \cap B_{j}\right)} a_{i} \\
& =\sum_{i} \sum_{j} \int_{k_{j}\left(A_{i} \cap B_{j}\right)}\left(a_{i} \circ \phi_{i, j}\right) \cdot \operatorname{det}\left(J \phi_{i, j}\right) \\
& =\sum_{j} \sum_{i} \int_{k_{j}\left(A_{i} \cap B_{j}\right)} b_{j} \\
& =\sum_{j} \int_{k_{j}\left(B_{j}\right)} b_{j} .
\end{aligned}
$$

Note that $\operatorname{det}\left(J \phi_{i, j}\right)>0$ since we are using an orienting atlas.

Definition 10.5. The support of an $n$-form $\omega$ is the closed set

$$
\operatorname{supp}(\omega):=\overline{\left\{p \in M \mid \omega_{p} \neq 0\right\}} \subseteq M
$$

Remark 10.6. If $M$ itself is compact, then all $n$-forms have compact support.
Exercise 10.7. Convince yourself that an $n$-form $\omega \in \Omega^{n} M$ with compact support is integrable if and only if it is locally integrable, meaning that around any point, there is a chart $(U, h)$ such that $\omega \circ h^{-1}: h(U) \rightarrow \mathbb{R}$ is integrable on $h(U) \subseteq \mathbb{R}^{n}$. Therefore, if $M$ is compact, then all $n$-forms are integrable.

Exercise 10.8. Let $-M$ be $M$ with the opposite orientation and let $\omega \in \Omega^{n} M$ be integrable. Show that $\int_{-M} \omega=\int_{M}-\omega$.

Proposition 10.9. Let $M$ and $N$ be manifolds of dimension $n$. Consider $f: M \rightarrow$ $N$ an orientation-preserving diffeomorphism. If $\omega \in \Omega^{n} N$ is integrable, then the pullback $f^{*} \omega$ is integrable on $M$ and

$$
\int_{M} f^{*} \omega=\int_{N} \omega
$$

Proof. Exercise.
10.2. Manifolds with boundary. Now the goal is to prove Stokes' theorem on oriented manifolds:

$$
\int_{\partial M} \omega=\int_{M} d \omega
$$

for some $(n-1)$-form $\omega$ with compact support. But to do so, we need a good definition of $\partial M$, the boundary of $M$ and look at manifolds with boundary.

Definition 10.10. Let $\mathbb{R}_{-}^{n}$ denote the half space

$$
\mathbb{R}_{-}^{n}:=\left\{x \in \mathbb{R}^{n} \mid x_{1} \leq 0\right\}
$$

with the subspace topology inherited from $\mathbb{R}^{n}$. Define its boundary to be

$$
\partial \mathbb{R}_{-}^{n}:=\left\{x \in \mathbb{R}^{n} \mid x_{1}=0\right\}
$$

For $U \subseteq \mathbb{R}_{-}^{n}$, define the boundary of $U$ to be its intersection with $\partial \mathbb{R}_{-}^{n}$

$$
\partial U:=U \cap \partial \mathbb{R}_{-}^{n}=U \cap\left(\{0\} \times \mathbb{R}^{n-1}\right)
$$

Remark 10.11. First note that it is possible for $\partial U=\emptyset$. Also, the boundary $\partial U$ is different from the boundary of $U$ in the topological sense, i.e., from $U \backslash U^{\circ}$ (cf. Definition B.1).

Definition 10.12. Let $U \subseteq \mathbb{R}_{-}^{n}$ be open. A map $f: U \rightarrow \mathbb{R}^{k}$ is differentiable at $p \in U$ if there exist an open neighborhood $U_{p} \subseteq \mathbb{R}^{n}$ of $p$ and $\tilde{f}: U_{p} \rightarrow \mathbb{R}^{k}$ a differentiable map (in the normal sense) such that $\left.f\right|_{U \cap U_{p}}=\left.\tilde{f}\right|_{U \cap U_{p}}$. We call $\tilde{f}$ a local extension of $f$.

Definition 10.13. Let $U, V \subseteq \mathbb{R}_{-}^{n}$ be open. A differentiable map $f: U \rightarrow V$ is a diffeomorphism if it is bijective and has a differentiable inverse.


Figure 16. Differentiability at $\partial U$.

Lemma 10.14. Let $U, V \subseteq \mathbb{R}_{-}^{n}$ be open and let $f: U \rightarrow V$ be a diffeomorphism. Then $f(\partial U)=\partial V$. Hence, $\left.f\right|_{\partial U}: \partial U \xrightarrow{\cong} \partial V$ is a diffeomorphism between open sets of $\mathbb{R}^{n-1}$.

Proof. Let $p \in \partial U$, and let $\tilde{f}: U_{p} \rightarrow \mathbb{R}^{n}$ be a local extension of $f$. Suppose for contradiction that $f(p) \in V \backslash \partial V$.


Figure 17. What's wrong with this picture?
Since $f^{-1}$ is continuous, $\left(f^{-1}\right)^{-1}\left(U \cap U_{p}\right)=f\left(U \cap U_{p}\right) \subseteq V$ is open in $\mathbb{R}_{-}^{n}$. Let $V_{p} \subseteq f\left(U \cap U_{p}\right)$ be a neighborhood of $f(p)$ with the further restriction that $V_{p} \subseteq$ $V \backslash \partial V$. Such $V_{p}$ exists since $f(p) \notin \partial V$ by assumption. Define $\tilde{V}=f\left(U \cap U_{p}\right)$, and consider the restriction $\left.f^{-1}\right|_{\tilde{V}}: \tilde{V} \rightarrow U \cap U_{p} \subset \mathbb{R}^{n}$. Our goal is to show that $U \cap U_{p}$, which is open in $\mathbb{R}_{-}^{n}$, is actually open in $\mathbb{R}^{n}$. That will yield a contradiction, since $U \cap U_{p}$ contains no open ball about $p$.

Since $\left.f \circ f^{-1}\right|_{\tilde{V}}=\mathrm{id}_{\tilde{V}}$, for any $x \in \tilde{V}$, we have

$$
J\left(f \circ f^{-1}\right)_{x}=J f_{f^{-1}(x)}\left(J f^{-1}\right)_{x}=I_{n}
$$

where $I_{n}$ is the identity matrix. In particular, $J f_{x}^{-1}$ is invertible for all $x$. By the inverse function theorem, $\left.f^{-1}\right|_{\tilde{V}}$ is a local diffeomorphism (cf. Definition A.6) and, hence, an open map. Therefore, $f^{-1}(\tilde{V})=U \cap U_{p} \subseteq U$ is a neighborhood of $p$ that is open in $\mathbb{R}^{n}$. This contradicts the assumption that $p \in \partial U$. Thus, $f(\partial U) \subseteq \partial V$. The other inclusion follows from applying the same argument to $f^{-1}$.

Lemma 10.15. Let $U, V \subseteq \mathbb{R}_{-}^{n}$ be open and let $f: U \rightarrow V$ be a diffeomorphism. Let $p \in \partial U$. Then the well-defined differential $d f_{p}: \mathbb{R}^{n} \xrightarrow{\cong} \mathbb{R}^{n}$ maps $\partial \mathbb{R}_{-}^{n}$ to $\partial \mathbb{R}_{-}^{n}$, $\mathbb{R}_{-}^{n}$ to $\mathbb{R}_{-}^{n}$, and $\mathbb{R}_{+}^{n}:=\left\{x \in \mathbb{R}^{n} \mid x_{1} \geq 0\right\}$ to $\mathbb{R}_{+}^{n}$.

Proof. Since we can extend $f$ to some neighborhood of $p$ in $\mathbb{R}^{n}$, the differential $d f_{p}$ is well-defined. The goal is to show that the Jacobian $J f_{p}$ is of the form

$$
J f_{p}=\left[\begin{array}{c|ccc}
\frac{\partial f_{1}}{\partial x_{1}}(p) & 0 & \cdots & 0 \\
\hline \frac{\partial f_{2}}{\partial x_{1}}(p) & & & \\
\vdots & & \left(\left.J f\right|_{\partial \mathbb{R}_{-}^{n}}\right)_{p} & \\
\frac{\partial f_{n}}{\partial x_{1}}(p) & & &
\end{array}\right]
$$

with $\frac{\partial f_{1}}{\partial x_{1}}(p)>0$. Then the result follows. First recall that

$$
\frac{\partial f_{i}}{\partial x_{j}}(p)=\lim _{t \rightarrow 0} \frac{f_{i}\left(p+t e_{j}\right)-f_{i}(p)}{t}
$$

By Lemma 10.14 , for $j \geq 2$, we have

$$
\frac{\partial f_{1}}{\partial x_{j}}(p)=\lim _{t \rightarrow 0} \frac{f_{1}\left(p+t e_{j}\right)-f_{1}(p)}{t}=\lim _{t \rightarrow 0} \frac{0-0}{t}=0
$$

So the first row of $J f_{p}$ except for the first entry is zero. At the same time, since $f_{1}\left(p+t e_{1}\right) \leq 0$,

$$
\frac{\partial f_{1}}{\partial x_{1}}(p)=\lim _{t \rightarrow 0^{-}} \frac{f_{1}\left(p+t e_{1}\right)-f_{1}(p)}{t}=\lim _{t \rightarrow 0^{-}} \frac{f_{1}\left(p+t e_{1}\right)-0}{t} \geq 0
$$

This forces $\frac{\partial f_{1}}{\partial x_{1}}(p)>0$ or otherwise $\operatorname{det}\left(J f_{p}\right)=0$.
Definition 10.16. An $n$-dimensional manifold with boundary is a second-countable, Hausdorff topological space that is locally homeomorphic to open subsets of $\mathbb{R}_{-}^{n}$ with differentiable transition functions. A point $p \in M$ is in the boundary of $M$ if there is some (hence every) chart $(U, h)$ at $p$ such that $h(p) \in \partial h(U) \subseteq \mathbb{R}_{-}^{n}$. The collection of all such points is denoted $\partial M$.

Remark 10.17. Again note the difference between $\partial M$ and the topological boundary of $M$.

Remark 10.18. If $M$ is an $n$-dimensional manifold with boundary, then the restrictions

$$
h_{U \cap \partial M}: U \cap \partial M \rightarrow \partial h(U) \subseteq \partial \mathbb{R}_{-}^{n}
$$

of charts $(U, h)$ on $M$ turn $\partial M$ into an $(n-1)$-dimensional manifold (without boundary).

Exercise 10.19. Let $M$ and $N$ be $n$-dimensional manifolds with boundary and let $f: M \rightarrow N$ be a diffeomorphism between manifolds with boundary. Show that $f(\partial M)=\partial N$ and that the restriction $\left.f\right|_{\partial M}: \partial M \rightarrow \partial N$ is a diffeomorphism.

Example 10.20. Let $M$ be an $n$-dimensional manifold. Then it is an $n$-dimensional manifold with boundary and $\partial M=\emptyset$. To see this, note that for a chart $(U, h)$, we can create a finite collection of smaller charts $\left\{\left(U_{i}, h_{i}\right)\right\}_{i=1}^{m}$ such that $U=\bigcup_{i=1}^{m} U_{i}$
and $h_{i}\left(U_{i}\right) \subseteq \mathbb{R}_{-}^{n} \backslash \partial \mathbb{R}_{-}^{n}$. Conversely, if $M$ is a manifold with boundary, then it is a manifold if $\partial M=\emptyset$.

Example 10.21. The closed ball $D^{n}:=\left\{x \in \mathbb{R}^{n}| | x \mid \leq 1\right\}$ is a manifold with boundary $S^{n-1}$. The cylinder $[0,1] \times S^{1}$ is a manifold with boundary equal to the dijoint union of two circles.

Definition 10.22. Let $M$ be a manifold with boundary and let $p \in \partial M$. Define the tangent space to $M$ at $p$ to be

$$
T_{p} M:=T_{p}^{\mathrm{alg}} M \stackrel{\cong}{\cong} T_{p}^{\mathrm{phy}} M
$$

as described in section 4 . Let $(U, h)$ be a chart at $p$ and define

$$
T_{p}^{+} M:=\left(d h_{p}\right)^{-1}\left(\mathbb{R}_{+}^{n}\right), \quad T_{p}^{-} M:=\left(d h_{p}\right)^{-1}\left(\mathbb{R}_{-}^{n}\right)
$$

Note that this definition does not depend on the choice of charts by Lemma 10.15.
Remark 10.23. The inclusion $\partial M \hookrightarrow M$ gives an inclusion of tangent spaces at $p \in \partial M$, and in fact we have

$$
T_{p} \partial M=\overline{T_{p}^{+} M} \cap \overline{T_{p}^{-M}}
$$

where the bar indicates topological closure in $\mathbb{R}^{n}$.
Definition 10.24. Let $M$ be a manifold with boundary and let $p \in \partial M$. We call elements in $T_{p}^{-} M \backslash T_{p} \partial M$ the inward-pointing tangent vectors, and elements in $T_{p}^{+} M \backslash T_{p} \partial M$ the outward-pointing tangent vectors. Note that this definition can be given without embedding $M$ into some $\mathbb{R}^{N}$.

The definition of an oriented manifold with boundary is the same as for ordinary manifolds. The boundary then is then orientable, but we would like to fix a convention for its orientation.

Definition 10.25. Let $M$ be an $n$-dimensional oriented manifold with boundary and let $p \in \partial M$. We define the natural orientation on $\partial M$ as follows: an ordered basis $\left\langle w_{1}, \ldots, w_{n-1}\right\rangle$ for $T_{p} \partial M$ is positively oriented if for any outward-pointing tangent vector $v \in T_{p} M$, the ordered basis $\left\langle v, w_{1}, \ldots, w_{n-1}\right\rangle$ for $T_{p} M$ is positively oriented in $T_{p} M$.

Example 10.26. Let $D^{3}$ denote the solid unit ball in $\mathbb{R}^{3}$ with its orientation induced by $\mathbb{R}^{3}$. Its boundary is $\partial D^{3}=S^{2}$, and the natural orientation on $S^{2}$ is given by $\left\langle w_{1}, w_{2}\right\rangle$ as shown in Figure 18.

### 10.3. Stokes' theorem on manifolds.

Theorem 10.27. Let $M$ be an n-dimensional oriented manifold with boundary and let $\omega \in \Omega^{n-1} M$ be an $(n-1)$-form with compact support. Let $\iota$ : $\partial M \hookrightarrow M$ denote the inclusion and define $\int_{\partial M} \omega:=\int_{\partial M} \iota^{*} \omega$. Then $d \omega$ is integrable and

$$
\int_{M} d \omega=\int_{\partial M} \omega
$$



Figure 18. The natural orientation on $S^{2}=\partial D^{3}$.

Proof. We carry out the proof in three steps of increasing generality.
Case 1: $M=\mathbb{R}_{-}^{n}$. Write $\omega$ as

$$
\omega=\sum_{i=1}^{n} a_{i} d x_{1} \wedge \cdots \wedge \hat{d x}_{i} \wedge \cdots \wedge d x_{n}
$$

Then

$$
\begin{aligned}
d \omega & =\sum_{i=1}^{n} \sum_{j=1}^{n} \frac{\partial a_{i}}{\partial x_{j}} d x_{j} \wedge d x_{1} \wedge \cdots \wedge d \hat{x}_{i} \wedge \cdots \wedge d x_{n} \\
& =\sum_{i=1}^{n}(-1)^{i-1} \frac{\partial a_{i}}{\partial x_{i}} d x_{1} \wedge \cdots \wedge d x_{i} \wedge \cdots \wedge d x_{n}
\end{aligned}
$$

This gives us

$$
\int_{M} d \omega=\sum_{i=1}^{n} \int_{\mathbb{R}_{-}^{n}}(-1)^{i-1} \frac{\partial a_{i}}{\partial x_{i}}
$$

Use Fubini's theorem to first integrate $\int_{\mathbb{R}_{-}^{n}}(-1)^{i-1} \frac{\partial a_{i}}{\partial x_{i}}$ with respect to $x_{i}$. For $i \neq 1$, by definition we have

$$
\begin{aligned}
\int_{x_{i}=-\infty}^{\infty} \frac{\partial a_{i}}{\partial x_{i}}= & \lim _{t \rightarrow \infty} \int_{0}^{t} \frac{\partial a_{i}}{\partial x_{i}}+\lim _{t \rightarrow-\infty} \int_{t}^{0} \frac{\partial a_{i}}{\partial x_{i}} \\
= & \lim _{t \rightarrow \infty}\left(a_{i}\left(x_{1}, \ldots, x_{i-1}, t, x_{i+1}, \ldots, x_{n}\right)-a_{i}\left(x_{1}, \ldots, x_{i-1}, 0, x_{i+1}, \ldots, x_{n}\right)\right) \\
& +\lim _{t \rightarrow-\infty}\left(a_{i}\left(x_{1}, \ldots, x_{i-1}, 0, x_{i+1}, \ldots, x_{n}\right)-a_{i}\left(x_{1}, \ldots, x_{i-1}, t, x_{i+1}, \ldots, x_{n}\right)\right) \\
= & \left(0-a_{i}\left(x_{1}, \ldots, x_{i-1}, 0, x_{i+1}, \ldots, x_{n}\right)\right)+\left(a_{i}\left(x_{1}, \ldots, x_{i-1}, 0, x_{i+1}, \ldots, x_{n}\right)-0\right) \\
= & 0,
\end{aligned}
$$

where $\lim _{t \rightarrow \infty} a_{i}\left(x_{1}, \ldots, x_{i-1}, t, x_{i+1}, \ldots, x_{n}\right)=0$ since $\omega$ has compact support. For $i=1$, since $M=\mathbb{R}_{-}^{n}$, we have instead

$$
\int_{x_{1}=-\infty}^{0} \frac{\partial a_{1}}{\partial x_{1}}=\lim _{t \rightarrow-\infty} \int_{x_{1}=t}^{0} \frac{\partial a_{1}}{\partial x_{1}}=a_{1}\left(0, x_{2}, \ldots, x_{n}\right)
$$

Therefore,

$$
\int_{M} d \omega=\sum_{i=1}^{n} \int_{\mathbb{R}_{-}^{n}}(-1)^{i-1} \frac{\partial a_{i}}{\partial x_{i}}=\int_{\mathbb{R}_{-}^{n}} \frac{\partial a_{1}}{\partial x_{1}}=\int_{\mathbb{R}^{n-1}} a_{1}\left(0, x_{2}, \ldots, x_{n}\right)
$$

On the other hand, note that $\iota: \partial M \rightarrow M$ is given by

$$
\begin{aligned}
\iota: \partial M & \longrightarrow M \\
\left(x_{2}, \ldots, x_{n}\right) & \longmapsto\left(0, x_{2}, \ldots, x_{n}\right)
\end{aligned}
$$

Since $\iota_{1} \equiv 0$, we have

$$
\begin{aligned}
\int_{\partial M} \omega=\int_{\partial M} \iota^{*} \omega & =\sum_{i=1}^{n} \int_{\mathbb{R}^{n-1}} \iota^{*}\left(a_{i} d x_{1} \wedge \cdots \wedge \dot{d x_{i}} \wedge \cdots \wedge d x_{n}\right) \\
& =\sum_{i=1}^{n} \int_{\mathbb{R}^{n-1}} a_{i}\left(0, x_{2}, \ldots, x_{n}\right) d \iota_{1} \wedge \cdots \wedge \hat{\iota_{i}} \wedge \cdots \wedge d \iota_{n} \\
& =\int_{\mathbb{R}^{n-1}} a_{1}\left(0, x_{2}, \ldots, x_{n}\right) d x_{2} \wedge \cdots \wedge d x_{n} \\
& =\int_{M} d \omega
\end{aligned}
$$

Case 2: Now suppose that $M$ is an arbitrary oriented manifold with boundary and that $\operatorname{supp}(\omega) \subseteq U$ for some orienting chart $(U, h)$. We can use the chart to reduce to the previous case: Using the chart and the fact that $\operatorname{supp}(\omega) \subseteq U$, we may assume $M=U \subseteq \mathbb{R}_{-}^{n}$. Then extend $\omega$ to a form $\widetilde{\omega}$ on all of $\mathbb{R}_{-}^{n}$ by letting $\left.\widetilde{\omega}\right|_{U}=\omega$, and $\left.\widetilde{\omega}\right|_{U^{c}} \equiv 0$. This will not cause any problem since $\operatorname{supp}(\omega) \subseteq U$ is compact.

Case 3: Finally, suppose that $\omega \in \Omega^{n-1} M$ is any compactly supported form. For this case, as we describe below, we can break $\omega$ into a finite sum $\omega=\omega_{1}+\cdots+\omega_{r}$ of $(n-1)$-forms, each of which has compact support that is contained in a chart. Then the previous case applies and we are home.

Around each $p \in \operatorname{supp}(\omega)$, choose an orientation-preserving chart $\left(U_{p}, h_{p}\right)$ and a smooth and compactly-supported function $\lambda_{p}: M \rightarrow[0,1]$ such that $\lambda_{p}(p)>0$ and $\operatorname{supp}\left(\lambda_{p}\right) \subseteq U_{p}$ (i.e., a bump function). Then $\left\{\lambda_{p}^{-1}((0,1])\right\}_{p \in \operatorname{supp}(\omega)}$ is an open cover of $\operatorname{supp}(\omega) . \operatorname{Since} \operatorname{supp}(\omega)$ is compact, there are $p_{1}, \ldots, p_{r} \in \operatorname{supp}(\omega)$ such that

$$
\operatorname{supp}(\omega) \subseteq \bigcup_{i=1}^{r} \lambda_{p}^{-1}((0,1]) \stackrel{\text { call }}{=} X
$$

Define $r$ differentiable functions $\tau_{1}, \ldots, \tau_{r}$ by

$$
\begin{aligned}
\tau_{i}: X & \longrightarrow[0,1] \\
x & \longmapsto \frac{\lambda_{p_{i}}(x)}{\sum_{i=1}^{r} \lambda_{p_{i}}(x)} .
\end{aligned}
$$

Then $\sum_{i=1}^{r} \tau_{i}(x)=1$ for all $x \in X$, and we call $\tau_{1}, \ldots, \tau_{r}$ a partition of unity. To find the corresponding partition of $\omega$, define $\omega_{i} \in \Omega^{n-1} M$ by

$$
\omega_{i}(p)=\left\{\begin{array}{cc}
\tau_{i}(p) \omega(p) & \text { if } p \in X \\
0 & \text { otherwise }
\end{array}\right.
$$

One can see that $\operatorname{supp}\left(\omega_{i}\right) \subseteq U_{p}$ is compact. Also, $\omega_{i}$ is differentiable on $M$ and $\omega=\omega_{1}+\cdots+\omega_{r}$. Now by the previous case, we have

$$
\int_{M} d \omega_{i}=\int_{\partial M} \omega_{i},
$$

and finally, by linearity,

$$
\int_{M} d \omega=\int_{\partial M} \omega
$$

## 11. de Rham Cohomology

Recall that there is a contravariant functor from the category of smooth manifolds to the category of cochain complexes. That is, to each manifold $M$, we associate the de Rham complex

$$
0 \rightarrow \Omega^{0} M \xrightarrow{d} \Omega^{1} M \xrightarrow{d} \Omega^{2} M \xrightarrow{d} \cdots
$$

where at degree $k$ we have $\Omega^{k} M$, the $k$-forms on $M$, and $d: \Omega^{k} M \rightarrow \Omega^{k+1} M$ is exterior differentiation. Furthermore, a smooth map $f: M \rightarrow N$ between manifolds induces a map between cochains


The cohomology groups of the de Rham complex of a manifold, defined below, are important invariants of the manifold. Thus, if two manifolds have different cohomology groups, then they are not diffeomorphic. Furthermore, using Stokes' theorem, it can be shown that de Rham cohomology is dual to singular homology on the manifold, where the latter can be computed through a triangulation of the manifold and detects topological features of the manifold, for instance, the number of $k$-dimensional holes (see Appendix D).
11.1. Definition and first properties. Recall that a form in the kernel of $d$ is called closed and a form in the image of $d$ is called exact. In the context of de Rham cohomology, closed forms are called cocycles and exact forms are called coboundaries. ${ }^{6}$

Since $d^{2}=0$, every exact form is closed, or said another way, every coboundary is a cocycle. Thus, we can make the following definition:

Definition 11.1. The $k$-th cohomology group of the de Rham complex is the quotient

$$
H^{k} M:=\operatorname{ker}\left(\Omega^{k} M \xrightarrow{d} \Omega^{k+1} M\right) / \operatorname{im}\left(\Omega^{k-1} M \xrightarrow{d} \Omega^{k} M\right)
$$

If $\omega \in \Omega^{k} M$ is a cocycle, we denote the cohomology class of $\omega$ by

$$
[\omega]:=\omega+d\left(\Omega^{k-1} M\right)
$$

We say that cocycles $\omega$ and $\eta$ are cohomologous if $[\omega]=[\eta]$, i.e., if $\omega-\eta=d \alpha$ for some $\alpha \in \Omega^{k-1} M$.

The cohomology groups measure the extent to which the de Rham sequence is not exact: i.e., $H^{k} M=0$ if and only if $\operatorname{im} d_{k-1}=\operatorname{ker} d_{k}$. If its dimension as an $\mathbb{R}$-vector space is large, then the sequence is far from being exact in degree $k$. (See Appendix subsection D. 1 for the basics on exact sequences.)

[^5]Example 11.2. Let $M=\mathbb{R}$. Since $M$ one-dimensional, its de Rham complex is

$$
0 \rightarrow \Omega^{0} \mathbb{R} \xrightarrow{d} \Omega^{1} \mathbb{R} \rightarrow 0
$$

Note that

$$
\Omega^{0} \mathbb{R}=\{f: \mathbb{R} \rightarrow \mathbb{R} \mid f \text { is smooth }\}
$$

and that

$$
\Omega^{1} \mathbb{R}=\{f d x \mid f: \mathbb{R} \rightarrow \mathbb{R} \text { is smooth }\}
$$

Also, for a smooth function $f: \mathbb{R} \rightarrow \mathbb{R}, d f=0$ means that $f$ is a constant function. Thus, we have $\operatorname{ker}\left(\Omega^{0} \mathbb{R} \xrightarrow{d} \Omega^{1} \mathbb{R}\right) \cong \mathbb{R}$. At the same time, for any $f \in \Omega^{1} \mathbb{R}$, we can define $g(x)=\int_{0}^{x} f(t) d t$. Then $g$ is smooth and $d g=f$. Therefore, $\operatorname{im}\left(\Omega^{0} \mathbb{R} \xrightarrow{d}\right.$ $\left.\Omega^{1} \mathbb{R}\right)=\Omega^{1} \mathbb{R}$. In this way we can compute the cohomology groups:

$$
\begin{aligned}
& H^{0} M=\operatorname{ker}\left(\Omega^{0} \mathbb{R} \xrightarrow[\rightarrow]{d} \Omega^{1} \mathbb{R}\right) / \operatorname{im}\left(0 \rightarrow \Omega^{0} \mathbb{R}\right) \cong \mathbb{R} / 0=\mathbb{R} \\
& H^{1} M=\operatorname{im}\left(\Omega^{0} \mathbb{R} \xrightarrow{d} \Omega^{1} \mathbb{R}\right) / \Omega^{1} \mathbb{R} \cong \mathbb{R} / \mathbb{R}=0
\end{aligned}
$$

The cohomology groups are not only groups under addition, they are also $\mathbb{R}$ vector spaces. In fact, there is even more structure. Define the cohomology ring of $M$ to be

$$
H^{\bullet} M:=\bigoplus_{k \geq 0} H^{k} M
$$

where the product is induced by the wedge product on $\Omega^{\bullet} M$. This product on $H^{\bullet} M$ is defined to be

$$
\begin{aligned}
\wedge: H^{r} M \times H^{s} M & \longrightarrow H^{r+s} M \\
([\omega],[\eta]) & \longmapsto[\omega \wedge \eta]
\end{aligned}
$$

Theorem 11.3. The product on $H^{\bullet} M$ is well-defined, making $H^{\bullet} M$ into a graded, anti-commutative $\mathbb{R}$-algebra. ${ }^{7}$

Proof. Conceptualize the cohomology groups as cocycles modulo coboundaries. We first need to check that the wedge product of two cocycles is a cocycle. Suppose $\omega$ and $\eta$ are cocycles, then by the product rule for exterior differentiation, $\omega \wedge \eta$ is also a cocycle:

$$
d(\omega \wedge \eta)=d \omega \wedge \eta \pm \omega \wedge d \eta=0+0=0
$$

Next, we show that the product does not depend on the choice of representative. Let $\mu \in \Omega^{r-1} M$ and let $\nu \in \Omega^{s-1} M$. Then

$$
\begin{aligned}
d((\omega+d \mu) \wedge(\eta+d \nu)) & =d(\omega+d \mu) \wedge(\eta+d \nu)+(-1)^{r}(\omega+d \mu) \wedge d(\eta+d \nu) \\
& =d \omega \wedge(\eta+d \nu)+(-1)^{r}(\omega+d \mu) \wedge d \eta \\
& =d \omega \wedge \eta+d \omega \wedge d \nu+(-1)^{r}(\omega \wedge d \eta+d \mu \wedge d \eta) \\
& =\left(d \omega \wedge \eta+(-1)^{r} \omega \wedge d \eta\right)+\left(d \omega \wedge d \nu+(-1)^{r} d \mu \wedge d \eta\right)
\end{aligned}
$$

[^6]We are trying to show that the difference between this form and $d(\omega \wedge \eta)$ is a coboundary, i.e., in the image of $d$. By the product rule $d(\omega \wedge \eta)=d \omega \wedge \eta+$ $(-1)^{r} \omega \wedge d \eta$. The difference is

$$
\left.d \omega \wedge d \nu+(-1)^{r} d \mu \wedge d \eta\right)=d(\omega \wedge d \nu+d \mu \wedge \eta)
$$

Remark 11.4. Recall that a smooth map $f: M \rightarrow N$ between manifolds induces a map between $k$-forms $f^{*}: \Omega^{k} N \rightarrow \Omega^{k} M$. In fact, $f$ also induces a group homomorphism between the $k$-th homology groups $f^{*, k}: H^{k} N \rightarrow H^{k} M$ that also commutes with the exterior differentiation $d$. In this way, $H^{\bullet}$ is a contravariant functor from the category of smooth manifolds to the category of graded, anti-commutative algebras over $\mathbb{R}$. We use $f^{*}: H^{\bullet} N \rightarrow H^{\bullet} M$ to denote the collection of induced maps $f^{*, k}$. If $f$ is constant, then $f^{*, k} \equiv 0$ for all $k>0$.

Proposition 11.5. Let $M$ be a smooth manifold of dimension $n$. Then $H^{k} M=0$ for all $k>n$.

Proof. Note that $\Omega^{k} M=0$ for $k>n$.
Proposition 11.6. Suppose that $M$ has c connected components. Then $H^{0} M \cong$ $\mathbb{R}^{c}$. In particular, $H^{0} M \cong \mathbb{R}$ if $M$ is connected.

Proof. If $f \in \operatorname{ker}\left(\Omega^{0} M \xrightarrow{d} \Omega^{1} M\right)$, then locally we have $d f=\sum_{i=1}^{n} \frac{\partial f}{\partial x_{i}} d x_{i}=0$, which means that $\frac{\partial f}{\partial x_{i}}=0$ for all $i$. Therefore, $f$ is locally constant. Since $M$ is locally path connected, its connected components agree with its path-connected components. Then $f$ is constant on each connected component: take a path $\gamma$ in one of these components and we have $(f \circ \gamma)^{\prime} \equiv 0$. (In fact, if $g: X \rightarrow Y$ is a locally constant continuous mapping of topological spaces, and points are closed in $Y$, e.g., $Y$ is Hausdorff, then $g$ is constant on connected components of $X$.)

Exercise 11.7. Let $f: M \rightarrow N$ be a smooth map between manifolds. Suppose that $M$ and $N$ are both connected. Show that $f^{*, 0}: H^{0} N \rightarrow H^{0} M$ is the identity map on $\mathbb{R}$.

Proposition 11.8. Let $M$ be an orientable manifold of dimension n. Suppose that $M$ closed, i.e., compact with $\partial M=0$. Then $H^{n} M \neq 0$.

Proof. It suffices to show that there is some $n$-form on $M$ that is not exact. Orient $M$ and choose $\omega \in \Omega^{n} M$ such that $\int_{M} \omega \neq 0$. If there were $\eta \in \Omega^{n-1} M$ such that $\omega=d \eta$, then by the Stokes' theorem, $\int_{M} \omega=\int_{M} d \eta=\int_{\partial M} \eta=0$ since $\partial M=\emptyset$ by assumption.

### 11.2. Homotopy invariance of de Rham cohomology.

Definition 11.9. Let $f, g: M \rightarrow N$ be smooth maps between manifolds. We say that $f$ and $g$ are (smoothly) homotopic, denoted $f \sim g$, if there is a smooth map

$$
h:[0,1] \times M \longrightarrow N
$$

such that $h(0, x)=f(x)$ and $h(1, x)=g(x)$ for all $x \in M$ (see Figure 19). In that case, we call $h$ a homotopy from $f$ to $g$ and write $f \underset{\sim}{\sim} g$.

We say that $f$ is null-homotopic if $f$ is homotopic to some constant function. We say that $M$ is contractible if $\mathrm{id}_{M}$ is null homotopic.


Figure 19. Homotopy between mapping $f, g: \mathbb{R} \rightarrow \mathbb{R}^{2}$.

Example 11.10. This example shows that $\mathbb{R}^{n}$ is contractible. Here is a homotopy from the identity to the constant zero-mapping:

$$
\begin{aligned}
h:[0,1] \times \mathbb{R}^{n} & \rightarrow \mathbb{R}^{n} \\
(t, x) & \mapsto(1-t) x
\end{aligned}
$$

At time zero, we have $h_{0}(x):=h(0, x)=x=\operatorname{id}_{\mathbb{R}^{n}}(x)$, and at time 1 , we have $h_{1}(x):=h(1, x)=0$.

Exercise 11.11. Show that $\sim$ is an equivalence relation on the set of mappings $M \rightarrow N$.

Theorem 11.12. Let $f, g: M \rightarrow N$ be smooth maps that are homotopic. Then they induce the same map on the cohomology groups, i.e.,

$$
f^{*}=g^{*}: H^{\bullet} N \rightarrow H^{\bullet} M
$$

Proof. Let $h:[0,1] \times M \rightarrow N$ denote the homotopy between $f$ and $g$ such that $h(0, x)=f(x)$ and $h(1, x)=g(x)$ for all $x \in M$. The goal is to use $h$ to construct a collection of maps $s: \Omega^{k} N \rightarrow \Omega^{k-1} M$ (shown below, though the diagram does not

necessarily commute) such that for any $\omega \in \Omega^{k} N$,

$$
\begin{equation*}
g^{*} \omega-f^{*} \omega=(s d+d s) \omega \tag{5}
\end{equation*}
$$

Then if $\omega$ is closed, i.e., $d \omega=0$, we have

$$
g^{*} \omega-f^{*} \omega=(d s) \omega=d(s \omega)
$$

which implies that $\left[f^{*} \omega\right]=\left[g^{*} \omega\right]$ and, hence, $f^{*}=g^{*}$. Thus, it suffices to show the existence the mappings $s$.

To reduce the problem further, for each $t \in[0,1]$, define

$$
\begin{aligned}
i_{t}: M & \longrightarrow[0,1] \times M \\
x & \longmapsto(t, x) .
\end{aligned}
$$

We have that $\left(h \circ i_{0}\right)(x)=h(0, x)=f(x)$ for all $x \in M$ and, similarly, $h \circ i_{1}=g$. For each $k$, we want to define the "prism operator" $P: \Omega^{k}([0,1] \times M) \rightarrow \Omega^{k-1} M$ such that

$$
\begin{equation*}
i_{1}^{*}-i_{0}^{*}=d P+P d \tag{6}
\end{equation*}
$$

We can let $s:=P \circ h^{*}$ :

$$
\Omega^{k} N \xrightarrow{h^{*}} \Omega^{k}([0,1] \times M) \xrightarrow{P} \Omega^{k-1} M,
$$

and using the fact that the exterior derivative and the pullback operators commute, we see that Equation 5 is satisfied:
$g^{*}-f^{*}=i_{1}^{*} h^{*}-i_{0}^{*} h^{*}=\left(i_{1}^{*}-i_{0}^{*}\right) h^{*}=(d P+P d) h^{*}=d\left(P h^{*}\right)+\left(P h^{*}\right) d=d s+s d$,
as required.
So the problem becomes finding the prism operator $P$. Define

$$
\begin{aligned}
P: \Omega^{k}([0,1] \times M) & \longrightarrow \Omega^{k-1} M \\
\omega & \longmapsto \int_{t=0}^{1} \omega\left(\frac{\partial}{\partial t},-\right)
\end{aligned}
$$

where $t$ is the coordinate on $[0,1]$. We need to clarify the meaning of $\omega\left(\frac{\partial}{\partial t},-\right)$. We are thinking of $\omega$ as a multilinear alternating form acting on tuples of tangent vectors. In local coordinates, $\partial / \partial t$ and $d t$ are dual, i.e., $d t(\partial / \partial t)=1$. So in local coordinates, if $d t$ occurs in $\omega$, by flipping signs if necessary, we make $d t$ the leading term in the wedge product. Then $\omega\left(\frac{\partial}{\partial t},-\right)$ is the form obtained by setting $d t$ equal to 1 . If $d t$ does not occur, then $\omega\left(\frac{\partial}{\partial t},-\right)=0 .{ }^{8}$ We pause here for an example of the prism operator.

Example 11.13. Let $M=\mathbb{R}^{3}$, and consider the $\omega \in \Omega^{3}\left([0,1] \times \mathbb{R}^{3}\right)$ given by

$$
\omega=\left(3 t^{2}+2 x t+x^{2} y\right) d t \wedge d x \wedge d y+\left(t x+t^{2} y\right) d x \wedge d y \wedge d z
$$

Then

$$
\begin{aligned}
P \omega & =P\left(\left(3 t^{2}+2 x t+x^{2} y\right) d t \wedge d x \wedge d y\right)+P\left(\left(t x+t^{2} y\right) d x \wedge d y \wedge d z\right) \\
& =P\left(\left(3 t^{2}+2 x t+x^{2} y\right) d t \wedge d x \wedge d y\right)+0 \\
& =\left(\int_{t=0}^{1}\left(3 t^{2}+2 x t+x^{2} y\right)\right) d x \wedge d y \\
& =\left(1+x+x^{2} y\right) d x \wedge d y \in \Omega^{2} M
\end{aligned}
$$

[^7]A 0 appears as a summand in the second step of this calculation since $d t$ does not appear in $\left(t x+t^{2} y\right) d x \wedge d y \wedge d z$.

We now resume the proof. The only thing that remains is to show that $P$ satisfies Equation 6. This is a local question, so we check it in local coordinates $t, x_{1}, \ldots, x_{m}$, where $m=\operatorname{dim} M$, and exterior differentiation, the pullback, and integrals are linear, we may assume $\omega$ has a single term. That term may or may not involve the differential $d t$. We consider each of those cases separately:

Case 1: Suppose that $\omega=\ell(t, x) d t \wedge d x_{\mu}$.
Consider the left-hand side of Equation 6 first. Pulling back by $i_{0}(x)=(0, x)$, we have

$$
i_{0}^{*}(\omega)=\ell(0, x) d 0 \wedge d x=0
$$

Similarly, $i_{1}^{*} \omega=0$. Therefore $i_{1}^{*} \omega-i_{0}^{*} \omega=0$. Therefore, to verify Equation 6, we must check that $d P \omega=-P d \omega$. Using the fact that $d t \wedge d t=0$, compute

$$
\begin{aligned}
d(P \omega) & =d\left(\left(\int_{t=0}^{1} \ell(t, x)\right) d x_{\mu}\right) \\
& =\left(\sum_{i=1}^{n} \frac{\partial}{\partial x_{i}}\left(\int_{0}^{1} \ell(x, t) d t\right) d x_{i}\right) \wedge d x_{\mu} \\
& =\left(\sum_{i=1}^{n}\left(\int_{0}^{1} \frac{\partial \ell}{\partial x_{i}} d t\right) d x_{i}\right) \wedge d x_{\mu} \\
& =\sum_{i=1}^{n}\left(\int_{0}^{1} \frac{\partial \ell}{\partial x_{i}} d t\right) d x_{i} \wedge d x_{\mu}
\end{aligned}
$$

and

$$
\begin{aligned}
P(d \omega) & =P\left(\sum_{i=1}^{n} \frac{\partial \ell}{\partial x_{i}} d x_{i} \wedge d t \wedge d x_{\mu}\right) \\
& =P\left(-\sum_{i=1}^{n} \frac{\partial \ell}{\partial x_{i}} d t \wedge d x_{i} \wedge d x_{\mu}\right) \\
& =-\sum_{i=1}^{n}\left(\int_{t=0}^{1} \frac{\partial \ell}{\partial x_{i}}\right) d x_{i} \wedge d x_{\mu}
\end{aligned}
$$

as required.
Case 2: Suppose that $\omega=\ell(x, t) d x_{\mu}$.
Now $P \omega=0$ and, hence, $d P \omega=0$. We need to show that $i_{1}^{*} \omega-i_{0}^{*} \omega=P d \omega$ :

$$
\begin{aligned}
P d \omega & =P\left(\frac{\partial \ell}{\partial t} d t \wedge d x_{\mu}+\sum_{i=1}^{n} \frac{\partial \ell}{\partial x_{i}} d x_{i} \wedge d x_{\mu}\right) \\
& =\left(\int_{t=0}^{1} \frac{\partial \ell}{\partial t}\right) d x_{\mu}+0 \\
& =(\ell(x, 1)-\ell(x, 0)) d x_{\mu}
\end{aligned}
$$

$$
=i_{1}^{*} \omega-i_{0}^{*} \omega
$$

Corollary 11.14. If $f: M \rightarrow N$ is null-homotopic, then $f^{*, k} \equiv 0$ for all $k>0$.
Corollary 11.15. If $M$ is contractible, then $H^{k} M=0$ for all $k>0$.
Corollary 11.16 (The Poincaré Lemma). If $U \subseteq \mathbb{R}^{n}$ is open and star-shaped (meaning that there is $x \in U$ such that for all $y \in U$, the line segment connecting $x$ and $y$ lies entirely in $U$ ), then $H^{k} U=0$ for all $k>0$.

Exercise 11.17. Prove the above three corollaries.
Definition 11.18. A mapping $f: M \rightarrow N$ defines a homotopy equivalence if there is a smooth mapping $g: N \rightarrow M$ such that $f \circ g \sim \operatorname{id}_{N}$ and $g \circ f \sim \operatorname{id}_{M}$. If so, we say that $M$ and $N$ are homotopy equivalent.

Remark 11.19. In particular, diffeomorphisms defines homotopy equivalences.
Theorem 11.20. If $M$ and $N$ are homotopy equivalent, then $H^{k} M \cong H^{k} N$ for all $k$.

Proof. This follows immediately from Theorem 11.12.
Remark 11.21 (Continuity versus smoothness.). Let $M$ and $N$ be smooth manifolds. We have defined homotopies for smooth functions $M \rightarrow N$, and further our homotopies, themselves are smooth functions. It turns out that all of our results hold if we allow our mappings to just be continuous rather than smooth. In detail: we say continuous mappings $f, g: M \rightarrow N$ (where each of $M$ and $N$ are still smooth manifolds) are (continuously) homotopic if there exists a continous mapping $h:[0,1] \times M \rightarrow N$ such that $h(0, x)=f(x)$ and $h(1, x)=g(x)$ for all $x \in M$. It can be shown that if $f: M \rightarrow N$ is a continuous mapping, then it is continuously homotopic to a smooth mapping $\tilde{f}: M \rightarrow N$, and if two smooth mapping $f, g: M \rightarrow N$ are continuously homotopic, then they are smoothly homotopic.

The reason the above results are important is that they show that de Rham cohomology is actually a topological invariant: if $M$ and $N$ are smooth manifolds that are homeomorphic, then there is an isomorphism of their de Rham cohomology rings. Without the above results, we could only make that conclusion if $M$ and $N$ were diffeomorphic-a much stronger condition.

Definition 11.22. Let $f: M \rightarrow N$ be a mapping of manifolds. Then
(1) $f$ is an immersion if $d f_{p}: T_{p} M \rightarrow T_{f(p)} N$ is injective (see subsection 4.4 for the definition of $d f_{p}$-locally, in terms of the "physical" version of tangent space, it is the mapping determined by the Jacobian of $f$ );
(2) the pair $(M, f)$ is a submanifold if $f$ is an injective immersion; if $M \subseteq N$, then we say $M$ is a submanifold if $(M, \iota)$ is a submanifold where $\iota$ is the inclusion mapping;
(3) $f$ is an embedding if it is a one-to-one immersion and also a homeomorphism onto $\operatorname{im} f \subseteq N$ where the latter set is given the subspace topology.


Figure 20. The difference between an immersion, a submanifold, and an embedding ([13]).

Definition 11.23. Let $S$ be an embedded submanifold of $M$, and let $\iota: S \hookrightarrow M$ denote the inclusion. A retraction from $M$ to $S$ is a map $r: M \rightarrow S$ such that $r \circ \iota=\mathrm{id}_{S}$. A deformation retraction from $M$ to $S$ is a map $F:[0,1] \times M \rightarrow M$ such that:
(1) $F(0,-)=\mathrm{id}_{M}$,
(2) there is a retraction $r: M \rightarrow S$ such that $F(1,-)=r$, and
(3) for all $t \in[0,1],\left.F(t,-)\right|_{S}=\mathrm{id}_{S}$.

Exercise 11.24. Show that a deformation retraction defines a homotopy equivalence.

Example 11.25. Let $M$ be the open Möbius strip as in Example 6.5. Note that there is a deformation retraction from $M$ to the unit circle $S^{1}$ (Figure 21). So they


Figure 21. The Möbius band deformation retracts to its central circle.
have the same cohomology groups. Similarly, the punctured plane $\mathbb{R}^{2} \backslash\{(0,0)\}$ has a deformation retraction to $S^{1}$ (Figure 22). Therefore, one can conclude that

$$
H^{k} M=H^{k}\left(\mathbb{R}^{2} \backslash\{(0,0)\}\right)=H^{k} S^{1}
$$

[Note to Dave:] We still need to give a direct proof that

$$
H^{k} S^{1}=\left\{\begin{array}{cc}
\mathbb{R} & \text { if } k=0,1 \\
0 & \text { otherwise }
\end{array}\right.
$$



Figure 22. The punctured plane deformation retracts to the unit circle.

Remark 11.26. Let $M$ be an orientable, closed $n$-dimensional manifold. Notice that the linear map

$$
\begin{aligned}
\int_{M}: H^{n} M & \longrightarrow \mathbb{R} \\
{[\omega] } & \longmapsto \int_{M} \omega
\end{aligned}
$$

is well-defined by Stokes theorem:

$$
\int_{M}(\omega+d \eta)=\int_{M} \omega+\int_{M} d \eta=\int_{M} \omega+\int_{\partial M} \eta=\int_{M} \omega
$$

since $\partial M=\emptyset$.
Corollary 11.27. Let $M$ be an orientable, closed $n$-dimensional manifold and let $f: M \rightarrow N$ be smooth. Then the composition

$$
H^{n} N \xrightarrow{f^{*}} H^{n} M \xrightarrow{\int_{M}} \mathbb{R}
$$

is homotopy invariant, i.e., if $f, g: M \rightarrow N$ are homotopic, then $\int_{M} \circ f^{*}=\int_{N} \circ g^{*}$.
As an application of the homotopy invariance of de Rham cohomology, we prove the hairy ball theorem ${ }^{9}$, stated as follows:

Theorem 11.28. Every vector field on an even-dimensional sphere has at least one zero.

Proof. Let $v$ be a nowhere vanishing vector field on the $n$-sphere $S^{n}$ with $n$ even. In other words, $v$ is a section of tangent bundle, $v: S^{n} \rightarrow T S^{n}$ such that $v(x) \neq 0$ for all $x \in S^{n}$. Let $D^{n+1}$ denote the solid unit ball in $\mathbb{R}^{n+1}$ with boundary $\partial D^{n+1}=$ $S^{n}$. Now consider the antipodal involution (i.e., its square is the identity) on $S^{n}$,

$$
\begin{aligned}
\tau: S^{n} & \longrightarrow S^{n} \\
x=\left(x_{1}, \ldots, x_{n+1}\right) & \longmapsto-x=\left(-x_{1}, \ldots,-x_{n+1}\right) .
\end{aligned}
$$

[^8]We can think of $v(x)$ as the pointer towards $-x$ (see Figure 23) and construct a homotopy $\operatorname{id}_{S^{n}} \stackrel{h}{\sim} \tau$ :

$$
h(t, x)=\cos (t \pi) x+\sin (t \pi) \frac{v(x)}{|v(x)|} .
$$

Here, we are identifying $T_{x} S^{n}$ with vectors in $\mathbb{R}^{n+1}$ perpendicular to $x \in \mathbb{R}^{n}$ (thinking of tangent vectors in terms of curves on $S^{n}$ passing through $x$ ). One can check that $\operatorname{im}(h) \subset S^{n}$, that $h(0,-)=\operatorname{id}_{S^{n}}$, and that $h(1,-)=\tau$.


Figure 23. The vector $v(x)$ as a pointer.
Now by Corollary 11.27 , for any $n$-form $\omega$ on $S^{n}$, we have

$$
\int_{S^{n}} \omega=\int_{S^{n}} \tau^{*} \omega
$$

At the same time, note that $\tau$ is an orientation-reversing diffeomorphism when $n$ is even $(\operatorname{det}(J \tau)=-1$ if and only if $n$ is even). So we have

$$
-\int_{S^{n}} \omega=\int_{S^{n}} \tau^{*} \omega
$$

forcing

$$
\int_{S^{n}} \omega=0
$$

However, since there are $n$-forms $\omega$ such that $\int_{S^{n}} \omega$ (bump functions, for example, or use that fact that $S^{n}$ is orientable and hence has a nonvanishing $n$-form), we have a contradiction.

Note that $S^{1}$, which is odd-dimensional, clearly has a nonvanishing vector field.
11.3. The Mayer-Vietoris sequence. We end this section with a long exact sequence that is a standard tool for computing cohomology. ${ }^{10}$

Let $M$ be a smooth manifold and let $U, V \subseteq M$ be open. Define the inclusion maps $i_{U}, i_{V}, j_{U}, j_{V}$ as shown in the following commutative diagram:


[^9]We want to use these inclusion maps to define cochain maps $i$ and $j$ such that the following sequence of cochain complexes is exact:

$$
0 \rightarrow \Omega^{\bullet}(U \cup V) \xrightarrow{i} \Omega^{\bullet} U \oplus \Omega^{\bullet} V \xrightarrow{j} \Omega^{\bullet}(U \cap V) \rightarrow 0,
$$

where the boundary map of $\Omega^{\bullet} U \oplus \Omega^{\bullet} V$ is $(d, d)$. This short exact sequence will then induce a long exact sequence called the Mayer-Vietoris sequence:

$$
\cdots \xrightarrow{\partial} H^{k}(U \cup V) \xrightarrow{i} H^{k} U \oplus H^{k} V \xrightarrow{j} H^{k}(U \cap V) \xrightarrow{\partial} H^{k+1}(U \cup V) \xrightarrow{i} \cdots,
$$

To see how this sequence would be useful in computing cohomology, suppose $M=$ $U \cup V$ and that it is easy to compute the cohomology of $U, V$, and $U \cap V$. Then the cohomology groups of $M$ are sandwiched in an exact sequence with known groups, which yields a lot of information. For instance, suppose that $H^{k-1}(U \cap V)=$ $H^{k}(U)=H^{k}(V)=0$. In that case, from the Mayer-Vietoris sequence, we know the following sequence is exact:

$$
0 \rightarrow H^{k} M \rightarrow 0
$$

which tells us that $H^{k} M=0$, too.
Lemma 11.29. Using the notation above, there is an exact sequence of cochain complexes

$$
0 \rightarrow \Omega^{\bullet}(U \cup V) \xrightarrow{i} \Omega^{\bullet} U \oplus \Omega^{\bullet} V \xrightarrow{j} \Omega^{\bullet}(U \cap V) \rightarrow 0,
$$

where $i$ and $j$ are defined to be

$$
\begin{gathered}
i(\omega)=\left(i_{U}^{*} \omega, i_{V}^{*} \omega\right) \\
j\left(\omega_{U}, \omega_{V}\right)=j_{U}^{*} \omega_{U}-j_{V}^{*} \omega_{V}
\end{gathered}
$$

Proof. That $i$ and $j$ are cochain maps is easy to check, as is injectivity of $i$. We need to show, at each level $k$, that $\operatorname{im}(i)=\operatorname{ker}(j)$ and that $j$ is surjective.

We first show that $\operatorname{im}(i)=\operatorname{ker}(j)$. Let $\omega \in \Omega^{k}(U \cup V)$. Since taking cohomology is functorial, it preserves the commutative diagram in (7), which implies

$$
(j \circ i)(\omega)=j\left(i_{U}^{*} \omega, i_{V}^{*} \omega\right)=j_{U}^{*} i_{U}^{*} \omega-j_{V}^{*} i_{V}^{*} \omega=0
$$

So $\operatorname{im}(i) \subseteq \operatorname{ker}(j)$. Let $\omega_{U} \in \Omega^{k} U$ and $\omega_{V} \in \Omega^{k} V$. If $j\left(\omega_{U}, \omega_{V}\right)=j_{U}^{*} \omega_{U}-j_{V}^{*} \omega_{V}=0$, then $\left.\omega_{U}\right|_{U \cap V}=\left.\omega_{V}\right|_{U \cap V}$. So we can glue $\omega_{U}$ and $\omega_{V}$ together along $U \cap V$ and obtain a form $\omega \in U \cup V$ and hence $\left(\omega_{U}, \omega_{V}\right)=i(\omega)$.

To see that $j$ is surjective, let $\omega \in \Omega^{k}(U \cap V)$. Let $\left\{\lambda_{U}, \lambda_{V}\right\}$ be a partition of unity on $U \cup V$ subordinate to the cover $\{U, V\}$, i.e., $\lambda_{U}, \lambda_{V}: U \cup V \rightarrow[0,1]$ are smooth and compactly supported with $\operatorname{supp}\left(\lambda_{U}\right) \subseteq U$ and $\operatorname{supp}\left(\lambda_{V}\right) \subseteq V$. Define

$$
\omega_{U}=\lambda_{U} \omega, \quad \omega_{V}=\lambda_{V} \omega
$$

By defining $\omega_{U}(p)=0$ outside of the support of $\lambda_{U}$, we may consider $\omega_{U}$ to be a form defined on all of $U$ and with the property that $j_{U}^{*} \omega_{U}=\lambda_{U} \omega$ on $U \cap V$. A similar remark holds for $\omega_{V}$. It follows that

$$
\begin{aligned}
j\left(\omega_{U},-\omega_{V}\right) & =j^{*} \omega_{U}+j^{*} \omega_{V} \\
& =\lambda_{U} \omega+\lambda_{V} \omega \\
& =\omega
\end{aligned}
$$

and we are done.
Theorem 11.30 (Mayer-Vietoris). Let $U, V \subseteq M$ be open. Then there is a long exact sequence

$$
\cdots \longrightarrow H^{k}(U \cup V) \longrightarrow H^{k} U \oplus H^{k} V \longrightarrow H^{k}(U \cap V) \longrightarrow H^{k+1}(U \cup V) \longrightarrow \cdots
$$

Sketch of proof. To prove something a little more general, suppose

$$
0 \rightarrow A^{\bullet} \rightarrow B^{\bullet} \rightarrow C^{\bullet} \rightarrow 0
$$

be a short exact sequence of cochain complexes. We define $H^{k}(A)=\operatorname{ker} d_{k}^{A} / \operatorname{im} d_{k-1}^{A}$, and similarly for $B$ and $C$. Then check that the following natural mappings are well-defined and the horizontal and vertical sequences are exact:


The result then follows from the snake lemma.
Finally, let us see how to use the Mayer-Vietoris sequence to compute the cohomology.

Example 11.31. Consider $M=S^{1} \subseteq \mathbb{R}^{2}$. We cover it with two open half circles $U$ and $V$ (overlapping a little bit) as shown in Figure 24. The Mayer-Vietoris sequence


Figure 24. Cover $S^{1}$ with two open half circles.
is:
$0 \rightarrow H^{0} S^{1} \rightarrow H^{0} U \oplus H^{0} V \rightarrow H^{0}(U \cap V) \rightarrow H^{1} S^{1} \rightarrow H^{1} U \oplus H^{1} V \rightarrow H^{1}(U \cap V) \rightarrow 0$.
Notice that each of $U$ and $V$ is diffeomorphic to an open interval of $\mathbb{R}$ which is contractible. So $H^{0} U=H^{0} V=\mathbb{R}$ and $H^{1} U=H^{1} V=0$. At the same time, the intersection $U \cap V$ can be contracted to two points, so $H^{0}(U \cap V)=\mathbb{R}^{2}$ and
$H^{1}(U \cap V)=0$. We already know that $S^{1}$ is connected, so $H^{0} S^{1}=\mathbb{R}$. The sequence then becomes

$$
0 \rightarrow \mathbb{R} \rightarrow \mathbb{R}^{2} \rightarrow \mathbb{R}^{2} \rightarrow H^{1} S^{1} \rightarrow 0
$$

Since the sequence is exact, the alternating sum of the dimensions of terms is 0 (via rank-nullity). We conclude that $\operatorname{dim} H^{1} S^{1}=1$, and, hence, $H^{1} S^{1}=\mathbb{R}$.

Next, consider the two-sphere. We cover $S^{2}$ by two open half spheres as shown in Figure 25. Both $U$ and $V$ are homotopy equivalent to a point. Their intersection is


Figure 25. Cover $S^{2}$ with two open half spheres
the cylinder, which is homotopic equivalent to $S^{1}$. So the Mayer-Vietoris sequence is:

$$
0 \rightarrow \mathbb{R} \rightarrow \mathbb{R}^{2} \rightarrow \mathbb{R} \rightarrow H^{1} S^{2} \rightarrow 0 \rightarrow \mathbb{R} \rightarrow H^{2} S^{2} \rightarrow 0
$$

We can break this into two exact sequence and conclude that

$$
H^{k} S^{2}= \begin{cases}\mathbb{R} & \text { if } k=0,2 \\ 0 & \text { otherwise }\end{cases}
$$

Exercise 11.32. Generalize Example 11.31 by proving that the cohomology groups of the $n$-sphere is

$$
H^{k} S^{n}= \begin{cases}\mathbb{R} & \text { if } k=0, n \\ 0 & \text { otherwise }\end{cases}
$$

Exercise 11.33. Compute the cohomology of the twice punctured plane,

$$
\mathbb{R}^{2} \backslash\{(-1,0),(1,0)\}
$$

What about the plane with $n$ holes?
Example 11.34. Let $T$ denote the torus and consider a cover $T=U \cup V$ by two open cylinders as shown in Figure 26. Use $A$ and $B$ to denote the two connected components of $U \cap V$. Note that $U, V, A$, and $B$ can all be contracted to $S^{1}$.


Figure 26. Cover the torus with two open cylinders.
So $U \cap V$ is homotopy equivalent to two circles and $H^{2}(U \cap V)=0$ (since the
union of two circles is a one-dimensional manifold). One can then use the cover $U \cap V=A \cup B$ and Mayer-Vietoris to show that $H^{1}(U \cap V) \cong H^{1} A \oplus H^{1} B=\mathbb{R}^{2}$ (or use the fact that the cohomology of a manifold is the direct sum of the cohomology of its components). The Mayer-Vietoris sequence becomes

$$
0 \rightarrow \mathbb{R} \rightarrow \mathbb{R}^{2} \rightarrow \mathbb{R}^{2} \rightarrow H^{1} T \rightarrow \mathbb{R}^{2} \xrightarrow{j} \mathbb{R}^{2} \xrightarrow{\partial} H^{2} T \rightarrow 0
$$

where $\partial$ is the connecting homomorphism, and $j: H^{1} U \oplus H^{1} V \rightarrow H^{1}(U \cap V)$ is induced by the map described in Lemma 11.29. By the exactness of the sequence, we have

$$
H^{2} T=\operatorname{im}(\partial) \cong \mathbb{R}^{2} / \operatorname{ker}(\partial)=\mathbb{R}^{2} / \operatorname{im}(j)
$$

To find $\operatorname{im}(j)$, note that an element $(a, b) \in H^{1} U \oplus H^{1} V$ is mapped to $(a-b, a-b) \in$ $H^{1}(U \cap V)$ under $j$. So $\operatorname{im}(j) \cong \mathbb{R}$ and $H^{2} T \cong \mathbb{R}$. A dimension count then gives us that $H^{1} T \cong \mathbb{R}^{2}$.

Note: Roughly, a torus has two independent one-dimensional holes and one twodimensional hole.

## 12. Differential Forms on Riemannian Manifolds

A Riemannian manifold is a manifold with a metric. The metric allows us to talk about geometric properties of the manifold, for example, distances and angles. some introduction
12.1. Scalar products. Let $V$ be a finite dimensional vector space over $\mathbb{R}$ with $\operatorname{dim}(V)=n$.

Definition 12.1. A symmetric bilinear form $\langle-,-\rangle: V \times V \rightarrow \mathbb{R}$ is nondegenerate if the map

$$
\begin{aligned}
b: V & \longrightarrow V^{*} \\
v & \longmapsto\langle v,-\rangle
\end{aligned}
$$

is an isomorphism. A nondegenerate symmetric bilinear form is called a scalar product on $V$.

Proposition 12.2. Let $v_{1}, \ldots, v_{n}$ be a basis for $V$ and suppose that $\langle-,-\rangle$ is a symmetric bilinear form on $V$. Let $G$ be the $n \times n$-matrix defined by

$$
G_{i, j}=\left\langle v_{i}, v_{j}\right\rangle .
$$

Then
(1) If $u=\sum_{i=1}^{n} a_{i} v_{i}$ and $w=\sum_{i=1}^{n} b_{i} v_{i}$, then

$$
\langle u, v\rangle=\left[\begin{array}{lll}
a_{1} & \cdots & a_{n}
\end{array}\right] G\left[\begin{array}{c}
b_{1} \\
\vdots \\
b_{n}
\end{array}\right] ;
$$

(2) The form $\langle-,-\rangle$ is nondegenerate if and only if $\operatorname{det}(G) \neq 0$.

Exercise 12.3. Prove the above proposition.
Definition 12.4. Let $\langle-,-\rangle$ be a scalar product on $V$. A basis $v_{1}, \ldots, v_{n}$ of $V$ is called an orthonormal basis if

$$
\left\langle v_{i}, v_{j}\right\rangle=\left\{\begin{array}{cl}
0 & \text { if } i \neq j \\
\pm 1 & \text { if } i=j
\end{array}\right.
$$

Remark 12.5. Given a scalar product, we can always find an orthonormal basis with respect to that scalar product (recall Gram-Schmidt).

Example 12.6. The Euclidean space $\mathbb{R}^{n}$ with the usual inner product has an orthonormal basis $e_{1}, \ldots, e_{n}$.

Remark 12.7. Let $\langle-,-\rangle$ be a scalar product on $V$ and let $e_{1}, \ldots, e_{n}$ be an orthonormal basis. There is a scalar product on $V^{*}$ induced by the product on $V$ :

$$
\left\langle e_{i}^{*}, e_{j}^{*}\right\rangle=\left\langle e_{i}, e_{j}\right\rangle,
$$

and the matrices corresponding to the two scalar products with respect to the two bases are the same.

Remark 12.8. Let $\langle-,-\rangle$ be a scalar product on $V$. By the universal property of symmetric products, we have a unique map $s: \operatorname{Sym}^{2} V \rightarrow \mathbb{R}$ making the following diagram commute:


Also, recall the isomorphism $\left(\operatorname{Sym}^{2} V\right)^{*} \cong \operatorname{Sym}^{2} V^{*}$ in Proposition 5.15 given by

$$
\begin{aligned}
\operatorname{Sym}^{l} V^{*} \longrightarrow\left(\operatorname{Sym}^{l} V\right)^{*} \\
\phi_{1} \cdots \phi_{l} \longmapsto\left[v_{1} \cdots v_{l} \mapsto \frac{1}{l!} \sum_{\sigma \in S_{n}} \prod_{i=1}^{l} \phi_{\sigma(i)}\left(v_{i}\right)\right] .
\end{aligned}
$$

Therefore, a scalar product on $V$ is an element of $\mathrm{Sym}^{2} V^{*}$.
Example 12.9. Let $M$ be a manifold of dimension 4 and let $p \in M$. Suppose that $T_{p} M$ has a basis $e_{1}, e_{2}, e_{3}, e_{4}$. Consider $f=e_{1}^{* 2}+e_{2}^{* 2}+e_{3}^{* 2}-e_{4}^{* 2} \in \operatorname{Sym}^{2} T_{p}^{*} M$. It corresponds to a symmetric bilinear form on $T_{p} M$ defined to be

$$
\begin{aligned}
f\left(e_{i}, e_{j}\right) & =e_{1}^{*}\left(e_{i}\right) e_{1}^{*}\left(e_{j}\right)+e_{2}^{*}\left(e_{i}\right) e_{2}^{*}\left(e_{j}\right)+e_{3}^{*}\left(e_{i}\right) e_{3}^{*}\left(e_{j}\right)-e_{4}^{*}\left(e_{i}\right) e_{4}^{*}\left(e_{j}\right) \\
& =\left\{\begin{array}{cc}
0 & \text { if } i \neq j \\
1 & \text { if } i=j \neq 4 \\
-1 & \text { if } i=j=4
\end{array}\right.
\end{aligned}
$$

with the corresponding matrix

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

Since $G$ is nonsingular, $f$ is a scalar product on $T_{p} M$.
Proposition 12.10. Suppose that $\langle-,-\rangle$ is a scalar product on $V$. Let $e_{1}, \ldots, e_{n}$ be a basis of $V$. Then the composition

$$
\begin{aligned}
& I: \Lambda^{k} V \xrightarrow{\Lambda^{k} b} \Lambda^{k} V^{*} \xrightarrow{\cong}\left(\Lambda^{k} V\right)^{*} \\
& e_{\mu} \longmapsto\left\langle e_{\mu_{1}},-\right\rangle \wedge \cdots \wedge\left\langle e_{\mu_{n}},-\right\rangle \longmapsto\left[e_{\nu} \mapsto \operatorname{det}\left(\left\langle e_{\mu_{i}}, e_{\nu_{j}}\right\rangle\right)\right]
\end{aligned}
$$

induces a scalar product on $\Lambda^{k} V$.
Proof. Bilinear: Straightforward.
Symmetric: Note that

$$
I\left(e_{\mu}\right)\left(e_{\nu}\right)=\operatorname{det}\left(\left\langle e_{\mu_{i}}, e_{\nu_{j}}\right\rangle\right)=\operatorname{det}\left(\left\langle e_{\nu_{j}}, e_{\mu_{i}}\right\rangle\right)=I\left(e_{\nu}\right)\left(e_{\mu}\right)
$$

Nondegenerate: Now suppose that $e_{1}, \ldots, e_{n}$ is an orthonormal basis of $V$. Observe that

$$
I\left(e_{\mu}\right)\left(e_{\nu}\right)=\operatorname{det}\left(\left\langle e_{\mu_{i}}, e_{\nu_{j}}\right\rangle\right)=\left\{\begin{array}{cl}
0 & \text { if } \mu \neq \nu ; \\
\prod_{i=1}^{k}\left\langle e_{\mu_{i}}, e_{\nu_{i}}\right\rangle & \text { if } \mu=\nu
\end{array}= \pm 1\right.
$$

So the matrix corresponding to $I$ with respect to the basis $\left\{e_{\mu}\right\}$ is a diagonal matrix with diagonal entries being $\pm 1$. Hence $I$ is nondegenerate.

Theorem 12.11 (Sylvester's Law of Intertia). Let $\langle-,-\rangle$ be a scalar product on $V$. Then there exists a basis for $V$ such that the matrix for $V$ with respect to this basis is diagonal of the form

$$
G=\left(\begin{array}{c|c}
I_{r} & 0 \\
\hline 0 & -I_{s}
\end{array}\right),
$$

where $I_{k}$ is the identity matrix of size $k$. The integer $s$ is called the index of $\langle-,-\rangle$ and is independent of the choice of basis.

Proof. ?
Definition 12.12. A scalar product on $V$ is positive definite if its index is 0 .
12.2. The star operator. Now let $V$ be a finite dimensional oriented vector space over $\mathbb{R}$ and let $\langle-,-\rangle$ be a scalar product on $V$.

Definition 12.13. Suppose that $e_{1}, \ldots, e_{n}$ is a positively oriented orthonormal basis for $V$. The volumn form of $V$ is the $n$-form over $V^{*}$ :

$$
\omega_{V}:=e_{1}^{*} \wedge \cdots \wedge e_{n}^{*}
$$

Lemma 12.14. Let $e_{1}, \ldots, e_{n}$ and $v_{1}, \ldots, v_{n}$ be orthonormal bases for $V$. Suppose that $v_{j}=\sum_{i=1}^{n} a_{i, j} e_{i}$ and $v_{j}^{*}=\sum_{i=1}^{n} b_{i, j} e_{i}^{*}$. Define matrices $A$ and $B$ with $A_{i, j}=$ $a_{i, j}$ and $B_{i, j}=b_{i, j}$. Then

$$
B=\left(A^{\top}\right)^{-1}
$$

Exercise 12.15. Prove Lemma 12.14.
Proposition 12.16. The volumn form of $V$ is the unique $n$-form sending each positively oriented orthonormal basis to 1 . More generally, let $v_{1}, \ldots, v_{n}$ be a positively oriented basis for $V$ and let $G$ be the matrix with $G_{i, j}=\left\langle v_{i}, v_{j}\right\rangle$. Then

$$
\omega_{V}=\sqrt{\operatorname{det}(G)} v_{1}^{*} \wedge \cdots \wedge v_{n}^{*}
$$

and in particular, if $v_{1}, \ldots, v_{n}$ is orthonormal, then $\omega=v_{1}^{*} \wedge \cdots \wedge v_{n}^{*}$.
Proof. Let $e_{1}, \ldots, e_{n}$ be a positively oriented orthonormal basis for $V$. Let $\widetilde{I}$ be the matrix for $b: V \rightarrow V^{*}$ with respect to the basis $e_{1}, \ldots, e_{n}$, i.e., $\widetilde{I}=\operatorname{diag}\left(\epsilon_{1}, \ldots, \epsilon_{n}\right)$, where $\epsilon_{i}=\left\langle e_{i}, e_{i}\right\rangle \in\{ \pm 1\}$. Suppose that $v_{j}=\sum_{i=1}^{n} a_{i, j} e_{i}$. Then we have a commutative diagram

i.e., $G=A^{\top} \widetilde{I} A$. So $\operatorname{det}(G)=\operatorname{det}(A)^{2} \operatorname{det}(\widetilde{I})$. Since both $v_{1}, \ldots, v_{n}$ and $e_{1}, \ldots, e_{n}$ are positively oriented, $\operatorname{det}(A)>0$. Hence,

$$
\operatorname{det}(A)=\sqrt{|\operatorname{det}(G)|}
$$

Finally, by Lemma 12.14, we have

$$
v_{1}^{*} \wedge \cdots \wedge v_{n}^{*}=\operatorname{det}\left(\left(A^{\top}\right)^{-1}\right) e_{1}^{*} \wedge \cdots \wedge e_{n}^{*}
$$

which is the same as

$$
e_{1}^{*} \wedge \cdots \wedge e_{n}^{*}=\sqrt{|\operatorname{det}(G)|} v_{1}^{*} \wedge \cdots \wedge v_{n}^{*}
$$

Theorem 12.17 (The star operator). Let $V$ be a finite-dimensional oriented vector space over $\mathbb{R}$ with a scalar product $\langle-,-\rangle$ and let $\omega_{V}$ be its volume form. Then for each $k \geq 0$, there exists a unique linear map

$$
*: \Lambda^{k} V^{*} \rightarrow \Lambda^{n-k} V^{*}
$$

satisfying

$$
\eta \wedge(* \zeta)=\langle\eta, \zeta\rangle \omega_{V}
$$

for all $\eta, \zeta \in \Lambda^{k} V^{*}$. We call the linear mapping the Hodge $*$-operator, or the star operator.

Proof. Let $e_{1}, \ldots, e_{n}$ be an orthonormal basis for $V$ and use $\mathcal{B}=\left\{e_{\mu}^{*}\right\}_{1 \leq \mu_{1}<\cdots<\mu_{k} \leq n}$ to denote the corresponding basis for $\Lambda^{k} V^{*}$. Since $*$ is determined by its action on the basis vectors, without loss of generality, suppose that $\zeta=e_{\mu}^{*}$ and write $* \zeta=\sum_{\nu} a_{\nu} e_{\nu}^{*} \in \Lambda^{n-k} V^{*}$. We want to compute the coefficients $\left\{a_{\nu}\right\}_{\nu}$ such that

$$
\eta \wedge(* \zeta)=\langle\eta, \zeta\rangle \omega_{V}
$$

for all $\eta \in \Lambda^{k} V^{*}$. Again, we only need to consider the case $\eta=e_{\gamma}^{*} \in \mathcal{B}$. Note that

$$
\langle\eta, \zeta\rangle=\left\langle e_{\gamma}^{*}, e_{\mu}^{*}\right\rangle=\operatorname{det}\left(\left\langle e_{\gamma_{i}}, e_{\mu_{j}}\right\rangle\right)=\left\{\begin{aligned}
0 & \text { if } \gamma \neq \mu \\
\epsilon_{\mu} & \text { if } \gamma=\mu
\end{aligned}\right.
$$

where $\epsilon_{\mu}=\prod_{i=1}^{k}\left\langle e_{\gamma_{i}}, e_{\mu_{i}}\right\rangle$. So $\langle\eta, \zeta\rangle \in\{ \pm 1\}$ if $\gamma=\mu$. On the other hand, by definition we have

$$
\eta \wedge * \zeta=a_{\widetilde{\gamma}} e_{\gamma}^{*} \wedge e_{\widetilde{\gamma}}^{*},
$$

where $\widetilde{\gamma}$ is the index formed by $[n] \backslash\left\{\gamma_{1}, \ldots, \gamma_{k}\right\}$ arranged in increasing order. Let $\tau_{\gamma}$ be the permutation sending $(1, \ldots, n)$ to $\left(\gamma_{1}, \ldots, \gamma_{k}, \widetilde{\gamma}_{1}, \ldots, \widetilde{\gamma}_{n-k}\right)$. We have

$$
\eta \wedge * \zeta=a_{\widetilde{\gamma}} \operatorname{sign}\left(\tau_{\gamma}\right) \omega_{V}
$$

Therefore, by assumption,

$$
a_{\widetilde{\gamma}} \operatorname{sign}\left(\tau_{\gamma}\right)=\langle\eta, \zeta\rangle=\left\{\begin{array}{cl}
0 & \text { if } \gamma \neq \mu \\
\epsilon_{\mu} & \text { if } \gamma=\mu
\end{array}\right.
$$

Hence,

$$
a_{\tilde{\gamma}}=\operatorname{sign}\left(\tau_{\gamma}\right)\langle\eta, \zeta\rangle=\left\{\begin{array}{cl}
0 & \text { if } \gamma \neq \mu \\
\operatorname{sign}\left(\tau_{\gamma}\right) \epsilon_{\mu} & \text { if } \gamma=\mu
\end{array}\right.
$$

So we define $* \zeta=* e_{\mu}^{*}$ to be

$$
* e_{\mu}^{*}:=\operatorname{sign}\left(\tau_{\mu}\right) \epsilon_{\mu} e_{\widetilde{\mu}}^{*}
$$

Remark 12.18. In summary, the star operator is defined to be

$$
\begin{aligned}
& *: \Lambda^{k} V^{*} \longrightarrow \Lambda^{n-k} V^{*} \\
& e_{\mu}^{*} \longmapsto \operatorname{sign}\left(\tau_{\mu}\right) \epsilon_{\mu} e_{\stackrel{\mu}{\mu}}^{*},
\end{aligned}
$$

where $\epsilon_{\mu}=\prod_{i=1}^{k}\left\langle e_{\mu_{i}}, e_{\mu_{i}}\right\rangle= \pm 1, \widetilde{\mu}=\left(\widetilde{\mu}_{1}, \ldots, \widetilde{\mu}_{n-k}\right)$ with $\widetilde{\mu}_{1}<\ldots<\widetilde{\mu}_{n-k}$ and $\left\{\widetilde{\mu}_{1}, \ldots, \widetilde{\mu}_{n-k}\right\}=[n] \backslash\left\{\mu_{1}, \ldots, \mu_{k}\right\}$, and $\tau_{\mu}$ is the permutation that sends $(1, \ldots, n)$ to $\left(\mu_{1}, \ldots, \mu_{k}, \widetilde{\mu}_{1}, \ldots, \widetilde{\mu}_{n-k}\right)$. Hence, $e_{\mu}^{*} \wedge e_{\widetilde{\mu}}^{*}=\operatorname{sign}\left(\tau_{\mu}\right) \omega_{V}$.

Example 12.19. Consider $\mathbb{R}^{3}$ with the usual scalar product and let $e_{1}, e_{2}, e_{3}$ be the standard basis for $\mathbb{R}^{3}$. Then

$$
*\left(e_{1}^{*} \wedge e_{2}^{*}\right)=e_{3}^{*}, \quad *\left(e_{1}^{*} \wedge e_{3}^{*}\right)=-e_{2}^{*}, \quad *\left(e_{2}^{*} \wedge e_{3}^{*}\right)=e_{1}^{*}
$$

Example 12.20. Consider $\mathbb{R}^{4}$ with the scalar product of index 1 and let $e_{1}, \ldots, e_{4}$ be an orthonormal basis. Let $G=\operatorname{diag}(1,1,1,-1)$ be the corresponding matrix for this scalar product with respect to the basis. Then

$$
*\left(e_{1}^{*} \wedge e_{3}^{*}\right)= \pm e_{2}^{*} \wedge e_{4}^{*}
$$

We have $\mu=(1,3)$ and we want to find the sign. Note that $\tau_{\mu}=(23), \operatorname{sos} \operatorname{sign}\left(\tau_{\mu}\right)=$ -1 . Also, $\epsilon_{\mu}=\left\langle e_{1}, e_{1}\right\rangle \cdot\left\langle e_{3}, e_{3}\right\rangle=1$. Therefore,

$$
*\left(e_{1}^{*} \wedge e_{3}^{*}\right)=-e_{2}^{*} \wedge e_{4}^{*}
$$

For another example,

$$
*\left(e_{1}^{*} \wedge e_{2}^{*} \wedge e_{4}^{*}\right)=\operatorname{sign}(34) \cdot 1 \cdot 1 \cdot(-1) e_{3}^{*}=e_{3}^{*}
$$

Lemma 12.21. Let $V$ be an oriented vector space over $\mathbb{R}$ and let $\langle-,-\rangle$ be a scalar product of index $s$. For each $k$,

$$
* *=(-1)^{k(n-k)+s} \operatorname{id}_{\Lambda^{k} V^{*}}
$$

Hence, * is an isomorphism.
Exercise 12.22. Prove Lemma 12.21.
Exercise 12.23. Let $V$ be an oriented vector space over $\mathbb{R}$ and let $\langle-,-\rangle$ be a scalar product of index $s$. Show that for all $\eta, \zeta \in \Lambda^{k} V^{*}$

$$
\langle * \eta, * \zeta\rangle=(-1)^{s}\langle\eta, \zeta\rangle .
$$

### 12.3. Poincaré duality.

Definition 12.24. A semi-Riemannian manifold of index $s$ is a manifold $M$ together with a family

$$
\langle-,-\rangle=\left\{\langle-,-\rangle_{p}\right\}_{p \in M}
$$

of scalar products $\langle-,-\rangle_{p}$ of index $s$ on $T_{p} M$ for all $p \in M$ varying smoothly with $p$ (i.e., the entries in the corresponding matrix are smooth locally around $p$.) If all the forms are positive definite, i.e., $s=0$, then we call $(M,\langle-,-\rangle)$ a Riemannian manifold.

Remark 12.25. The collection of scalar products $\langle-,-\rangle_{p}$ of a Riemannian manifold is sometimes called a Riemannian metric. Note that every oriented manifold can be given such a metric using the usual scalar product on $\mathbb{R}^{n}$ and a partition of unity.

Remark 12.26. We saw in Remark 12.8 that a scalar product on a vector space $V$ corresponds with an element of the symmetric product $\mathrm{Sym}^{2} V^{*}$. Thus, we can define a semi-Riemannian manifold as a manifold together with a section of the bundle $\operatorname{Sym}^{2} T^{*} M$ (the fiber at each $p \in M$ is $\operatorname{Sym}^{2} T_{p}^{*} M$ ) such that the corresponding bilinear form on $T_{p} M$ is nondegenerate for all $p \in M$.

Suppose that $M$ is an oriented semi-Riemannian $n$-manifold of index $s$. For each $p \in M$, on the tangent space $T_{p} M$ we get a volume form $\omega_{T_{p} M}$ and a star operator *: $\Lambda^{k} T_{p}^{*} M \rightarrow \Lambda^{n-k} T_{p}^{*} M$ that vary smoothly with $p$. Gluing them together, we obtain a volume form $\omega_{M} \in \Omega^{n} M$ on $M$ and a star operator $*: \Omega^{k} M \rightarrow \Omega^{n-k} M$ on $M$.

Definition 12.27. The coderivative $\delta_{k}: \Omega^{n-k} M \rightarrow \Omega^{n-k-1} M$ is defined by

$$
\delta_{k}:=(-1)^{k} * d *^{-1}
$$

The coderivative allows us to have the following commutative diagram:


Now let $M$ be a closed Riemannian $n$-manifold. We will show that $*$ induces an isomorphism of cohomology groups $H^{k} M \cong H^{n-k} M$, which gives the Poincaré duality. By Lemma 12.21, for all $\zeta \in \Omega^{k+1} M$ we have

$$
d * \zeta=(-1)^{k(n-k)} * * d * * *^{-1} \zeta=(-1)^{k(n-k)+(k+1)(n-k-1)} * * d *^{-1} \zeta=(-1)^{k} * \delta \zeta
$$

Exercise 12.28. Let $\eta \in \Omega^{k} M$ and $\zeta \in \Omega^{k+1} M$. Recall the product rule

$$
d(\eta \wedge * \zeta)=d \eta \wedge * \zeta+(-1)^{k} \eta \wedge d * \zeta
$$

Show that

$$
d(\eta \wedge * \zeta)=d \eta \wedge * \zeta+\eta \wedge * \delta \zeta
$$

Remark 12.29. The coderivative $\delta$ is independent of the orientation of $M$. We define a positive-definite scalar product on $\Omega^{k} M$ using the scalar product on each of $T_{p} M$ as follows: For all $\eta, \zeta \in \Omega^{k} M$, put

$$
\langle\langle\eta, \zeta\rangle\rangle:=\int_{M}\langle\eta, \zeta\rangle \omega_{M}=\int_{M} \eta \wedge * \zeta .
$$

Since by assumption $M$ is closed, by Stokes' theorem, $\int_{M} d(\eta \wedge * \zeta)=0$. So the above exercise shows that for all $\eta \in \Omega^{k} M$ and $\zeta \in \Omega^{k+1} M$,

$$
\langle\langle d \eta, \zeta\rangle\rangle+\langle\langle\eta, \delta \zeta\rangle\rangle=0 .
$$

This meas that $-\delta$ is the adjoint of $d$, i.e.,

$$
\langle\langle d \eta, \zeta\rangle\rangle=\langle\langle\eta,-\delta \zeta\rangle\rangle .
$$

At the same time, the scalar product on $\Omega^{k} M$ gives us for free that

$$
\operatorname{ker}\left(d_{k}\right)=\left(\operatorname{im}\left(\delta_{n-k-1}\right)\right)^{\perp}, \quad \operatorname{ker}\left(\delta_{n-k}\right)=\left(\operatorname{im}\left(d_{k-1}\right)\right)^{\perp} .
$$

Theorem 12.30. Let $M$ be an oriented closed Riemannian n-manifold. Then

$$
\Omega^{k} M=\operatorname{ker}\left(d_{k}\right) \oplus \operatorname{im}\left(\delta_{n-k-1}\right)=\operatorname{ker}\left(\delta_{n-k}\right) \oplus \operatorname{im}\left(d_{k-1}\right)
$$

as an orthogonal direct sum with respect to the scalar product on $\Omega^{k} M$ defined by

$$
\langle\langle\eta, \zeta\rangle\rangle:=\int_{M} \eta \wedge * \zeta .
$$

Remark 12.31. The proof is beyond the scope of these notes, but the interested reader can consult Chapter 6 of [13].

Corollary 12.32. For $M$ as above,

$$
\begin{aligned}
\operatorname{ker}\left(d_{k}\right) & =\operatorname{im}\left(d_{k-1}\right) \oplus\left(\operatorname{ker}\left(d_{k}\right) \cap \operatorname{ker}\left(\delta_{n-k}\right)\right) \\
\operatorname{ker}\left(\delta_{n-k}\right) & =\operatorname{im}\left(\delta_{n-k-1}\right) \oplus\left(\operatorname{ker}\left(d_{k}\right) \cap \operatorname{ker}\left(\delta_{n-k}\right)\right) .
\end{aligned}
$$

This decomposition led us to consider the harmonic forms on $M$.
Definition 12.33. The harmonic $k$-forms on $M$ are

$$
\mathcal{H}^{k} M=\left\{\eta \in \Omega^{k} M \mid d \eta=0 \text { and } \delta \eta=0\right\} .
$$

Theorem 12.34 (Hodge Theorem). Every de Rham cohomology class of an oriented closed Riemannian manifold is represented by a well-defined harmonic form. More precisely, for each $k$, the canonical map

$$
\begin{aligned}
\mathcal{H}^{k} M & \longrightarrow H^{k} M \\
\eta & \longmapsto[\eta]
\end{aligned}
$$

is an isomorphism.
Proof. This follows from Corollary 12.32.
Remark 12.35. As a result, we have an orthonormal direct sum decomposition

$$
\Omega^{k} M=d \Omega^{k-1} M \oplus \delta \Omega^{k+1} M \oplus \mathcal{H}^{k} M
$$

Proposition 12.36. The star operator $*: \mathcal{H}^{k} \rightarrow \mathcal{H}^{n-k} M$ is an isomorphism.
Proof. Let $\eta \in \mathcal{H}^{k} M$. So $d \eta=\delta \eta=0$. Then

$$
d * \eta= \pm \delta * \eta=0, \quad \delta * \eta= \pm * d \eta=0
$$

Hence, $*\left(\mathcal{H}^{k} M\right) \subseteq \mathcal{H}^{n-k} M$. It is an isomorphism since $* *= \pm \mathrm{id}_{M}$.
Theorem 12.37 (Poincaré Duality). Let $M$ be an oriented closed Riemannian $n$-manifold. Then $H^{k} M \cong H^{n-k} M$ as described in the following diagram:


Corollary 12.38. Let $M$ be a connected oriented closed Riemannian n-manifold. Then $H^{n} M \cong \mathbb{R}$.

## 13. Toric Varieties

An $n$-dimensional manifold over $\mathbb{C}$ is a second countable Hausdorff space where the charts are homeomorphic to open subsets of $\mathbb{C}^{n}$ and the transition maps are holomorphic (locally given by convergent power series). In this section we look at smooth toric varieties as examples of complex manifolds. The name "toric" comes from the action on the variety by the algebraic torus $\left(\mathbb{C}^{*}\right)^{n}$ (where $\mathbb{C}^{*}$ is the set of nonzero complex numbers, i.e., $\left.\mathbb{R}^{2} \backslash\{(0,0)\}\right)$.
13.1. Toric varieties from fans. Let $N$ be a lattice of rank $n$, meaning that it is a free $\mathbb{Z}$-module generated by some basis $w_{1}, \ldots, w_{n} \in N$. So we have $N \cong \mathbb{Z}^{n}$ as $\mathbb{Z}$-modules. Consider the vector space over $\mathbb{R}$

$$
N_{\mathbb{R}}:=N \otimes_{\mathbb{Z}} \mathbb{R} \cong \mathbb{R}^{n}
$$

obtained from extension of scalars. We can think of $N$ as a collection of points in $N \otimes_{\mathbb{Z}} \mathbb{R}$ that have integer coordinates. Let $M=\operatorname{hom}(N, \mathbb{Z})$ be the dual lattice and use $M_{\mathbb{R}}=M \otimes_{\mathbb{Z}} \mathbb{R}$ to denote the corresponding vector space. There is a dual pairing $\langle-,-\rangle: M_{\mathbb{R}} \times N_{\mathbb{R}} \rightarrow \mathbb{R}$ defined by

$$
\begin{aligned}
\langle-,-\rangle: M_{\mathbb{R}} \times N_{\mathbb{R}} & \longrightarrow \mathbb{R} \\
(f, v) & \longmapsto f(v) .
\end{aligned}
$$

Let $e_{1}, \ldots, e_{n}$ be the standard basis for $\mathbb{R}^{n}$ (hence a basis for $N_{\mathbb{R}}$ ). Then in coordinates, the dual pairing can be written as

$$
\left\langle\sum_{i=1}^{n} a_{i} e_{i}^{*}, \sum_{i=1}^{n} b_{i} e_{i}\right\rangle=\sum_{i=1}^{n} a_{i} b_{i} .
$$

Further identifying $\left(\mathbb{R}^{n}\right)^{*}$ with $\mathbb{R}^{n}$, the pairing $M \times N$ is just the ordinary inner product.

A convex polyhedral cone is the set

$$
\sigma=\left\{\sum_{i=1}^{s} r_{i} v_{i} \mid r_{i} \geq 0\right\}
$$

for some $v_{1}, \ldots, v_{s} \in N_{\mathbb{R}}$. The cone is rational if $v_{i} \in N$. It is strongly convex if it does not contain a line through the origin. The dual cone $\sigma^{\vee} \subseteq M_{\mathbb{R}}$ is defined to be the set

$$
\sigma^{\vee}:=\left\{u \in M_{\mathbb{R}} \mid\langle u, v\rangle \geq 0 \text { for all } v \in \sigma\right\}
$$

A face $\tau$ of $\sigma$ is the intersection of $\sigma$ with any supporting hyperplane: $\tau=\sigma \cap u^{\perp}=$ $\{v \in \sigma \mid\langle u, v\rangle=0\}$ for some $u \in \sigma^{\vee}$. A face of a cone is also a cone.

Let $\sigma$ be a strongly convex rational polyhedral cone in $N_{\mathbb{R}}$. There is a semigroup associated with $\sigma$ :

$$
S_{\sigma}:=\sigma^{\vee} \cap M
$$

The group algebra corresponding to $\sigma$ is

$$
\mathbb{C}\left[S_{\sigma}\right]:=\mathbb{C}\left[e_{u} \mid u \in S_{\sigma}\right]
$$

whose elements are polynomials in the symbols $e_{u}$ with the relations

$$
e_{u} e_{v}:=e_{u+v} \quad \text { and } e_{0}:=1
$$

It turns out that $S_{\sigma}$ is finitely generated over $\mathbb{Z}_{\geq 0}$, i.e., So $\mathbb{C}\left[S_{\sigma}\right]$ is finitely generated; that is, fixing $u_{1}, \ldots, u_{s}$, a minimal set of generators for $S_{\sigma}$, we have

$$
\mathbb{C}\left[S_{\sigma}\right] \cong \mathbb{C}\left[x_{1}, \ldots, x_{s}\right] / I
$$

where $I$ is the ideal generated by binomials corresponding to the relations between the $u_{i}$ 's:

$$
x_{1}^{a_{1}} \cdots x_{n}^{a_{n}}-x_{1}^{b_{1}} \cdots x_{n}^{b_{n}}
$$

where $\sum_{i=1}^{s} a_{i} u_{i}=\sum_{i=1}^{s} b_{i} u_{i}$. By Hilbert's theorem, a finite number of these generators will suffice. Denote the zero set or vanishing set of $I$ as

$$
U_{\sigma}:=Z(I):=\left\{p \in \mathbb{C}^{s} \mid f(p)=0 \text { for all } f \in I\right\}
$$

Alternatively, if we treat $u_{i} \in \mathbb{Z}^{n}$ as exponents for monomials in $x_{1}, \ldots, x_{n}$, we have

$$
\mathbb{C}\left[S_{\sigma}\right] \cong \mathbb{C}\left[x^{u_{1}}, \ldots, x^{u_{s}}\right] \subseteq \mathbb{C}\left[x_{1}^{ \pm 1}, \ldots, x_{n}^{ \pm 1}\right]
$$

where for $a \in \mathbb{Z}^{n}$, we use the notation $x^{a}:=\prod_{=1}^{n} x_{i}^{a_{i}}$.
Definition 13.1. Let $\sigma$ be a strongly convex polyhedral cone in $N$. The affine toric variety associated to $\sigma$, denoted $U_{\sigma}$, is the zero set constructed as above.

Example 13.2. Let $n=1$. Note that $\tau=\{0\}$ is a cone. Its dual cone is $\tau^{\vee}=$ $M_{\mathbb{R}} \cong \mathbb{R}$, which is generated by $v_{1}=e_{1}^{*}=1$ and $v_{2}=-e_{1}^{*}=-1$ with the single relation $v_{1}+v_{2}=0$. So

$$
\mathbb{C}\left[S_{\tau}\right]=\mathbb{C}\left[x_{1}^{v_{1}}, x_{1}^{v_{2}}\right]=\mathbb{C}\left[x_{1}, x_{1}^{-1}\right] \cong \mathbb{C}[x, y] /(x y-1)
$$

and the zero set of $x y-1$ is

$$
\begin{aligned}
U_{\tau}=\left\{(u, 1 / u) \in \mathbb{C}^{2} \mid u \neq 0\right\} & \cong \mathbb{C}^{*}:=\mathbb{C} \backslash\{0\} \\
(u, 1 / u) & \mapsto u \\
(z, 1 / z) & \hookleftarrow z
\end{aligned}
$$

In general, for an arbitrary $n$ and $\tau=\{0\}$, we have

$$
\mathbb{C}\left[S_{\tau}\right]=\mathbb{C}\left[x_{1}^{ \pm 1}, \ldots, x_{n}^{ \pm 1}\right] \cong \mathbb{C}\left[x_{1}, y_{1}, \ldots, x_{n}, y_{n}\right] /\left(x_{1} y_{1}-1, \ldots, x_{n} y_{n}-1\right)
$$

and $U_{\tau}=\left(\mathbb{C}^{*}\right)^{n}$, the $n$-torus. For any cone $\sigma$, the affine toric variety $U_{\sigma}$ turns out to contain the torus as a dense open subset.

If $\tau$ is a facet (i.e., a codimension one face) of $\sigma$, then there is an inclusion $U_{\tau} \subseteq U_{\sigma}$ which can be described as follows: Suppose that $\tau=\sigma \cap u^{\perp}$ for some $u \in \sigma^{\vee}$. Without loss of generality, say $u=v_{s}$ is the last generator of $\sigma$. Then $S_{\sigma} \subseteq S_{\tau}$ since the generators of $S_{\tau}$ are the generators of $S_{\sigma}$ and $-v_{s}$, and as a result, $\mathbb{C}\left[S_{\sigma}\right] \subseteq \mathbb{C}\left[S_{\tau}\right]$ as a subalgebra. To be more explicit, the relations amongst the generators of $S_{\tau}$ are the relations amongst those of $S_{\sigma}$ plus the additional relation $v_{s}+\left(-v_{s}\right)=0$. This means that we have

$$
\mathbb{C}\left[S_{\sigma}\right] \subseteq \mathbb{C}\left[S_{\tau}\right] \cong \mathbb{C}\left[y_{1}, \ldots, y_{s}, y_{s+1}\right] /\left(I+\left(y_{s} y_{s+1}-1\right)\right)
$$

where we recall that $\mathbb{C}\left[S_{\sigma}\right] \cong \mathbb{C}\left[y_{1}, \ldots, y_{s}\right] / I$. Therefore, $U_{\tau}$ is subject to one more restriction than $U_{\sigma}$, and we can identitify $U_{\tau}$ as

$$
U_{\tau} \cong U_{\sigma} \backslash\left\{y_{s}=0\right\}
$$

i.e., points in $U_{\sigma}$ that have nonzero entry in the last coordinate.

Example 13.3. Consider the case $n=2$, and take $\sigma$ to be generated by $e_{2}$ and $2 e_{1}-e_{2}$ as shown in the left of Figure 27. The cone in the right is the dual lattice $\sigma^{\vee}$ with the three generators of $S_{\sigma}$ being circled. The three generators are


Figure 27. A cone $\sigma$ and its dual. The generators of $S_{\sigma}$ are circled.
$u_{1}=e_{1}^{*}=(1,0), u_{2}=e_{1}^{*}+e_{2}^{*}=(1,1)$ and $u_{3}=e_{1}^{*}+2 e_{2}^{*}=(1,2)$. The corresponding group algebra is

$$
\begin{equation*}
\mathbb{C}\left[S_{\sigma}\right]=\mathbb{C}\left[x^{u_{1}}, x^{u_{2}}, x^{u_{3}}\right]=\mathbb{C}\left[x_{1}, x_{1} x_{2}, x_{1} x_{2}^{2}\right] \cong \mathbb{C}\left[y_{1}, y_{2}, y_{3}\right] /\left(y_{2}^{2}-y_{1} y_{3}\right), \tag{8}
\end{equation*}
$$

and $U_{\sigma}=\left\{\left(y_{1}, y_{2}, y_{3}\right) \in \mathbb{C}^{3} \mid y_{2}^{2}=y_{1} y_{3}\right\}$ is a conic surface whose real points are pictured in Figure 28. Now let $\tau=\mathbb{R}_{\geq 0} e_{2}$ be the cone generated by $e_{2}$. Then $\tau^{\vee}$ is


Figure 28. The real points of the affine toric variety $U_{\sigma}$ in Example 13.3.
the upper half plane, and $S_{\tau}$ is generated by $e_{1}^{*},-e_{1}^{*}, e_{2}^{*}$. We have

$$
\begin{equation*}
\mathbb{C}\left[S_{\tau}\right]=\mathbb{C}\left[x_{1}, x_{1}^{-1}, x_{2}\right] \cong \mathbb{C}\left[z_{1}, z_{2}, z_{3}\right] /\left(z_{1} z_{2}-1\right) \tag{9}
\end{equation*}
$$

So

$$
\begin{aligned}
U_{\tau} & \simeq \\
\left(z_{1}, z_{2}, z_{3}\right) & \rightarrow\left(z_{1}, z_{3}\right),
\end{aligned}
$$

and the $\operatorname{map} U_{\tau} \rightarrow U_{\sigma}$ can be described by

$$
\begin{aligned}
& \mathbb{C}^{*} \times \mathbb{C} \longrightarrow U_{\sigma} \backslash\left\{\left(y_{1}, y_{2}, y_{3}\right) \in U_{\sigma} \mid y_{1} \neq 0\right\} \\
& \left(z_{1}, z_{3}\right) \longmapsto\left(z_{1}, z_{1} z_{3}, z_{1} z_{3}^{2}\right) .
\end{aligned}
$$

Here is a way to find the above mapping. Under the isomorphisms Equation 8 and Equation 9, we have

$$
y_{1} \leftrightarrow x_{1}, \quad y_{2} \leftrightarrow x_{1} x_{2}, \quad y_{3} \leftrightarrow x_{1} x_{2}^{2}
$$

and

$$
z_{1} \leftrightarrow x_{1}, \quad z_{2} \leftrightarrow x_{1}^{-1}, \quad z_{3} \leftrightarrow x_{2}
$$

Eliminating the $x_{i}$, we find $z_{1}=y_{1}, y_{2}=z_{1} z_{3}$, and $y_{3}=z_{1} z_{3}^{2}$.
A fan $\Delta$ in $N_{\mathbb{R}}$ is a collection of strongly convex rational polyhedral cones $\sigma$ in $N_{\mathbb{R}}$ such that
(1) If $\tau$ is a face of $\sigma$, then $\tau \in \Delta$;
(2) If $\sigma, \tau \in \Delta$, then $\sigma \cap \tau \in \Delta$.

Construction of a toric variety from a fan. Let $\Delta$ be a fan in $N_{\mathbb{R}}$. The toric variety associated to $\Delta$, denoted $X(\Delta)$ is constructed by starting with the disjoint union of the affine toric varieties $U_{\sigma}$ for all $\sigma \in \Delta$. We then glue these varieties together according to the following instructions. For each pair of cones $\sigma, \sigma^{\prime} \in \Delta$ we have induced mappings

$$
\mathbb{C}\left[S_{\sigma}\right] \hookrightarrow \mathbb{C}\left[S_{\sigma \cap \sigma^{\prime}}\right] \hookleftarrow \mathbb{C}\left[S_{\sigma^{\prime}}\right] \quad \leftrightarrow \rightsquigarrow \quad U_{\sigma} \hookleftarrow U_{\sigma \cap \sigma^{\prime}} \hookrightarrow U_{\sigma^{\prime}}
$$

Write

$$
U_{\sigma} \stackrel{\phi_{\sigma, \sigma^{\prime}}}{---->} U_{\sigma^{\prime}}
$$

for the mapping that sends the image of each point $p \in U_{\sigma \cap \sigma^{\prime}}$ in $U_{\sigma}$ to the image of $p$ in $U_{\sigma^{\prime}}$. The dashed arrow indicates that the mapping is not defined on the whole domain and codomain. (However, it is a bijection on the images of $U_{\sigma \cap \sigma^{\prime}}$.) In this way, we glue $U_{\sigma}$ to $U_{\sigma^{\prime}}$ along $U_{\sigma \cap \sigma^{\prime}}$ and $\phi_{\sigma, \sigma^{\prime}}$. For each pair of cones of $\Delta$, we perform these gluing to arrive at the toric variety $X(\Delta)$. We then have $U_{\sigma} \subseteq X(\Delta)$ for each $\sigma \in \Delta$, and the mappings $\phi_{\sigma, \sigma^{\prime}}$ then become transition functions. We omit the verification that all of these gluings are compatible.

Definition 13.4. Let $\Delta$ be a fan in $N_{\mathbb{R}}$. The toric variety associated to $\Delta$, denoted $X(\Delta)$, is constructed from $\Delta$ by gluing the affine toric varieties $U_{\sigma}$ for all $\sigma \in \Delta$.

Example 13.5. Consider the fan $\Delta$ in $\mathbb{R}$ consisting of the three cones as shown in Figure 29. We have


Figure 29. A fan $\Delta$ such that $X(\Delta)=\mathbb{P}^{1}$.

$$
\mathbb{C}\left[S_{\sigma_{1}}\right]=\mathbb{C}[x], \quad \mathbb{C}\left[S_{\sigma_{2}}\right]=\mathbb{C}[y], \quad \mathbb{C}\left[S_{\tau}\right]=\mathbb{C}[x, y] /(x y-1)
$$

So $U_{\sigma_{1}}=U_{\sigma_{2}}=\mathbb{C}, U_{\tau}=\mathbb{C}^{*}$, and the gluing instruction is given by

$$
\begin{aligned}
\phi_{\sigma_{1}, \sigma_{2}}: U_{\sigma_{1}} & \longrightarrow U_{\sigma_{2}} \\
x & \longmapsto 1 / x .
\end{aligned}
$$

Therefore, $X(\Delta)=\mathbb{P}^{1}$.
Example 13.6. Let $n=2$. Consider the fan $\Delta$ with cones $\sigma_{1}, \sigma_{2}, \sigma_{3}$ and their faces, where $\sigma_{1}$ is generated by $e_{1}, e_{2}, \sigma_{2}$ is generated by $e_{2},-e_{1}-e_{2}$, and $\sigma_{3}$ is generated by $e_{1},-e_{1}-e_{2}$ as shown on the left in Figure 30.



Figure 30. Fan $\Delta$ in $\mathbb{R}^{2}$ such that $X(\Delta)=\mathbb{P}^{2}$.
We claim that the corresponding toric variety is $\mathbb{P}^{2}$. Indeed, one can see that

$$
\mathbb{C}\left[S_{\sigma_{1}}\right]=\mathbb{C}[u, v], \quad \mathbb{C}\left[S_{\sigma_{2}}\right]=\mathbb{C}\left[u^{-1}, u^{-1} v\right], \quad \mathbb{C}\left[S_{\sigma_{3}}\right]=\mathbb{C}\left[v^{-1}, v u^{-1}\right]
$$

and as a result,

$$
U_{\sigma_{1}}=U_{\sigma_{2}}=U_{\sigma_{3}}=\mathbb{C}^{2}
$$

However, notice that their intersections are $\mathbb{C}^{*} \times \mathbb{C}$ and the gluing maps are given by

$$
\begin{array}{ccccc}
U_{\sigma_{1}} & \cdots & U_{\sigma_{2}} & \cdots & U_{\sigma_{3}} \\
(u, v) & \longmapsto & (1 / u, v / u) & \longmapsto & (1 / v, v / u) .
\end{array}
$$

Relabeling $u$ and $v$ as $x_{2} / x_{1}$ and $x_{3} / x_{1}$, we see that the gluing morphisms are our usual transition maps of $\mathbb{P}^{2}$.

Exercise 13.7. Let $e_{n+1}=-e_{1}-\ldots-e_{n}$. Show that $\mathbb{P}^{n}$ can be constructed as a toric variety using the fan $\Delta$ with cones generated by the proper subsets of $e_{1}, \ldots, e_{n+1}$. To start with, let $\sigma_{i}$ be the cone over $e_{1}, \ldots, \hat{e}_{i}, \ldots, e_{n+1}$, a maximal cone in $\Delta$. One can think of $U_{\sigma_{i}}$ as the standard chart $U_{i}$ of $\mathbb{P}^{n}$ and the gluing morphisms are exactly the usual transition functions. You might find this package of SageMath helpful when finding the generators of the dual cone $\sigma_{i}^{\vee}$.

Example 13.8 (Hirzebruch Surfaces). Consider the fan depicted in Figure 31, where the slanting arrow passes through the point $(-1, a)$ for some $a \in \mathbb{Z}_{>0}$. The corresponding toric variety is called the Hirzebruch surface and is denoted $H_{a}$. The reader should construct the dual cones and verify that

$$
\mathbb{C}\left[S_{\sigma_{1}}\right]=\mathbb{C}[u, v], \quad \mathbb{C}\left[S_{\sigma_{2}}\right]=\mathbb{C}\left[u, v^{-1}\right]
$$



Figure 31. Fan for the the Hirzebruch surface.

$$
\mathbb{C}\left[S_{\sigma_{3}}\right]=\mathbb{C}\left[u^{-1}, u^{-a} v^{-1}\right], \quad \mathbb{C}\left[S_{\sigma_{4}}\right]=\mathbb{C}\left[u^{-1}, u^{a} v\right]
$$

Gluing $U_{\sigma_{1}}$ to $U_{\sigma_{2}}$ and gluing $U_{\sigma_{3}}$ to $U_{\sigma_{4}}$ yields two copies of $\mathbb{C} \times \mathbb{P}^{1}$. The other two gluing maps are given by

$$
\begin{array}{cccccc}
U_{\sigma_{1}} & -\cdots & U_{\sigma_{4}} & U_{\sigma_{2}} & \cdots & U_{\sigma_{3}} \\
(u, v) & \longmapsto & \left(1 / u, u^{a} v\right), & (x, y) & \longmapsto & \left(1 / x, x^{a} / y\right) .
\end{array}
$$

We would now like to define a projection $H_{a} \rightarrow \mathbb{P}^{1}$ whose fibers are all $\mathbb{P}^{1}$, i.e., the mapping is surjective and the inverse image of each point in $\mathbb{P}^{1}$ is a copy of $\mathbb{P}^{1}$. We say $H_{a}$ is a $\mathbb{P}^{1}$ bundle over $\mathbb{P}^{1}$. To construct the projection, consider the mapping $\mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$ given by the first projection $(x, y) \mapsto x$. It maps the lattice $N=\mathbb{Z}^{2} \in \mathbb{R}^{2}$ to the lattice $N^{\prime}=\mathbb{Z} \subset \mathbb{R}$. It also maps each cone $\sigma_{1}$ and $\sigma_{2}$ to the cone $\tau_{1}:=\mathbb{R} \cdot 1$ and each cone $\sigma_{3}$ and $\sigma_{4}$ to the cone $\tau_{2}:=\mathbb{R} \cdot(-1)$. Let $\Delta^{\prime}$ be the fan in $\mathbb{R}$ with maximal cones $\tau_{1}$ and $\tau_{2}$. As we saw earlier, $X\left(\Delta^{\prime}\right)=\mathbb{P}^{1}$. In a way specified in general below, the mapping $N \rightarrow N^{\prime}$ has induced a mapping of fans $\Delta \rightarrow \Delta^{\prime}$, and in turn this induces a mapping of toric varieties

$$
H_{a}=X(\Delta) \rightarrow X\left(\Delta^{\prime}\right)=\mathbb{P}^{1}
$$

We have the following commutative mapping in which the dashed arrows are gluing mappings


These mappings glue together to give the projection $H_{a} \rightarrow \mathbb{P}^{1}$. Note that the inverse image of every point in $\mathbb{P}^{1}$ is a $\mathbb{P}^{1}$.

In general, suppose we have a $\mathbb{Z}$-linear mapping $\phi: N \rightarrow N^{\prime}$ of lattices and fans $\Delta$ and $\Delta^{\prime}$ in $N$ and $N^{\prime}$, respectively, having the property that for each cone $\sigma \in \Delta$, there is a cone $\sigma^{\prime} \in \Delta^{\prime}$ such that $\phi(\sigma) \subseteq \sigma^{\prime}$, then-just as for the example of the Hirzebruch surface, above - there is a natural mapping $X(\Delta) \rightarrow X\left(\Delta^{\prime}\right)$ induced by $\phi$. In the example of the Hirzebruch surface we used $\phi(x, y)=x$. Note that, since $a>0$, the mapping $(x, y) \mapsto y$ would not satisfy the condition.
13.2. Toric varieties from polytopes. A rational polytope $P$ is the convex hull over a finite set of vertices $X \subseteq \mathbb{Q}^{n} \subseteq \mathbb{R}^{n}$. We can also write $P$ as the bounded intersection of a finite collection of half spaces in $\mathbb{R}^{n}$, that is,

$$
P=\left\{x \in \mathbb{R}^{n} \mid A x \geq-b\right\}
$$

for some $A \in M_{n \times n}(\mathbb{Q})$ and $b \in \mathbb{Q}^{n}$. A (proper) face $F$ of $P$ of dimension $n-k$ is the intersection with $k$ supporting hyperplanes of the form

$$
\left\{v \in P \mid\left\langle a_{i}, v\right\rangle=-b_{i}\right\},
$$

where $a_{i}$ is the $i$-th row of $A$. A facet is a face of dimension $n-1$.
Given a rational polytope $P \in M_{\mathbb{R}} \cong \mathbb{R}^{n}$, define a fan $\Delta_{P}$ whose rays are the inward pointing normals to the facets of $P$. There is a cone $\sigma_{v}$ for each vertex $v$ of $P$, determined by the normals of the facets incident on $v$. A nice feature of this construction is the dual cone of $\sigma_{v}$ is formed by looking at $v \in P$ and extending its edges indefinitely (and translating to the origin).

Example 13.9. Consider the square $P$ with vertices $(0,0),(1,0),(2,1),(0,1)$ as shown in the left of Figure 32. We get $X\left(\Delta_{P}\right)=H_{1}$, the Hirzebruch surface, considered earlier, with $a=1$.

$\Delta_{P}$


Figure 32. The fan $\Delta_{P}$ associated to a polytope $P$. The cones of $\Delta_{P}$ are in bijection with vertices of $P$. The dual cones are determined by the corresponding vertex figures. For instance, the blue triangle at $v_{3}$ determines the dual cone $\sigma_{3}^{\vee}$

Exercise 13.10. Let $n=2$ and let $P$ be the convex polytope in $M_{\mathbb{R}}$ with vertices $(-1,1),(-1,-1),(2,-1)$. What is the corresponding toric variety $X\left(\Delta_{P}\right)$ ? Describe $U_{\sigma}$ for each $\sigma \in \Delta_{P}$ and the gluing maps between the two-dimensional cones of $\Delta_{P}$.
13.3. Cohomology of toric varieties. We can compute the cohomology of some nice toric varieties, namely those that are smooth and complete.

A cone $\sigma \subseteq N_{\mathbb{R}}$ is smooth if it is generated by a subset of a $\mathbb{Z}$-basis of $N$. A fan $\Delta \subseteq N_{\mathbb{R}}$ is complete if its support is all of $N_{\mathbb{R}}$, i.e., its fans cover $N_{\mathbb{R}}$. Suppose that $\Delta$ is a complete fan and that each cone $\sigma \in \Delta$ is smooth. In this case, it turns out that the corresponding toric variety $X(\Delta) \stackrel{\text { call }}{=} X$ is a compact orientable complex manifold.

Let $\Delta(1)$ denote the set of all 1-dimensional faces of $\Delta$. For each $D \in \Delta(1)$, let $n_{D}$ be the first lattice point along $D$. We define the Chow ring of $X$ to be the quotient ring graded by degree:

$$
A^{\bullet}(X):=\mathbb{Z}[D \in \Delta(1)] /(I+J)
$$

where

$$
\begin{aligned}
& I=\left(\prod_{D \in S} D \mid S \subseteq \Delta(1) \text { does not generated a cone in } \Delta\right) \\
& J=\left(\sum_{D \in \Delta(1)}\left\langle m, n_{D}\right\rangle D \mid m \in M\right)
\end{aligned}
$$

where $\left\langle m, n_{D}\right\rangle$ denotes the inner product.
The cohomology groups can be calculated through an extension of scalars:
Theorem 13.11. Let $\Delta$ and $X$ be as above. We have

$$
H^{k} X=\left\{\begin{array}{cl}
0 & \text { if } k \text { is odd } \\
A^{\ell}(X) \otimes_{\mathbb{Z}} \mathbb{R} & \text { if } k=2 \ell
\end{array}\right.
$$

(To form $A^{\bullet}(X) \otimes \mathbb{R}$, just replace $\mathbb{Z}$ by $\mathbb{R}$ in the definition. If $A^{\ell}(X)$ is spanned over $\mathbb{Z}$ by $t$ monomials, then $H^{2 \ell} X=\mathbb{R}^{t}$.)

Example 13.12. Let $\Delta$ be the fan in Example 13.6 whose associated toric variety is $X=\mathbb{P}^{2}$. Let $D_{1}=\mathbb{R}_{\geq 0} e_{1}, D_{2}=\mathbb{R}_{\geq 0} e_{2}, D_{3}=\mathbb{R}_{\geq 0}\left(-e_{1}-e_{2}\right)$. Then

$$
\begin{aligned}
I= & \left(D_{1} D_{2} D_{3}\right), \\
J= & \left(\sum_{i=1}^{3}\left\langle e_{j}, n_{D_{i}}\right\rangle D_{i} \mid j=1,2\right) \\
= & \left(\langle(1,0),(1,0)\rangle D_{1}+\langle(1,0),(0,1)\rangle D_{2}+\langle(1,0),(-1,-1)\rangle D_{3},\right. \\
& \left.\langle(0,1),(1,0)\rangle D_{1}+\langle(0,1),(0,1)\rangle D_{2}+\langle(0,1),(-1,-1)\rangle D_{3}\right) \\
= & \left(D_{1}-D_{3}, D_{2}-D_{3}\right) .
\end{aligned}
$$

We see that
$A \cdot\left(\mathbb{P}_{\mathbb{C}}^{2}\right)=\mathbb{Z}\left[D_{1}, D_{2}, D_{3}\right] /\left(D_{1} D_{2} D_{3}, D_{1}-D_{3}, D_{2}-D_{3}\right) \cong \mathbb{Z}\left[D_{3}\right] /\left(D_{3}^{3}\right)=\mathbb{Z}+\mathbb{Z} D_{3}+\mathbb{Z} D_{3}^{2}$.
Therefore, we have the table

$$
\begin{array}{c|ccccc}
k & 0 & 1 & 2 & 3 & 4 \\
\hline H^{k}\left(\mathbb{P}_{\mathbb{C}}^{2}\right) & \mathbb{R} & 0 & \mathbb{R} & 0 & \mathbb{R}
\end{array}
$$

Exercise 13.13. What are the cohomology groups of $\mathbb{P}^{n}$ ?
Example 13.14. Let $H_{a}$ be the Hirzebruch surface as described in Example 13.8. Let $D_{1}=\mathbb{R}_{\geq 0} e_{1}, D_{2}=\mathbb{R}_{\geq 0} e_{2}, D_{3}=\mathbb{R}_{\geq 0}\left(-e_{1}+a e_{2}\right), D_{4}=\mathbb{R}_{\geq 0}\left(-e_{2}\right)$. Then one can check that the two ideals are

$$
\begin{aligned}
& I=\left(D_{1} D_{3}, D_{2} D_{4}\right) \\
& J=\left(D_{1}-D_{3}, D_{2}+a D_{3}-D_{4}\right)
\end{aligned}
$$

Therefore, the Chow ring is

$$
\begin{aligned}
A^{\bullet}\left(H_{a}\right) & =\mathbb{Z}\left[D_{1}, D_{2}, D_{3}, D_{4}\right] /\left(D_{1} D_{3}, D_{2} D_{4}, D_{1}-D_{3}, D_{2}+a D_{3}-D_{4}\right) \\
& \cong \mathbb{Z}\left[D_{1}, D_{2}\right] /\left(D_{1}^{2}, D_{2}\left(D_{2}+a D_{1}\right)\right)
\end{aligned}
$$

$$
=\mathbb{Z}+\left(\mathbb{Z} D_{1}+\mathbb{Z} D_{2}\right)+\mathbb{Z} D_{1} D_{2}
$$

using the facts $D_{3}=D_{1}, D_{4}=D_{2}+a D_{3}$, and $D_{2}^{2}=-a D_{1} D_{2}$ in $A^{\bullet}\left(H_{a}\right)$. So we have the table

$$
\begin{array}{c|ccccc}
k & 0 & 1 & 2 & 3 & 4 \\
\hline H^{k}\left(H_{a}\right) & \mathbb{R} & 0 & \mathbb{R}^{2} & 0 & \mathbb{R}
\end{array}
$$

The symmetry in the above two examples is due to Poincaré duality.
13.4. Homogeneous coordinates. David Cox ([3]) introduced the homogeneous coordinates for a toric variety, generalizing those for projective space. In the case of simplicial toric varieties, we can then write the toric variety as a quotient of $\mathbb{C}^{n}$ modulo the action of a group, again generalizing the case of $\mathbb{P}^{n}=\mathbb{C}^{n+1} /(x \sim \lambda x$ : $\left.\lambda \in \mathbb{C}^{*}\right)$.

Let $\Delta$ be a fan in $N \cong \mathbb{Z}^{n}$, and let $\Delta(1)$ be the collection of all 1-dimensional faces (rays) of $\Delta$. For each $D \in \Delta(1)$, let $n_{D} \in N$ be the first lattice point along $D$. Let $X=X(\Delta)$ be the corresponding toric variety. Define the the codimension Chow group $A_{n-1}(X)$, (related to the Chow group $A^{\bullet}$, defined previously only for smooth complete toric varieties) to be the quotient

$$
A_{n-1}(X):=\mathbb{Z}^{\Delta(1)} /\left(\sum_{D \in \Delta(1)}\left\langle m, n_{D}\right\rangle D \mid m \in M\right)
$$

where $\mathbb{Z}^{\Delta(1)}=\left\{\sum_{D \in \Delta(1)} a_{D} D \mid a_{D} \in \mathbb{Z}\right\} \cong \mathbb{Z}^{|\Delta(1)|}$ is the free $\mathbb{Z}$-module generated by $\Delta(1)$. (Note that $A_{n-1}(X) \cong A^{1}(X)$ if $X$ is smooth and complete.) We have an injection of the lattice $M$ (in which the dual cones sit) via

$$
\begin{aligned}
M & \rightarrow \mathbb{Z}^{\Delta(1)} \\
m & \mapsto D_{m}:=\sum_{D \in \Delta(1)}\left\langle m, n_{D}\right\rangle D
\end{aligned}
$$

and a short exact sequence

$$
0 \rightarrow M \rightarrow \mathbb{Z}^{\Delta(1)} \rightarrow A_{n-1}(X) \rightarrow 0
$$

Let $S$ be the polynomial ring

$$
S:=\mathbb{C}\left[x_{D} \mid D \in \Delta(1)\right] .
$$

For each $a=\sum_{D \in \Delta(1)} a_{D} D \in \mathbb{Z}^{\Delta(1)}$ with $a_{D} \geq 0$ for all $D$, there is a corresponding monomial

$$
x^{a}:=\prod_{D \in \Delta(1)} x_{D}^{a_{D}} \in S
$$

and its degree, denoted $\operatorname{deg}\left(x^{a}\right)$ is defined to be its class $[a] \in A_{n-1}(X)$. Thus, the polynomial ring $S$ is graded by $A_{n-1}(X)$.

Definition 13.15. The homogeneous coordinate ring of $X$ is the polynomial ring $S$ graded by $A_{n-1}(X)$, meaning that if we write

$$
S_{\alpha}:=\bigoplus_{\operatorname{deg}\left(x^{a}\right)=\alpha} \mathbb{C} x^{a}
$$

then

$$
S \cong \bigoplus_{\alpha \in A_{n-1}(X)} S_{\alpha}
$$

and one can check that $S_{\alpha} \cdot S_{\beta} \subseteq S_{\alpha+\beta}$.
Example 13.16. The fan for $\mathbb{P}^{n}$ has $n+1$ rays: $D_{i}=\mathbb{R} e_{i}$ for $i=1, \ldots, n$, corresponding to the standard basis vectors, and $D_{n+1}=\mathbb{R}\left(-e_{1}-\cdots-e_{n}\right)$. To compute $A_{n-1} \mathbb{P}^{n}=A^{1} \mathbb{P}^{n}$, we first find

$$
\begin{aligned}
\left(\sum_{D \in \Delta(1)}\left\langle m, n_{D}\right\rangle D \mid m \in M\right) & =\left(\sum_{i=1}^{n+1}\left\langle e_{j}, n_{D_{i}}\right\rangle D_{i} \mid j=1, \ldots, n\right) \\
& =\left(D_{1}-D_{n+1}, D_{2}-D_{n+1}, \ldots, D_{n}-D_{n+1}\right)
\end{aligned}
$$

Therefore,

$$
\begin{aligned}
A_{n-1} \mathbb{P}^{n}=\mathbb{Z}^{\Delta(1)} /\left(D_{1}-D_{n+1}, D_{2}-D_{n+1}, \ldots, D_{n}-D_{n+1}\right)=\mathbb{Z} D_{n+1} & \cong \mathbb{Z} \\
D_{i} & \mapsto D_{n+1}
\end{aligned}>1 .
$$

Letting $x_{i}=x_{D_{i}}$ for $i=1, \ldots, n+1$, we have the homogeneous coordinate ring

$$
S=\mathbb{C}\left[x_{1}, \ldots, x_{n+1}\right] .
$$

The degree of a monomial $x_{1}^{a_{1}} \cdots x_{n+1}^{a_{n+1}}$ is the class of $a=\sum_{i=1}^{n+1} a_{i} D_{i}$ in $A_{n-1} \mathbb{P}^{n}$. But in $A_{n-1} \mathbb{P}^{n}$, we have

$$
\operatorname{deg}\left(x^{a}\right)=[a]=\left[\sum_{i=1}^{n+1} a_{i} D_{i}\right]=\left[\left(\sum_{i=1}^{n+1} a_{i}\right) D_{n+1}\right]
$$

Identifying $A_{n-1} \mathbb{P}^{n}$ with $\mathbb{Z}$, as above, we get

$$
\operatorname{deg} x^{a}=\sum_{i=1}^{n+1} a_{i}
$$

the usual degree.
Example 13.17. Let $H_{a}$ denote the Hirzebruch surface as described in Example 13.8. Recall that in Example 13.14 we computed

$$
\begin{aligned}
A_{n-1}\left(H_{a}\right)=\mathbb{Z}^{\Delta(1)} /\left\langle D_{1}-D_{3}, D_{2}+a D_{3}-D_{4}\right\rangle & \rightarrow \mathbb{Z}\left\{D_{1}, D_{2}\right\} \\
D_{1} & \mapsto D_{1} \\
D_{2} & \mapsto D_{2} \\
D_{3} & \mapsto D_{1} \\
D_{4} & \mapsto a D_{1}+D_{2}
\end{aligned}
$$

where $\mathbb{Z}\left\{D_{1}, D_{2}\right\}:=\operatorname{Span}_{\mathbb{Z}}\left\{D_{1}, D_{2}\right\}$. Then, finally

$$
\begin{aligned}
A_{n-1}\left(H_{a}\right) \cong & \mathbb{Z}\left\{D_{1}, D_{2}\right\}
\end{aligned} \begin{aligned}
& \mathbb{Z}^{2} \\
& \\
& \\
& a D_{1}+b D_{2} \mapsto(a, b)
\end{aligned}
$$

So the homogeneous coordinate ring for $H_{a}$ is $S=\mathbb{C}\left[x_{1}, x_{2}, x_{3}, x_{4}\right]$ and it is graded by $\mathbb{Z}^{2}$ with each indeterminate having degree

$$
\operatorname{deg}\left(x_{1}\right)=\operatorname{deg}\left(x_{3}\right)=(1,0), \quad \operatorname{deg}\left(x_{2}\right)=(0,1), \quad \operatorname{deg}\left(x_{4}\right)=\operatorname{deg}\left(x_{2} x_{3}^{a}\right)=(a, 1)
$$

For example, $\operatorname{deg}\left(x_{1} x_{2}^{2} x_{3}^{3} x_{4}\right)=(1,0)+2(0,1)+3(1,0)+(a, 1)=(4+a, 3)$. The monomial $x_{1}^{2} x_{2} x_{3}^{3} x_{4}$ has degree $(5+a, 2)$. The degrees of these monomials in $S$
differ even though they would have the same degree under the usual grading (in which each $x_{i}$ had degree 1 ).

Quotients. Now assume that $X$ is simplicial, meaning that each of its cones $\sigma$ has $\operatorname{dim}(\sigma)$ rays. (It suffices to check the maximal-dimensional cones.) Our goal is to view $X$ as a quotient space under some group action, generalizing the construction of projective space. Consider the following short exact sequence of $\mathbb{Z}$-modules:

$$
\begin{aligned}
0 \longrightarrow M & \longrightarrow \mathbb{Z}^{\Delta(1)} \longrightarrow A_{n-1}(X) \longrightarrow 0 \\
m & \longmapsto \sum_{D \in \Delta(1)}\left\langle m, n_{D}\right\rangle D .
\end{aligned}
$$

The nonzero complex numbers, $\mathbb{C}^{*}$, form an abelian group under multiplication, and hence has a $\mathbb{Z}$-module structure: for $a \in \mathbb{Z}$ and $z \in \mathbb{C}^{*}$, we have $a \cdot z:=z^{a}$. It then makes sense to consider the mappings $\operatorname{hom}_{\mathbb{Z}}\left(\mathbb{Z}^{\Delta(1)}, \mathbb{C}^{*}\right)$, i.e, the $\mathbb{Z}$-linear mappings from $\mathbb{Z}^{\Delta(1)}$ to $\mathbb{C}^{*}$. These are determined by the images of $D \in \Delta(1)$. For instance, if $\Delta(1)=\left\{D_{1}, \ldots, D_{k}\right\}$ and $g_{1}, \ldots, g_{k}$ are any elements of $\mathbb{C}^{*}$, there is a corresponding mapping

$$
a_{1} D_{1}+\cdots+a_{k} D_{k} \mapsto g_{1}^{a_{1}} \cdots g_{k}^{a_{k}} \in \mathbb{C}^{*}
$$

Since the choices for the $g_{i}$ are arbitrary and completely determine the mapping, we have

$$
\operatorname{hom}_{\mathbb{Z}}\left(\mathbb{Z}^{\Delta(1)}, \mathbb{C}^{*}\right) \cong\left(\mathbb{C}^{*}\right)^{\Delta(1)}
$$

Next, consider $\operatorname{hom}_{\mathbb{Z}}\left(A_{n-1}(X), \mathbb{C}^{*}\right)$. If $\Delta(1)=\left\{D_{1}, \ldots, D_{k}\right\}$ and $g_{1}, \ldots, g_{k} \in$ $\mathbb{C}^{*}$, we can still attempt to define a mapping $A_{n-1}(X) \rightarrow \mathbb{C}^{*}$ by sending $D_{i} \mapsto$ $g_{i}$, as above. However, since there are relations among the $D_{i}$ in $A_{n-1}(X)$, the corresponding relations must hold among the $g_{i}$. In other words, the choices for the $g_{i}$ are now constrained. For instance, if $D_{3}=2 D_{1}+D_{2}$, then since $D_{3} \mapsto g_{3}$ and $2 D_{1}+D_{2} \mapsto g_{1}^{2} g_{2}$, we require that $g_{3}=g_{1}^{2} g_{2}$, i.e. $g_{1}^{2} g_{2} g_{3}^{-1}=1$. In general, we define the group $G$ by
$G:=\operatorname{hom}_{\mathbb{Z}}\left(A_{n-1}(X), \mathbb{C}^{*}\right) \cong\left\{g \in\left(\mathbb{C}^{*}\right)^{\Delta(1)} \mid \prod_{D \in \Delta(1)} g_{D}^{\left\langle m, n_{D}\right\rangle}=1\right.$ for all $\left.m \in M\right\}$.
The conditions $\prod_{D \in \Delta(1)} g_{D}^{\left\langle m, n_{D}\right\rangle}=1$, encode all the required relations. It suffices to let $m$ range over a basis for $M$. So if $M=\mathbb{Z}^{n}$, the relations are $\prod_{D \in \Delta(1)} g_{D}^{\left\langle e_{i}, n_{D}\right\rangle}=1$ for $i=1, \ldots, n$.

The inclusion

$$
G \subseteq\left(\mathbb{C}^{*}\right)^{\Delta(1)} \subseteq \mathbb{C}^{\Delta(1)}
$$

gives a natural action of $G$ on $\mathbb{C}^{\Delta(1)}$ :

$$
g \cdot x:=\left(g_{D} x_{D}\right)_{D \in \Delta(1)},
$$

where $g \in G, x \in \mathbb{C}^{\Delta(1)}$.
For a face $\sigma \in \Delta$, let $\sigma(1):=\Delta(1) \cap \sigma$ denote the rays in $\sigma$ and define the monomial

$$
x^{\hat{\sigma}}:=\prod_{D \notin \sigma(1)} x_{D} \in S .
$$

Let $B$ be the monomial ideal

$$
B:=\left(x^{\hat{\sigma}} \mid \sigma \in \Delta\right)=\left(x^{\hat{\sigma}} \mid \sigma \text { is a maximal cone in } \Delta\right) \subseteq S
$$

(note that the generators of $B$ encode the structure of $\Delta$ ) and let $Z$ be the zero set of $B$,

$$
Z:=\left\{x \in \mathbb{C}^{\Delta(1)} \mid x^{\hat{\sigma}}=0 \text { for all } \sigma \in \Delta\right\}
$$

Theorem 13.18. Let $X=X(\Delta)$ be a simplicial toric variety. Then,
(1) $\mathbb{C}^{\Delta(1)} \backslash Z$ is invariant under the action by $G$,
(2) The toric variety $X$ is the quotient of $\mathbb{C}^{\Delta(1)} \backslash Z$ by the action of $G$, i.e.,

$$
X \cong\left(\mathbb{C}^{\Delta(1)} \backslash Z\right) /(x \sim g \cdot x \mid g \in G)
$$

An element in this quotient is called the homogeneous coordinates of a point in $X$.
Example 13.19. Let $X=\mathbb{P}^{n}$. The maximal cones of the fan in this case consists of all subsets of size $n$ from the vectors $e_{1}, \ldots, e_{n},-e_{1}-\cdots-e_{n}$. If $\sigma$ is one of these cones, then it omits exactly one of these vectors. Hence,

$$
B=\left(x_{1}, \ldots, x_{n+1}\right)
$$

and $Z=\{0\} \subset \mathbb{C}^{n+1}$. Since $A_{n-1} \mathbb{P}^{n} \cong \mathbb{Z}$, we have $G=\operatorname{hom}\left(A_{n-1}\left(\mathbb{P}^{n}\right), \mathbb{C}, C^{*}\right) \cong$ $\mathbb{C}^{*}$. In detail, since $A_{n-1}\left(\mathbb{P}^{n}\right)$ is the span of the $D_{i}$ modulo the relations $D_{i}-D_{n+1}$ for $i=1, . . n$, we have

$$
\begin{aligned}
G & =\left\{g \in \mathbb{C}^{n+1} \mid g_{i} g_{n+1}^{-1}=1 \text { for } i=1, \ldots, n\right\} \\
& =\left\{g \in \mathbb{C}^{n+1} \mid g_{i}=g_{n+1} \text { for } i=1, \ldots, n\right\} \\
& \cong \mathbb{C}^{*}
\end{aligned}
$$

where the isomorphism in the final step is given by $g \mapsto g_{n+1}$. Using this isomophism, the group action on $\mathbb{C}^{\Delta(1)}=\mathbb{C}^{n+1}$ is given by

$$
\lambda \cdot\left(x_{1}, \ldots, x_{n+1}\right)=\left(\lambda x_{1}, \ldots, \lambda x_{n+1}\right)
$$

where $\lambda=g_{n+1} \in \mathbb{C}^{*}$.
As claimed in Theorem 13.18 , the set $\mathbb{C}^{\Delta(1)} \backslash Z$, i.e., $\mathbb{C}^{n+1} \backslash\{0\}$ is invariant under the action by $G$ : if $\left(x_{1}, \ldots, x_{n+1}\right) \neq 0$, then $\lambda\left(x_{1}, \ldots, x_{n+1}\right) \neq 0$. Further, $\mathbb{P}^{n}$ is the quotient

$$
\mathbb{P}^{n} \cong\left(\mathbb{C}^{n+1} \backslash\{0\}\right) /\left(x \sim \lambda x \mid \lambda \in \mathbb{C}^{*}\right)
$$

Example 13.20. Now consider the Hirzebruch surface $H_{a}$. We have previously computed

$$
A_{n-1}\left(H_{a}\right)=\mathbb{Z}^{\Delta(1)} /\left\langle D_{1}-D_{3}, D_{2}+a D_{3}-D_{4}\right\rangle \cong \mathbb{Z}^{2}
$$

From Figure 31, we see

$$
B=\left\langle x_{3} x_{4}, x_{2} x_{3}, x_{1} x_{4}, x_{1} x_{2}\right\rangle
$$

Setting all of these generating monomials equal to zero and solving, we find

$$
Z=\left\{x \in \mathbb{C}^{4} \mid x_{1}=x_{3}=0 \text { or } x_{2}=x_{4}=0\right\}
$$

The relations for $A_{n-1}$ are generated by $D_{1}-D_{3}=0$ and $D_{4}=D_{2}+a D_{3}$. Using the first relation, the second can be rewritten as $D_{4}=a D_{1}+D_{2}$. Rewriting these multiplicatively gives

$$
\begin{aligned}
G \cong\left\{g \in\left(\mathbb{C}^{*}\right)^{4} \mid g_{1}=g_{3}, g_{4}=g_{1}^{a} g_{2}\right\} & \cong\left(\mathbb{C}^{*}\right)^{2} \\
g & \mapsto\left(g_{1}, g_{2}\right) \\
\left(g_{1}, g_{2}, g_{1}, g_{1}^{a} g_{2}\right) & \leftrightarrow\left(g_{1}, g_{2}\right)
\end{aligned}
$$

Therefore, the quotient description of the Hirzebruch surface is

$$
H_{a} \cong\left(\mathbb{C}^{4} \backslash Z\right) /\left(x \sim\left(\lambda x_{1}, \mu x_{2}, \lambda x_{3}, \lambda^{a} \mu x_{4}\right)\right)
$$

where $\lambda, \mu \in \mathbb{C}^{*}$.
13.5. Mapping toric varieties into projective spaces. In this section, we will assume that $X$ is a smooth, complete toric variety associated with the fan $\Delta$. Let $\Delta(1)=\left\{D_{1}, \ldots, D_{\ell}\right\}$ be the rays of $\Delta$, and let $S=\mathbb{C}\left[x_{1}, \ldots, x_{\ell}\right]$ be the homogeneous coordinate ring of $X$ with $x_{i}$ corresponding to $D_{i}$ and graded by $A_{n-1} X$. Recall the short exact sequence

$$
\begin{aligned}
0 \longrightarrow M & \longrightarrow \mathbb{Z}^{\Delta(1)} \longrightarrow A_{n-1}(X) \longrightarrow 0 \\
m & D_{m}
\end{aligned}
$$

where

$$
\begin{equation*}
D_{m}:=\sum_{D \in \Delta(1)}\left\langle m, n_{D}\right\rangle D \tag{10}
\end{equation*}
$$

An element of $\mathbb{Z}^{\Delta(1)}$ is called a divisor of $X$. If all of the coefficients of a divisor are nonnegative, then the divisor is said to be effective. Consider an effective divisor

$$
E=\sum_{D \in \Delta(1)} a_{D} D \in \mathbb{Z}^{\Delta(1)}
$$

and associate to it the polytope

$$
\begin{equation*}
P(E):=\left\{m \in M_{\mathbb{R}} \mid\left\langle m, n_{D}\right\rangle \geq-a_{D} \text { for all } D \in \Delta(1)\right\} \tag{11}
\end{equation*}
$$

where we recall that $n_{D}$ is the first lattice point along $D$. We also make the following assumption: $X=X\left(\Delta_{P}(E)\right)$.

Let $T=\left\{m_{1}, \ldots, m_{t+1}\right\} \subseteq P(E) \cap M$ be a collection of lattice points in $P(E)$ containing all of its vertices. Using the homogeneous coordinates of $X$ as described in Theorem 13.18, we get a mapping of $X$ into projective space as follows:

$$
\begin{align*}
\phi_{T}: X & \longrightarrow \mathbb{P}^{t}  \tag{12}\\
x & \longmapsto\left(x^{D_{m_{1}}+E}, \ldots, x^{D_{m_{t+1}+E}}\right),
\end{align*}
$$

with $D_{m_{i}}$ defined by Equation 10. The mapping $\phi_{T}$ is well-defined: Including the vertices of $P(E)$ in $T$ assures there is no point $p \in X$ such that $\phi(T)=0$. Further, since $\left[D_{m_{i}}\right]=[0] \in A_{n-1}(X)$ for all $i$, we have $\left[D_{m_{i}}+E\right]=[E]$, and the mapping is homogeneous of degree $[E]$. Therefore, scaling the homogeneous coordinates in the domain will scale each component of its corresponding point in the codomain.

The mapping $\phi_{T}$ will be an embedding if for each vertex $v$ of $P(E)$, as you travel along each each emanating from $v$, the first lattice point you reach is an element of $T$.

Example 13.21. Let $X=\mathbb{P}^{n}$ with divisors $D_{i}=\mathbb{R}_{\geq 0} e_{i}$ for $i=1, \ldots, n$ and $D_{n+1}=$ $\mathbb{R}_{\geq 0}\left(-e_{1}-\cdots-e_{n}\right)$, as usual. Consider the effective divisor $E=d D_{n+1}$ for some positive integer $d$. We have

$$
\begin{aligned}
P(E) & =\left\{m \in M_{\mathbb{R}} \mid\left\langle m, e_{i}\right\rangle \geq 0 \text { for } i=1, \ldots, n \text { and }\left\langle m,-\sum_{i=1}^{n} e_{i}\right\rangle \geq-d\right\} \\
& =\left\{m \in M_{\mathbb{R}} \mid m_{i} \geq 0 \text { for } i=1, \ldots, n \text { and } \sum_{i=1}^{n} m_{i} \leq d\right\}
\end{aligned}
$$

i.e., $P(E)$ is the simplex with vertices $0, d e_{1}, \ldots, d e_{n}$. When $T=P(E) \cap \mathbb{Z}^{n}$ contains all the lattice points in $P(E)$, we get the $d$-uple embedding $\mathbb{P}^{n} \rightarrow \mathbb{P}^{\binom{n+d}{d}-1}$ :

$$
\begin{aligned}
\mathbb{P}^{n} & \longrightarrow \mathbb{P}^{\binom{n+d}{d}-1} \\
\left(x_{0}, \ldots, x_{n}\right) & \longmapsto(\underbrace{x_{1}^{d}, x_{1}^{d-1} x_{2}, \ldots, x_{n+1}^{d}}_{\text {all monomials of degree } d}) .
\end{aligned}
$$

Example 13.22. Let $H_{2}$ denote the Hirzebruch surface having its four rays generated by $e_{1}, e_{2},-e_{1}+2 e_{2},-e_{2}$ :


Consider the effective divisor $E=2 D_{3}+3 D_{4}$. Taking the dot products of $m=$ $\left(m_{1}, m_{2}\right) \in M_{\mathbb{R}}=\mathbb{R}^{2}$ with the first lattice points along each $D_{i}$ and using Equation 11 gives the inequalities defining the corresponding polytope:

$$
P(E)=\left\{m \in M_{\mathbb{R}} \mid m_{1}, m_{2} \geq 0,-m_{1}+2 m_{2} \geq-2,-m_{2} \geq-3\right\}
$$

which is drawn in Figure 33. Consider the following set of lattice points in $P(E)$.

$$
T=\{(0,0),(0,1),(0,2),(1,2),(0,3),(3,1),(8,3)\}
$$

Using Equation 10 we compute the corresponding divisors necessary for our mapping (as prescribed by Equation 12):

$$
\begin{aligned}
D_{(0,0)}+E=2 D_{3}+3 D_{4}, & D_{(0,1)}+E=D_{2}+4 D_{3}+2 D_{4}, \\
D_{(0,2)}+E=2 D_{2}+6 D_{3}+D_{4} & D_{(1,2)}+E=D_{1}+2 D_{2}+5 D_{3}+D_{4}, \\
D_{(0,3)}+E=3 D_{3}+8 D_{4}, & D_{(3,1)}+E=3 D_{1}+D_{2}+D_{3}+2 D_{4} \\
D_{(8,3)}+E=8 D_{1}+3 D_{2} . &
\end{aligned}
$$



Figure 33. Polytope associated to the divisor $E=2 D_{3}+3 D_{4}$ on the Hirzebruch surface $\mathrm{H}_{2}$. We consider the embedding of $\mathrm{H}_{2}$ into $\mathbb{P}^{6}$ determined by the circled lattice points.

So the map into projective space is

$$
\begin{aligned}
\phi_{T}: H_{2} & \longrightarrow \mathbb{P}^{6} \\
\left(x_{1}, x_{2}, x_{3}, x_{4}\right) & \longmapsto\left(x_{3}^{2} x_{4}^{3}, x_{2} x_{3}^{4} x_{4}^{2}, x_{2}^{2} x_{3}^{6} x_{4}, x_{1} x_{2}^{2} x_{3}^{5} x_{4}, x_{3}^{3} x_{4}^{8}, x_{1}^{3} x_{2} x_{3} x_{4}^{2}, x_{1}^{8} x_{2}^{3}\right)
\end{aligned}
$$

To appreciate that $\phi_{T}$ respects scaling, recall that scaling, in the case of $H_{2}$ is by the group $\left(\mathbb{C}^{*}\right)^{2}$ and the group action is given by

$$
(\lambda, \mu) \cdot\left(x_{1}, x_{2}, x_{3}, x_{4}\right)=\left(\lambda x_{1}, \mu x_{2}, \lambda x_{3}, \lambda^{2} \mu x_{4}\right)
$$

where $(\lambda, \mu) \in\left(\mathbb{C}^{*}\right)^{2}$. We then check

$$
\phi_{T}\left(\lambda x_{1}, \mu x_{2}, \lambda x_{3}, \lambda^{2} \mu x_{4}\right)=\lambda^{8} \mu^{3} \phi_{T}\left(x_{1}, x_{2}, x_{3}, x_{4}\right)=\phi_{T}\left(x_{1}, x_{2}, x_{3}, x_{4}\right) \in \mathbb{P}^{6}
$$

There is a way to read off the monomial components in $\phi_{T}$ by computing the "lattice distances" of the corresponding lattice point from each of the facets of $P(E)$. As an example, consider the lattice point $(1,2) \in T$. Its corresponding component function in $\phi(T)$ is $x_{1} x_{2}^{2} x_{3}^{5} x_{4}$. We would like to see that its exponent vector, $(1,2,5,1)$ is the list of lattice distances from the facets of $P(E)$. First note that each facet of $P(E)$ corresponds to an indeterminate $x_{i}$ : the inward pointing normal of a facet determines a ray $D_{i}$ which determines $x_{i}$. So we label the facets accordingly:
$x_{4}$


To get to the lattice point $(1,2)$, we need to slide the $x_{1}$-facet over one, the $x_{2}$-facet up two, and the $x_{4}$-facet down one. That accounts for the first, second, and fourth components of the exponent vector $(1,2,5,1)$. The following diagram illustrates that the $x_{3}$-facet needs to be slid in its inward normal direction a total of five steps to reach $(1,2)$ :


The reader should check that, for instance, monomial $x_{1}^{3} x_{2} x_{3} x_{4}^{2}$, corresponding to the lattice point $(3,1)$, can be computed similarly.

In order to create an embedding of $H_{2}$ using the divisor $E=2 D_{3}+3 D_{4}$, one would needs to include not only the vertices in $T$, but also the first lattice points as you travel away from each vertex along edges incident to the vertex. In other words, $T$ must include at least the vertices pictured in Figure 34.


Figure 34. Polytope associated to the divisor $E=2 D_{3}+3 D_{4}$ on the Hirzebruch surface $H_{2}$. The circled lattice points are required to produce an embedding into projective space.

## 14. Grassmannians

Definition 14.1. If $V$ is a vector space over $\mathbb{k}$, then projective space on $V$, denoted $\mathbb{P}(V)$, is the collection of all one-dimensional subspaces of $V$. A special case is $\mathbb{P}_{\mathbb{k}}^{n}:=\mathbb{P}\left(\mathbb{k}^{n+1}\right)$ (or, usually, $\mathbb{P}^{n}$ when $\mathbb{k}$ is clear from context). An $r$-plane in $\mathbb{P}(V)$ is an $(r+1)$-dimensional subspace of $V$.

Proposition 14.2. Suppose $L$ is an $r$-plane, $M$ is an s-plane, and $L \cap M$ is a $k$ plane in $\mathbb{P}^{n}$. Then $k \geq r+s-n$.

Proof. Exercise.
Duality. An $(n-1)$-plane in $\mathbb{P}^{n}$ is called a hyperplane. It is the solution set to a linear equation

$$
H_{a}:=\left(a_{1}, \ldots, a_{n+1}\right) \cdot\left(x_{1}, \ldots, x_{n+1}\right)=a_{1} x_{1}+\cdots+a_{n+1} x_{n+1}=0
$$

for some $a:=\left(a_{1}, \ldots, a_{n+1}\right) \neq(0, \ldots, 0) \in \mathbb{k}^{n+1}$. Thus, $a$ is the normal vector to the hyperplane (in the case $\mathbb{k}=\mathbb{R}$ or $\mathbb{C}$ ). Further, $a$ is determined by the hyperplane up to scaling by a nonzero element of $\mathbb{k}$. Define $\left(\mathbb{P}^{n}\right)^{*}$ to be the dual projective space whose points are the hyperplanes in $\mathbb{P}^{n}$. Then we get a well-defined bijection between $\mathbb{P}^{n}$ and its dual:

$$
\begin{aligned}
\mathbb{P}^{n} & \rightarrow\left(\mathbb{P}^{n}\right)^{*} \\
a & \mapsto H_{a} .
\end{aligned}
$$

Definition 14.3. Let $V$ be a vector space over $\mathbb{k}$. The collection of $(r+1)$ dimensional subspaces of $V$ is called a Grassmannian and denoted by $G(r+1, V)$. It is also called the Grassmannian of $r$-planes in $\mathbb{P}(V)$ and then denoted by $\mathbb{G}_{r} \mathbb{P}(V)$. If $V=\mathbb{k}^{n+1}$, the Grassmannian is denoted by either $G(r+1, n+1)$ or $\mathbb{G}_{r} \mathbb{P}^{n}$. (We say $\mathbb{G}_{r} \mathbb{P}^{n}$ is a moduli space space for the set of $r$-planes in $\mathbb{P}^{n}$ since it parametrizes this set.)

Example 14.4. Some special cases of Grassmannians:

$$
\mathbb{G}(1, n+1)=\mathbb{G}_{0} \mathbb{P}^{n}=\mathbb{P}^{n}, \quad G(2,3)=\mathbb{G}_{1} \mathbb{P}^{2}=\left(\mathbb{P}^{2}\right)^{*}, \quad \mathbb{G}_{n-1} \mathbb{P}^{n}=\left(\mathbb{P}^{n}\right)^{*}
$$

The Grassmannian $G(2,4)=\mathbb{G}_{1} \mathbb{P}^{3}$ is the set of lines in three-space.
14.1. Manifold structure. We now fix our vector space $\mathbb{k}$ to be $\mathbb{C}$ (although a lot of what we do below will carry over $\mathbb{k}=\mathbb{R}$ or to arbitrary fields). Given $L \in G(r, n)=\mathbb{G}_{r-1} \mathbb{P}^{n-1}$, we can write

$$
L=\operatorname{Span}\left\{a_{1}, \ldots, a_{r}\right\}
$$

for some vectors $a_{1}, \ldots, a_{r} \in \mathbb{C}^{n}$. Let $A$ be the matrix whose rows are $a_{1}, \ldots, a_{r}$. Then $L=\operatorname{Span}\left\{b_{1}, \ldots, b_{r}\right\}$ for some other vectors $b_{i}$ if and only if there is an invertible $r \times r$ matrix $M$ such that $B=M A$, where $B$ is the matrix whose rows are the $b_{i}$. (Multiplying $A$ by $M$ on the left performs invertible row operations and, thus, does not change the rowspan.) Therefore,

$$
G(r, n)=\{r \times n \text { rank } r \text { matrices }\} /(A \sim M A: M r \times r, \text { invertible })
$$

Identifying the set of $r \times n$ matrices with $\mathbb{C}^{r \times n}=\mathbb{C}^{r n}$ induces a topology on the set of $r \times n$ matrices and the quotient topology on $G(r, n)$. (So a subset of $G(r, n)$ is open if and only if the set of all matrices representing points in that set forms an open subset of $\mathbb{C}^{r \times n}$.) The case $r=1$ recovers the usual construction of projective space $-A$ will be $1 \times n$ and $M=[\lambda]$ for some nonzero $\lambda \in \mathbb{C}$. In general, we consider an $r \times n$ matrix of rank $r$ to be the homogeneous coordinates for a point in $G(r, n)$.

We now seek an open covering of $G(r, n)$ and chart mappings generalizing those for projective space. We motivate the idea with an example:

Example 14.5. Consider

$$
L=\left(\begin{array}{llll}
1 & 0 & 3 & 1 \\
2 & 4 & 3 & 1
\end{array}\right) \in G(2,4)=\mathbb{G}_{1} \mathbb{P}^{3}
$$

a line in $\mathbb{P}^{3}$. With respect to the chart $\left(U_{4}, \phi_{4}\right)$ on $\mathbb{P}^{3}$, one may check that the line is parametrized by

$$
t \mapsto(1,0,3)+t((2,4,3)-(1,0,3)),
$$

i.e., the line containing the points $(1,0,3)$ and $(2,4,3)$. To find what we will soon define to be coordinates for $L$ with respect to a particular chart for $G(2,4)$, choose any two linearly independent columns and perform row operations to reduce that pair of columns to the identity matrix. Choosing the first and fourth columns gives:

$$
\begin{gathered}
\left(\begin{array}{cccc}
{ }^{\star} & 0 & 3 & { }^{\star} \\
2 & 4 & 3 & 1
\end{array}\right) \rightarrow\left(\begin{array}{rrrr}
\star & & \star \\
1 & 0 & 3 & 1 \\
0 & 4 & -3 & -1
\end{array}\right) \rightarrow\left(\begin{array}{rrrr}
\star & \\
1 & 4 & 0 & 0 \\
0 & 4 & -3 & -1
\end{array}\right) \rightarrow\left(\begin{array}{rrrc}
\star \\
1 & 4 & 0 & 0 \\
0 & -4 & 3 & 1
\end{array}\right) \\
L
\end{gathered}
$$

Performing the same row operations to the identity matrix $I_{2}$ yields the matrix

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

and $M L=L^{\prime}$. Since the operations are invertible, all matrices row equivalent to $L^{\prime}$ are equivalent to each other, i.e., represent the same point in $G(2,4)$. Knowing that the columns 1 and 4 have been fixed, the point with homogeneous coordinates $L$ can then be uniquely represented by the entries in columns 2 and 3 . If we then agree to read the entries of a matrix from left-to-right and top-to-bottom, we assign the unique point $(4,0,-4,3) \in \mathbb{C}^{4}$ to $L$. These are the coordinates of $L$ with respect to columns 1 and 4. Every $2 \times 4$ matrix whose first and fourth columns are linearly independent will have similar coordinates. Also, every point $x \in \mathbb{C}^{4}$ has a corresponding point in $\mathbb{G}_{1} \mathbb{P}^{3}$ represented by a matrix whose first and fourth columns form $I_{2}$ and whose second and third columns are given by the coordinates of $x$. The result is a chart, which we will call $\left(U_{1,4}, \phi_{1,4}\right)$.

We generalize the above example to create an atlas for $G(r, n)$. Let $j \in \mathbb{Z}^{r}$ with $1 \leq j_{1}<\cdots<j_{r} \leq n$. Given an $r \times n$ matrix $L$, let $L_{j}$ be the square $r \times r$
submatrix of $L_{j}$ formed by the columns with indices $j_{1}, \ldots, j_{r}$. Then define

$$
U_{j}:=\left\{L \in G(r, n) \mid \operatorname{rk} L_{j}=r\right\}
$$

Note that the rank of $L \in G(r, n)$ is independent of the choice of representative matrix. Define the corresponding chart mapping by

$$
\begin{align*}
\phi_{j}: U_{j} & \rightarrow \mathbb{C}^{r(n-1)}  \tag{13}\\
L & \mapsto \operatorname{flatten}_{j}\left(L_{j}^{-1} L\right),
\end{align*}
$$

where flatten ${ }_{j}$ creates a point in $\mathbb{C}^{r(n-r)}$ from an $r \times r$ matrix by reading off the entries in the columns not indexed by $j$ from left-to-right and top-to-bottom. Multiplying $L$ by $L_{j}^{-1}$ on the left performs row operations on $L$ so that the resulting matrix has $I_{r}$ as the submatrix indexed by $j$. To get the coordinates with respect to $\phi_{j}$, we then just read off the other entries in the matrix.

Proposition 14.6. Using the notation defined above, $\left\{\left(U_{j}, \phi_{j}\right)\right\}_{j}$ is an atlas for $G(r, n)$ :
(1) each $U_{j}$ is open in $G(r, n)$;
(2) the $U_{j}$ cover $G(r, n)$; and
(3) $\phi_{j}: U_{j} \rightarrow \mathbb{C}^{r(n-r)}$ is a homeomorphism.

Proof. Let $\mathbb{C}^{a \times b}$ denote that set of $a \times b$ matrices. Fix the isomorphism $\mathbb{C}^{a \times b} \cong$ $\mathbb{C}^{a b}$ by reading the entries of a matrix from left-to-right and top-to-bottom. The topology on $\mathbb{C}^{a \times b}$ is determined by insisting the isomophism is a homeomorphism. For each $j$, let $\pi_{j}$ be the projection mapping that sends an $r \times n$ matrix $L$ to the $r \times r$ submatrix $L_{j}$. Then we have a sequence of continuous mappings

$$
\begin{array}{ll}
\mathbb{C}^{r \times n} & \xrightarrow{\pi_{j}} \mathbb{C}^{r \times r} \\
L & \xrightarrow{\text { det }} \mathbb{C} \\
L & \mapsto L_{r}
\end{array}>\operatorname{det}\left(L_{r}\right) .
$$

The composition is continous, so the inverse image of $0 \in \mathbb{C}$ is closed in $\mathbb{C}^{r \times n}$. Call the complement of this set $\tilde{U}_{j}$. Then $\tilde{U}_{j}$ open, and the quotient of $\tilde{U}_{j}$ modulo our equivalence $A \sim M A$ defining the Grassmannian is $U_{j}$. Since we have given the Grassmannian the quotient topology, we conclude that $U_{j}$ is open.

A point in $G(r, n)$ is represented by an $r \times n$ matrix $L$ of rank $r$. Row reducing $L$ must then produce $I_{r}$ as a square submatrix. Let $j$ be the indices of the columns of that submatrix in $L$. Then $L \in U_{j}$.

We now consider a chart mapping $\phi_{j}: U_{j} \rightarrow \mathbb{C}^{r(n-r)}$. First note that $L \mapsto L_{j}^{-1}$ is a continuous function of $L$ : use the formula for the inverse using cofactors to see that the inverse is a rational function of the entries of the matrix. Then, since matrix multiplication and projections are continuous, it follows that $\phi_{j}$ is continuous. The inverse is the continuous mapping that sends a point in $p \in \mathbb{C}^{r(n-r)}$ to the matrix $L$ such that $L_{j}=I_{r}$ and whose other entries are obtained from $p$.

Definition 14.7. The standard atlas for $G(r, n)$ is $\left\{\left(U_{j}, \phi_{j}\right)\right\}_{j}$ where $j=\left(j_{1}, \ldots, j_{r}\right)$ with $1 \leq j_{1}<j_{2}<\cdots<j_{r} \leq n$, and $\phi_{j}$ is defined in Equation 13.

Example 14.8. If $L \in U_{1,2,4} \subset G(3,7)$, then it has a representative of the form

$$
\left(\begin{array}{lllllll}
1 & 0 & * & 0 & * & * & * \\
0 & 1 & * & 0 & * & * & * \\
0 & 0 & * & 1 & * & * & *
\end{array}\right)
$$

The $3(7-3)=12$ entries denoted by asterisks give the coordinates of the point via the chart $\left(U_{1,2,4}, \phi_{1,2,4}\right)$.

Exercise 14.9. The set of lines in three-space is $\mathbb{G}_{1} \mathbb{P}^{3}$, and $\operatorname{dim} \mathbb{G}_{1} \mathbb{P}^{3}=4$. Now, a line is determined by a point and a direction. So that sounds like six parameters should be necessary. Give an intuitive explanation for the fact that the set of lines in three-space should be four-dimensional.

### 14.2. The Plücker embedding.

Definition 14.10. Fix an ordering of the $\binom{n}{r}$ choices of column indices: $j: 1 \leq$ $j_{1}<\cdots<j_{r} \leq n$. The Plücker embedding is defined by

$$
\begin{aligned}
\Lambda: G(r, n) & \rightarrow \mathbb{P}\binom{n}{r}-1 \\
L & \mapsto\left(\operatorname{det}\left(L_{j}\right)\right)_{j}
\end{aligned}
$$

where $L_{j}$ is the submatrix of $L$ with columns indexed by $j$. The components of $\Lambda(L)$ are the Plücker coordinates of $L$. The $j$-th Plücker coordinate of $L$ is $\operatorname{det}\left(L_{j}\right)$.

Example 14.11. The Plücker embedding of $G(2,4)$ is

$$
\begin{gathered}
\Lambda: G(2,4) \rightarrow \mathbb{P}^{5} \\
\left(\begin{array}{cccc}
a_{1} & a_{2} & a_{3} & a_{4} \\
b_{1} & b_{2} & b_{3} & b_{4}
\end{array}\right) \mapsto\left(a_{1} b_{2}-a_{2} b_{1}, a_{1} b_{3}-a_{3} b_{1}, a_{1} b_{4}-a_{4} b_{1}, a_{2} b_{3}-a_{3} b_{2}, a_{2} b_{4}-a_{4} b_{2}, a_{3} b_{4}-a_{4} b_{3}\right)
\end{gathered}
$$

Proposition 14.12. The Plücker coordinates of $L \in G(r, n)$ are well-defined.
Proof. Since (any representative of) $L \in G(r, n)$ has rank $r$, it follows that $\operatorname{det}\left(L_{j}\right) \neq$ 0 for some $j$. Next, suppose that $A$ and $B$ are $r \times n$ matrices both representing $L$. Then there exists an $r \times r$ matrix $M$ such that $B=M A$. Therefore, for each choice $j$ of $r$ columns, we have $B_{j}=M A_{j}$. Hence, $\operatorname{det} B_{j}=\operatorname{det} M \operatorname{det} A_{j}$ for all $j$. So $\left(B_{j}\right)_{j}=\lambda\left(A_{j}\right)_{j}$ where $\lambda=\operatorname{det} M \neq 0$, and $\Lambda(A)=\Lambda(B) \in \mathbb{P}^{\binom{n}{r}-1}$.

Coordinate-free description. Let $G(r, V)$ denote the set of $r$-dimensional subspaces of a vector space $V$. Define the Plücker embedding of $G(r, V)$ by

$$
\begin{aligned}
G(r, V) & \rightarrow \mathbb{P}\left(\Lambda^{r} V\right) \\
W & \mapsto \Lambda^{r} W
\end{aligned}
$$

Picking a basis for $V$ recovers the Plücker embedding defined previously.
The Plücker relations. We now seek the equations defining the image of the Plücker embedding. Denote the coordinates of a point in $\mathbb{P}\binom{n}{r}$ by

$$
x(j)=x\left(j_{1}, \ldots, j_{r}\right)
$$

for each $j: 1 \leq j_{1}<\cdots<j_{r} \leq n$. For notational convenience, we adopt the following conventions to allow permutations of $j$ and repetitions of the indices:

$$
\begin{array}{rll}
x(\sigma(j)):= & \operatorname{sign}(\sigma) x(j) & \text { for each } \sigma \in \mathfrak{S}_{r}, \text { and } \\
x(j):= & 0 & \text { if } j_{s}=j_{t} \text { for some } s \neq t
\end{array}
$$

Definition 14.13 (Plücker relations). For each $I: 1 \leq i_{1}<\cdots<i_{r-1} \leq n$ and $J: 1 \leq j_{1}<\cdots<j_{r+1} \leq n$, the $I$, $J$-th Plücker relation is

$$
P_{I, J}:=\sum_{k=1}^{r+1}(-1)^{k} x\left(i_{1}, \ldots, i_{r-1}, j_{k}\right) x\left(j_{1}, \ldots, \widehat{j_{k}}, \ldots, j_{r+1}\right)
$$

Example 14.14. For $G(2,4)$ with $I=\{1\}$ and $J=\{2,3,4\}$,

$$
P_{I, J}=-x(1,2) x(3,4)+x(1,3) x(2,4)-x(1,4) x(2,3)
$$

In this case, all other choices for $I$ and $J$ yield this same relation up to sign. Therefore, the image of this Grassmannian under its Plücker embedding is

$$
\Lambda(G(2,4))=\left\{x \in \mathbb{P}^{5} \mid P_{\{1\},\{2,3,4\}}=0\right\}
$$

a quadric hypersurface in $\mathbb{P}^{5}$. The reader may want to check that the coordinates displayed in Example 14.11 satisfy this relation.
Theorem 14.15. The Plücker embedding $\Lambda: G(r, n) \rightarrow \mathbb{P}^{\binom{n}{r}}$ is injective with image equal to the zero-set of $\left\{P_{I, J}\right\}_{I, J}$.

Proof. We will just show that the image satisfies $P_{I, J}=0$ for each $I, J$. For injectivity, one just back-solves the relations. See [9] for details. Let $L=\left(a_{i j}\right)$ be an $r \times n$ matrix representing a point in $G(r, n)$. In the following calculation, we expand the first (red) determinant along its last row; the expression in the penultimate step evaluates to zero since the final (blue) matrix has a repeated row:

$$
\begin{aligned}
& P_{I, J}(\Lambda(L))=\sum_{k=1}^{r+1}(-1)^{k} \operatorname{det} L_{i_{1}, \ldots, i_{r-1}, j_{k}} \operatorname{det} L_{j_{1}, \ldots, \widehat{j_{k}}, \ldots, j_{r+1}} \\
& \left.=\sum_{k=1}^{r+1}(-1)^{k}|\underbrace{\left|\begin{array}{cc}
a_{1 j_{k}} \\
\cdots & \vdots \\
a_{r, j_{k}}
\end{array}\right|}_{\text {expand }}| \begin{array}{ccc}
\widehat{a_{1 j_{k}}} & \\
& \cdots & \vdots \\
& \cdots \\
a_{r j_{k}} &
\end{array} \right\rvert\, \\
& = \pm \sum_{k=1}^{r+1}(-1)^{k} \sum_{\ell=1}^{r}(-1)^{\ell} a_{\ell j_{k}}\left|\begin{array}{ccc} 
& \vdots & \\
\widehat{a_{\ell i_{1}}} & \cdots & \widehat{a_{\ell i_{r-1}}} \\
\vdots & & \\
& \widehat{a_{1 j_{k}}} & \\
\cdots & \vdots & \cdots \\
& \widehat{a_{r j_{k}}} &
\end{array}\right| \\
& = \pm \sum_{\ell=1}^{r}(-1)^{\ell}\left|\begin{array}{ccc} 
& \vdots & \\
\widehat{a_{\ell i_{1}}} & \cdots & \widehat{a_{\ell i_{r-1}}} \\
\vdots &
\end{array}\right|\left(\begin{array}{llll}
r+1 \\
k=1
\end{array}(-1)^{k} a_{\ell j_{k}}\left|\begin{array}{ccc} 
& \widehat{a_{1 j_{k}}} & \\
& \vdots & \ldots \\
\widehat{a_{r j_{k}}} &
\end{array}\right|\right)
\end{aligned}
$$

$$
\begin{aligned}
& = \pm \sum_{\ell=1}^{r}(-1)^{\ell}\left|\begin{array}{ccc} 
& \vdots & \\
\widehat{a_{\ell i_{1}}} & \cdots & \widehat{a_{\ell i_{r-1}}} \\
\vdots &
\end{array}\right|\left(\left|\begin{array}{cccc}
a_{\ell j_{1}} & \cdots & a_{\ell j_{k}} & \cdots \\
a_{1 j_{1}} & \cdots & a_{\ell j_{k}} & \cdots \\
& a_{1 j_{r+1}} \\
& & \vdots & \\
a_{r j_{1}} & \cdots & a_{r j_{k}} & \cdots \\
& a_{r j_{r+1}}
\end{array}\right|\right) \\
& =0
\end{aligned}
$$

14.3. Schubert varieties. Our goal is to describe the cohomology ring of a Grassmannian. In this section we describe sets whose classes will give a basis for that ring. There are many closely-related cohomology theories. We have carefully considered de Rham cohomology. Appendix D is an introduction to abstract simplicial homology, which can be dualized (by applying the function hom $(\cdot, R)$ ) to define cohomology groups). It is fairly straightforward to relate abstract simplicial cohomology to a cohomology theory for manifolds by, roughly, triangulating a manifold. In section 13, we gave an operational definition of the Chow ring of a smooth complete toric variety, and that will be the approach we take with Grassmannians. We will give a very rough idea of the construction of the Chow ring, in general, and give a recipe for computing it for the Grassmannian. We will then describe in detail the meaning of the Chow classes for the Grassmannian and use calculations in the Chow ring to solve interesting enumerative problems.
14.3.1. The Chow ring. An algebraic set is the set of solutions to a system of polynomial equations. An algebraic set is irreducible if it cannot be written as a proper union of algebraic sets. An irreducible algebraic set is called a variety.

Example 14.16. The algebraic set $X=\left\{(x, y) \in \mathbb{R}^{2} \bmod x\left(y-x^{2}\right)=0\right\}$ is reducible. It can be written as the union of the $y$-axis and a parabola, both of which are varieties:

$$
X=\left\{(x, y) \in \mathbb{R}^{2} \mid x=0\right\} \cup\left\{(x, y) \in \mathbb{R}^{2} \mid y-x^{2}=0\right\}
$$



If our defining polynomials come from the polynomial ring $\mathbb{k}\left[x_{1}, \ldots, x_{n}\right]$, then the corresponding algebraic set is a subset of $\mathbb{k}^{n}$. If $X$ is defined by the polynomials $f_{1}, \ldots, f_{k}$, we write $X=Z\left(f_{1}, \ldots, f_{k}\right)$. If the defining polynomials are homogeneous, then we may regard $X$ as a subset of projective space.

From now on, we will take $\mathbb{k}=\mathbb{R}$ or $\mathbb{C}$. When working over $\mathbb{C}$, there is the option of replacing $\mathbb{C}$ by $\mathbb{R}^{2}$ to define an algebraic set over $\mathbb{R}$.

Example 14.17. Consider the algebraic set $X=\left\{(w, z) \in \mathbb{C}^{2} \mid w=z^{2}\right\}$. This is a variety in $\mathbb{C}^{2}$, but since $\mathbb{C}=\mathbb{R}^{2}$, we may also consider it as a variety in $\mathbb{R}^{4}$. Letting $w=a+b i$ and $z=c+d i$ and equating real and imaginary parts in $w=z^{2}$, then $X \subset \mathbb{R}^{2}$ is defined by the system of equations

$$
a=c^{2}-d^{2} \quad \text { and } \quad b=2 c d
$$

Thus, considering $X$ as a subset of $\mathbb{R}^{4}$, it is the intersection of two quadric hypersurfaces ${ }^{11}$ giving a (two-dimensional) a surface in $\mathbb{R}^{4}$.

There are many ways to define the dimension of a variety $X \in \mathbb{k}^{n}$. We will state one possibility that does not require a long digression. Say $X$ is defined by the system of equations $f_{1}=\cdots=f_{k}=0$,

$$
\operatorname{dim} X=n-\max _{p \in X} \operatorname{rk}\left(\frac{\partial f_{i}}{\partial x_{j}}\right)_{i, j}
$$

A point $p$ at which the Jacobian matrix $\left(\frac{\partial f_{i}}{\partial x_{j}}\right)_{i, j}$ reaches its maximum rank is a smooth point, and a non-smooth point is a singular point. The set of smooth points of $X$ will form a manifold.

Example 14.18. Let $X=\left\{(x, y, z) \in \mathbb{R}^{3} \mid z^{2}=x y\right\}$, a cone, with a single defining equation $f=z^{2}-x y$. The Jacobian matrix is

$$
J=\left(\begin{array}{ccc}
\frac{\partial f}{\partial x} & \frac{\partial f}{\partial y} & \frac{\partial f}{\partial z}
\end{array}\right)=\left(\begin{array}{ccc}
-y & -x & 2 z
\end{array}\right)
$$

It has maximum rank of 1 , which occurs at all points of $X$ except $(0,0,0)$. The dimension of $X$ is $3-1=2$, and $(0,0,0)$ is the only singular point.

Let $X$ be a variety. An $r$-cycle of $X$ is a finite formal sum $\sum_{i} n_{i} V_{i}$ where $n_{i} \in \mathbb{Z}$ and $V_{i}$ is an $r$-dimensional subvariety of $X$. Let $Z_{r}(X)$ denote the set of all $r$ cycles, an abelian group under addition. A subvariety $V \subseteq X$ has codimension $r$ is $\operatorname{dim} V=\operatorname{dim} X-r$. So $Y$ is a hypersurface in $X$ if it has codimension one. There is a notion of rational equivalence of subvarieties, which we will not define, but which is an algebraic version of homotopy equivalence, If subvarieties $V$ and $W$ are rationally equivalent, we will write $V \sim W$. Rational equivalence preserves many essential properties. For instance, if $V \sim W$, then $\operatorname{dim} V=\operatorname{dim} W$.

Definition 14.19. Suppose that $X$ is a smooth variety of dimension $n$. Let $A^{r}=$ $Z_{n-r}(X) / \sim$. The Chow ring of $X$ is $A^{\bullet}(X)=\oplus_{r \geq 0} A^{r}(X)$.

To define the ring structure on $A^{\bullet}(X)$, let $[V] \in A^{r}(X)$ and $[W] \in A^{s}(X)$. Then let $[V] \cdot[W]=[V \cap W] \in A^{r+s}(X)$ where the intersection is done after deforming $V$ and $W$ (replacing $V$ and $W$ with rationally equivalent subvarieties) so that $V$ and $W$ meet transversally everywhere. To meet transversally means that the tangent spaces of $V$ and $W$ at each point $p$ of there intersection together span the tangent space of of $X$.

[^10]Define the degree of a variety $X$ in $\mathbb{k}^{n}$ or $\mathbb{P}_{\mathbb{k}}^{n}$ is the number of points in the intersection (over the complex numbers) of $X$ with a generic linear space $L$ of complementary dimension, i.e., such that $\operatorname{dim} L=n-\operatorname{dim} X$.

Example 14.20. Let $X=\mathbb{P}^{n}$. It turns out that

$$
\begin{aligned}
A^{\bullet}(X) & \simeq \\
& \mathbb{Z}[t] /\left(t^{n+1}\right)=\mathbb{Z}+\mathbb{Z} t+\cdots+\mathbb{Z} t^{n} \\
V & \mapsto(\operatorname{deg} V) t^{\operatorname{codim} V} .
\end{aligned}
$$

For example, let $n=2$, and take $p, q \in \mathbb{P}^{2}, X=Z\left(y z-x^{2}\right)$, and $Y=Z\left(z y^{2}-x^{3}-\right.$ $z x^{2}$ ). Then we have the following examples of calculations in the Chow ring:

- $2[p]+3[q]-[X]+4[Y]+7\left[\mathbb{P}^{2}\right] \mapsto 2 t^{2}+3 t^{2}-2 t+4(3 t)+7=7+10 t+5 t^{2}$.
- $\left(3[p]+[X]+\left[\mathbb{P}^{2}\right]\right)^{2} \mapsto\left(3 t^{2}+2 t+1\right)^{2}=1+4 t+10 t^{2}+12 t^{3}+9 t^{4}$.
- $[X][Y] \mapsto(2 t)(3 t)=6 t^{2}$.

The class of a point in $\mathbb{P}^{2}$ is given by $t^{2}$. Hence, the last calculation shows that $X$ and $Y$ meet in in 6 points after continuous deformation.
14.3.2. Schubert varieties. For details regarding the following, see [9]. Throughout, when necessary, we assume that $\mathbb{G}_{r} \mathbb{P}^{n}$ is embedded in $\mathbb{P}^{\binom{n}{r}}$ via the Plücker embedding.

Definition 14.21. A sequence $A_{0} \subsetneq \cdots \subsetneq A_{r}$ of linear subspaces of $\mathbb{P}^{n}$ is called a flag. The Schubert variety corresponding to a flag is

$$
\mathfrak{S}\left(A_{0}, \ldots, A_{r}\right)=\left\{L \in \mathbb{G}_{r} \mathbb{P}^{n} \mid \operatorname{dim}\left(L \cap A_{i}\right) \geq i \text { for all } i\right\}
$$

Proposition 14.22. We have

$$
\mathfrak{S}\left(A_{0}, \ldots, A_{r}\right)=\mathbb{G}_{r} \mathbb{P}^{n} \cap M
$$

for some linear subspace $M$ of $\mathbb{P}\left(\begin{array}{c}\binom{n+1}{r+1}-1\end{array}\right.$. The space $M$ is a hyperplane if and only if $\operatorname{dim} A_{0}=n-r-1$ and $\operatorname{dim} A_{i}=n-r+i$ for $i=1, \ldots, r .\left(S o\left(A_{r}, A_{r-1}, \ldots, A_{1}\right)=\right.$ $(n, n-1, \ldots, n-r+1)$ and $\operatorname{dim} A_{0}=n-r-1$ is one less than expected given the previous sequence.)

Proposition 14.23. If $A_{0} \subsetneq \cdots \subsetneq A_{r}$ and $B_{0} \subsetneq \cdots \subsetneq B_{r}$ are flags in $\mathbb{P}^{n}$ with $\operatorname{dim} A_{i}=\operatorname{dim} B_{i}$ for all $i$, then there is an invertible linear transformation of $\mathbb{P}^{\binom{n+1}{r+1}-1}$ that sends $\mathfrak{S}\left(A_{0}, \ldots, A_{r}\right)$ to $\mathfrak{S}\left(B_{0}, \ldots, B_{r}\right)$, and

$$
\left[\mathfrak{S}\left(A_{0}, \ldots, A_{r}\right)\right]=\left[\mathfrak{S}\left(B_{0}, \ldots, B_{r}\right)\right] \in A \cdot \mathbb{G}_{r} \mathbb{P}^{n}
$$

Definition 14.24. If $\operatorname{dim} A_{i}=a_{i}$ for all $i$, then we write $\mathfrak{S}\left(a_{0}, \ldots, a_{r}\right)$ or just $\left(a_{0}, \ldots, a_{r}\right)$ for the class $\left[\mathfrak{S}\left(A_{0}, \ldots, A_{r}\right)\right] \in A \cdot \mathbb{G}_{r} \mathbb{P}^{n}$. The $\left(a_{0}, \ldots, a_{r}\right)$ are called Schubert cycles or Schubert classes.

Theorem 14.25. The Chow ring $A \bullet \mathbb{G}_{r} \mathbb{P}^{n}$ is a free abelian group on the Schubert cycles $\left(a_{0}, \ldots, a_{r}\right)$ (i.e., it is generated by the Schubert cycles, and there are no nontrivial integer combinations of the Schubert cycles that are 0). The codimension of $\left(a_{0}, \ldots, a_{r}\right)$ is $k:=(r+1)(n-r)-\sum_{i}\left(a_{i}-i\right)$, i.e., $\left(a_{0}, \ldots, a_{r}\right) \in A^{k} \mathbb{G}_{r} \mathbb{P}^{n}$.

The above result might be considered the solution to Hilbert's 15th problem ([7]):

The problem consists in this: To establish rigorously and with an exact determination of the limits of their validity those geometrical numbers which Schubert especially has determined on the basis of the so-called principle of special position, or conservation of number, by means of the enumerative calculus developed by him.

Although the algebra of today guarantees, in principle, the possibility of carrying out the processes of elimination, yet for the proof of the theorems of enumerative geometry decidedly more is requisite, namely, the actual carrying out of the process of elimination in the case of equations of special form in such a way that the degree of the final equations and the multiplicity of their solutions may be foreseen.

Example 14.26. Consider the Grassmannian of planes in five-space, $\mathbb{G}_{2} \mathbb{P}^{5}$. Here, we consider the meaning of some Schubert varieties.
I. Let $\left(a_{0}, a_{1}, a_{2}\right)=(1,3,4)$. Then

$$
\mathfrak{S}\left(A_{0}, A_{1}, A_{2}\right)=\left\{L \in \mathbb{G}_{2} \mathbb{P}^{5} \mid \operatorname{dim}\left(L \cap A_{i}\right) \geq i \text { for all } i=0,1,2\right\}
$$

The meaning is given by

$$
\begin{aligned}
& \operatorname{dim} L \cap A_{0} \geq 0 \Rightarrow L \text { meets the line } A_{0} \text { in at least a point } \\
& \operatorname{dim} L \cap A_{1} \geq 1 \Rightarrow L \text { meets the solid } A_{1} \text { in at least a line } \\
& \operatorname{dim} L \cap A_{2} \geq 2 \Rightarrow L \text { meets the } 4 \text {-plane } A_{1} \text { in at least a plane. }
\end{aligned}
$$

Since the dimension of $L$ is 2 , the last condition means that $L$ lies in the 4 -plane $A_{2}$.
II. Let $\left(a_{0}, a_{1}, a_{2}\right)=(3,4,5)$. Then $\mathfrak{S}\left(A_{0}, A_{1}, A_{2}\right)$ is those planes $L$ in $\mathbb{P}^{5}$ such that $\operatorname{dim} L \cap A_{0} \geq 0, \quad \operatorname{dim} L \cap A_{1} \geq 1 \quad$ and $\quad \operatorname{dim} L \cap A_{2} \geq 2$.

However, by Proposition 14.2, these conditions are satisfied by all planes $L$. So there is not condition at all, and $\mathfrak{S}\left(A_{0}, A_{1}, A_{2}\right)=\mathbb{G}_{2} \mathbb{P}^{5}$.
III. Let $\left(a_{0}, a_{1}, a_{2}\right)=(1,4,5)$. Then $\mathfrak{S}\left(A_{0}, A_{1}, A_{2}\right)$ represents the condition that a 2-plane meets a given line in a least a point, i.e., the plane intersects a given line. The conditions $\operatorname{dim} L \cap A_{1} \geq 1$ and $\operatorname{dim} L \cap A_{2} \geq 2$ are non-conditions by Proposition 14.2.

When is it the case that $\operatorname{dim} L \cap A_{i} \geq i$ imposes no condition, i.e., when is it the case that every $r$-plane $L$ satisfies this restriction? By Proposition 14.2,

$$
\operatorname{dim} L \cap A_{i} \geq \operatorname{dim} L+\operatorname{dim} A_{i}-n=r+a_{i}-n
$$

for all $L$ and $A_{i}$. Therefore, $\operatorname{dim} L \cap A_{i} \geq i$ for all $L \in \mathbb{G}_{r} \mathbb{P}^{n}$ if $r+a_{i}-n \geq i$, i.e., if

$$
a_{i} \geq n-r+i
$$

For instance any of $a_{r} \geq n$, or $a_{r-1} \geq n-1$, or $a_{r-2} \geq n-2$ all impose no condition. Since $A_{0} \subsetneq \cdots \subsetneq A_{r} \subseteq \mathbb{P}^{n}$, we have $0 \leq a_{0}<\cdots<a_{r} \leq n$. So if $a_{i}=n-r+i$ for some $i$, then $a_{j}=n-r+j$ for all $j \geq i$.

Example 14.27. In $\mathbb{G}_{3} \mathbb{P}^{6}$

$$
(3,4,5)=\left[\mathbb{G}_{3} \mathbb{P}^{6}\right]
$$

In $(2,4,5)$, the entries $a_{1}=4$ and $a_{2}=5$ impose no conditions. So the only condition of consequence is that $\operatorname{dim} L \cap A_{0} \geq 0$. This is the condition that the solid (i.e., the 3-plane) $L$ intersects the plane $A_{0}$. (In 6 -space, there is just room for a solid and a plane to not meet.) Similarly, in ( $1,3,5$ ), the entry $a_{2}=5$ imposes no condition. So $(1,3,5)$ can be thought of as the class in the Chow ring represented by all solids $L$ that intersects a given $\operatorname{line}\left(\operatorname{dim} L \cap A_{0} \geq 0\right)$ and meet a given solid in at least a line $\left(\operatorname{dim} L \cap A_{1} \geq 1\right)$.

Example 14.28. A basis for the Chow ring $A \cdot \mathbb{G}_{1} \mathbb{P}^{3}$ consist of the Schubert classes $\left(a_{0}, a_{1}\right)$ with $0 \leq a_{0}<a_{1} \leq 3$. Here is a table of theses classes and their interpretation (the alternate notation for the class using curly braces is for use in the next section):

| codimension | class | condition |
| :---: | :---: | :--- |
| 0 | $(2,3)=\{0,0\}$ | no condition |
| 1 | $(1,3)=\{1,0\}$ | meet a given line |
| 2 | $(0,3)=\{2,0\}$ | pass through a given point |
| 2 | $(1,2)=\{1,1\}$ | lie in a given plane |
| 3 | $(0,2)=\{2,1\}$ | pass through a given point and lie in a given plane |
| 4 | $(0,1)=\{2,2\}$ | to be a certain line. |

For instance, the last class, ( 0,1 ), imposes the conditions $\operatorname{dim} L \cap A_{0} \geq 0$ and $\operatorname{dim} L \cap$ $A_{1} \geq 1$. So $L$ must contain the point $A_{0}$ and must meet the line $A_{1}$ in a space of dimension 1, i.e., $L$ must equal $A_{1}$, at which point $L$ contains the point $A_{0}$ automatically.
14.3.3. The Schubert calculus. In this section, we describe the ring structure for $A \cdot \mathbb{G}_{r} \mathbb{P}^{n}$. Given a Schubert class $\left(a_{0}, \ldots, a_{r}\right)$, define the integers

$$
\lambda_{i}=n-r-\left(a_{i}-i\right)
$$

for $i=0, \ldots, r$. Since $0 \leq a_{0}<\cdots<a_{r} \leq n$, it follows that each $a_{i}-i \geq 0$ and

$$
n-r \geq \lambda_{0} \geq \cdots \geq \lambda_{r} \geq 0
$$

Further,

$$
|\lambda|:=\sum_{i=0}^{r} \lambda_{i}=(r+1)(n-r)-\sum_{i=0}^{r}\left(a_{i}-i\right)=\operatorname{codim}\left(a_{0}, \ldots, a_{r}\right) .
$$

Thus,

$$
\left\{\lambda_{0}, \ldots, \lambda_{r}\right\} \in A^{|\lambda|} \mathbb{G}_{r} \mathbb{P}^{n}
$$

Notation: We denote the Schubert class $\left(a_{0}, \ldots, a_{r}\right)$ by $\left\{\lambda_{0}, \ldots, \lambda_{r}\right\}$.
For each $\lambda$ there is an associated Young diagram consisting of rows of left-justified unit boxes such that row $i$ has $\lambda_{i}$ boxes. The following example should make the definition clear.

Example 14.29. The Schubert class $(1,4,5,7)=\{4,2,2,1\} \in \mathbb{G}_{3} \mathbb{P}^{8}$ has Young diagram


The codimension is $|\{4,2,2,1\}|=9$ and the dimension of $\mathbb{G}_{3} \mathbb{P}^{8}$ is $(3+1)(8-3)=20$. Thus, a representative Schubert variety $\mathfrak{S}\left(A_{0}, \ldots, A_{r}\right)$ has dimension $20-9=11$.

The product of two Schubert classes $\{\lambda\}$ and $\{\mu\}$ will be a unique integer linear combination of Schubert classes. So we can write

$$
\{\lambda\} \cdot\{\mu\}=\sum_{\{\nu\} \in A \bullet \mathbb{G}_{r} \mathbb{P}^{n}} c_{\lambda \mu}^{\nu}\{\nu\} .
$$

The integers $c_{\lambda \mu}^{\nu}$ are called Littlewood-Richardson coefficients. It was shown in 2006 that the general problem of computing Littlewood-Richardson numbers is \#-P complete. Nevertheless, there are several ways of calculating them, of which we now describe one.

Definition 14.30. Let $\{\lambda\},\{\mu\} \in A \cdot \mathbb{G}_{r} \mathbb{P}^{n}$. Form the $\mu$-expansion of $\lambda$ in steps as follows. Let $Y$ be the Young diagram for $\lambda$. Use $\mu$ to build a new Young diagram in steps from $Y$ as follows: At the $i$-th step, add $\mu_{i}$ boxes to existing rows or columns with the restriction that no two of these boxes can be added to the same column. Next, write the number $i$ in each of the added boxes. The resulting shape must be a Young diagram. Then proceed to step $i+1$ (stopping with the last part of $\mu$ is reached). The expansion is strict if reading off the numbers right-to-left and top-to-bottom, each number $i$ occurs at least as many times as the number $i+1$.

Example 14.31. Some examples of $\{2,1\}$ expansions of $\{3,1\}$ :


The following expansions are not allowed-the first because there are two boxes with 0s in the same column, and the second because it is not a Young diagram:


Theorem 14.32. The Littlewood-Richardson number $c_{\lambda \mu}^{\nu}$ is the number of strict $\mu$ expansions of $\lambda$ resulting in $\nu$.

From the theorem, it is easy to see that $c_{\lambda \mu}^{\nu}=0$ unless $\lambda, \mu \subseteq \nu$, i.e., unless $\lambda_{i} \leq \nu_{i}$ and $\mu_{i} \leq \nu_{i}$ for all $i$. It is a little harder to see that $c_{\lambda \mu}^{\nu}=c_{\mu \lambda}^{\nu}$.

Remark 14.33. From now on, we will feel free to represent a Schubert class by its corresponding Young diagram. Take note of the following
(1) The Young diagrams that represent Schubert classes for $\mathbb{G}_{r} \mathbb{P}^{n}$ must fit in an $(r+$ $1) \times(n-r)$ box of unit squares. Thus, when computing products in the Chow ring using the formula $\{\lambda\} \cdot\{\mu\}=\sum_{\nu \in A \cdot \mathbb{G}_{r} \mathbb{P}^{n}} c_{\lambda \mu}^{\nu}\{\nu\}$ are those $\nu$ that fit in that box.
(2) The identity in $A \cdot \mathbb{G}_{r} \mathbb{P}^{n}$ is $\left[\mathbb{G}_{r} \mathbb{P}^{n}\right]$ since intersecting a subvariety of $\mathbb{G}_{r} \mathbb{P}^{n}$ with the whole space does not change the variety. We have $\left[\mathbb{G}_{r} \mathbb{P}^{n}\right]=(n-r, n-r+$ $1, \ldots, n)=\{0, \ldots, 0\}=\lambda$. Note that $|\lambda|=0$, which agrees with the fact that the class has codimension 0 .
(3) The class of a single point in $\mathbb{G}_{r} \mathbb{P}^{n}$, i.e., of an $r$-plane in $\mathbb{P}^{n}$, has the form $a=(0,1, \ldots, r)=\{n-r, \ldots, n-r\}=\lambda$, with Young diagram consisting of the $(r+1) \times(n-r)$ rectangle. Note that the codimension is $|\lambda|=(r+1)(n-r)=$ $\operatorname{dim} \mathbb{G}_{r} \mathbb{P}^{n}$, as expected.

Example 14.34. The reader should verify the following multiplication table for $A \bullet \mathbb{G}_{1} \mathbb{P}^{3}$ :


The class of $\square$ represents the condition "to meet a given line" and has codimension 1 , therefore $\square^{4}$ represents the condition "to meet four given lines". To determine how many lines in 3 -space meet four generic lines, we use the table to compute:

$$
\square^{4}=\square^{2}(\square+\square)=\square(2 \square)=2 \square
$$

Since $\square$ is the class of a point, i.e., of a line in 3-space, we conclude that there are two lines meeting four generic lines in 3 -space.

Here is a sketch of another proof that there are two lines meeting four general lines in 3 -space. Call the four general lines $L_{1}, L_{2}, L_{3}, L_{4}$. Walk along $L_{3}$, and at each point $p \in L_{3}$, stop and look out at the lines $L_{1}$ and $L_{2}$. They will appear to cross at some point. Draw a line from $p$ through that point. One may check that the collection of all lines drawn in this way forms a quadric surface $Q$ in 3space - the zero-set of a single equation of degree two. Through each point in $Q$, there is a line lying on $Q$ that meets $L_{1}, L_{2}$ and $L_{3}$. Now consider line $L_{4}$. A line and a surface in $\mathbb{P}^{3}$ must meet, and since $L_{4}$ is general, it will not be tangent to $Q$. Since $Q$ is defined by an equation of degree two, the line will meet in two points, each of which corresponds to a line we drew earlier. These two lines are exactly those meeting $L_{1}, L_{2}, L_{3}$, and $L_{4}$.

Finally, we mention a proof given by Schubert using his "principle of conservation of number": that the number of solutions will remain the same as the parameters of the configuration are continuously changed as long as the number of solutions stays finite. Suppose that $L_{1}$ and $L_{2}$ lie in a plane $P_{12}$ and $L_{3}$ and $L_{4}$ lie in a plane $P_{34}$. How many lines meet these four lines now that they are in special position? Since we are working in projective space, both $L_{1} \cap L_{2}$ and $L_{3} \cap L_{4}$ will be points. The line through these two points yields one solution. The intersection of $P_{12}$ and $P_{34}$ will be a line. Since that line sits in $P_{12}$ it will meet both $L_{1}$ and $L_{2}$, and it similarly meets both $L_{3}$ and $L_{4}$.

## Appendix A. Vector Calculus in Euclidean Spaces

A.1. Derivatives. Let $f: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ with the understanding that what we say below can be suitably modified to apply to the more general case where the domain and codomain of $f$ are arbitrary open subsets of $\mathbb{R}^{n}$ and $\mathbb{R}^{m}$, respectively.

Definition A.1. Let $U \subseteq \mathbb{R}^{m}$ and $V \subseteq \mathbb{R}^{n}$ be open. A function $f: U \rightarrow V$ is smooth if each of its components has all partial derivatives of all orders. If $f$ is bijective and has a smooth inverse, then we say that $f$ is a diffeomorphism.

Definition A.2. The function $f$ is differentiable at $p \in \mathbb{R}^{n}$ if there is a linear function $D f_{p}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ such that

$$
\lim _{h \rightarrow 0} \frac{\left|f(p+h)-f(p)-D f_{p}(h)\right|}{|h|}=0
$$

The function $D f_{p}$ is called the derivative of $f$ at $p$. If each of the component functions of $f$ have partial derivatives of all orders at $p$, then $f$ is smooth at $p$, also known as being in the $C^{\infty}$, (in which case $f$ is differentiable at $p$-see Theorem A.3.) We say $f$ is differentiable (resp. smooth), if it is differentiable (resp. smooth) at all $p \in \mathbb{R}^{n}$.

Note: In this text will assume all of our differentiable mappings are smooth and use the words "differentiable" and "smooth" interchangeably.

Theorem A.3. If $f$ is differentiable at $p \in \mathbb{R}^{n}$, then each partial derivative of each component function of $f$, i.e., $\partial f_{i}(p) / \partial x_{i}$, exists, and $D f_{p}$ is the linear map associated with the Jacobian matrix of $f$ at $p$ :

$$
J f_{p}:=\left[\begin{array}{ccc}
\frac{\partial f_{1}(p)}{\partial x_{1}} & \cdots & \frac{\partial f_{1}(p)}{\partial x_{n}} \\
\vdots & \ddots & \vdots \\
\frac{\partial f_{m}(p)}{\partial x_{1}} & \cdots & \frac{\partial f_{m}(p)}{\partial x_{n}}
\end{array}\right]
$$

Conversely, if each partial derivative of each component function exists and is continuous in an open set containing $p$, then $f$ is differentiable at $p$.

If $m=1$, then $J f_{p}$ is a single row vector called the gradient of $f$ at $p$, denoted $\nabla_{p} f$ or $\operatorname{grad}_{p} f$. If $n=1$ and $m$ is arbitrary, then $f$ is a parametrized curve and $J f_{p}$ is a single column vector $=$ the velocity of the curve at $p$. (In general, the columns of $J f_{p}$ may be thought of as tangent vectors spanning the tangent space to $f$ at p.)

In what follows below, we make statements about functions $f: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ with the understanding that, with minor adjustments, they apply equally well to functions of the form $f: U \rightarrow V$ for open subsets of $\mathbb{R}^{n}$ and $\mathbb{R}^{m}$.

Theorem A.4. (Chain Rule.) If $f: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ is differentiable at $u$, and $g: \mathbb{R}^{m} \rightarrow$ $\mathbb{R}^{p}$ is differentiable at $f(u)$, then the composition $g \circ f: \mathbb{R}^{n} \rightarrow \mathbb{R}^{p}$ is differentiable at $u$ and

$$
D(g \circ f)_{u}=D g_{f(u)} \circ D f_{u}
$$

or, in terms of Jacobian matrices,

$$
J(g \circ f)_{u}=J g_{f(u)} \circ J f_{u}
$$

Theorem A.5. (Inverse Function Theorem.) Suppose that all the partial derivatives of each component function of $f: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ exist and are continuous in an open set containing $u$ and $\operatorname{det} J f_{u} \neq 0$. Then there is an open set $W$ containing $u$ and an open set $V$ containing $f(u)$ such that $f: W \rightarrow V$ has a continuous inverse $f^{-1}: V \rightarrow W$ which is differentiable for all $v \in V$ and

$$
J\left(f^{-1}\right)_{v}=\left(J f_{f^{-1}(v)}\right)^{-1}
$$

Definition A.6. Let $U, V$ be open subsets of $\mathbb{R}^{n}$. A mapping $f: U \rightarrow V$ is a local diffeomorphism if for all $p \in U$ there exists an open neighborhood $\tilde{U} \subseteq U$ of $p$ such that $f(\tilde{U})$ is an open subset of $V$

$$
\left.f\right|_{\tilde{U}}: \tilde{U} \rightarrow f(\tilde{U})
$$

is a diffeomorphism.
Remark A.7. Note that if $f: U \rightarrow V$ is a local diffeomorphism, the $f(U)$ is open in $V$. Further, by the inverse function theorem, if $\operatorname{det}\left(J f_{u}\right) \neq 0$ for all $u \in U$, then $f$ is a local diffeomorphism.

Definition A.8. Let $f: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ be a differentiable function, and let $p \in \mathbb{R}^{n}$. The best affine approximation to $f$ at $p$ is the function

$$
A f_{p}(x):=f(p)+D f_{p}(x)
$$

## A.2. Classical integral vector calculus.

A.2.1. Integral over a set.

Definition A.9. Let $K$ be a bounded subset of $\mathbb{R}^{n}$, and let $f: K \rightarrow \mathbb{R}$ be a bounded function. To define the integral of $f$ over $K$, choose a rectangle $A$ containing $K$ and define the function $\tilde{f}$ on $A$ which agrees with $f$ on $K$ and is zero otherwise. The integral of $f$ over $K$ is then

$$
\int_{K} f:=\int_{A} \tilde{f}
$$

provided this integral exists.
Remark A. 10.
(1) One may show that the definition does not depend on the choice of the bounding rectangle, $A$.
(2) If $\int_{K} f$ exists, we say $f$ is integrable (over $K$ ).
(3) In class, we showed that if $f$ is a bounded function defined on any rectangle $A$ and is continuous except for a set of volume zero, then $f$ is integrable over $A$. It follows that if $K$, from above, is compact (i.e., closed and bounded), has boundary of volume zero, and $f$ is continous on $K$, then $f$ is integrable over $K$. Recall that the boundary of $K$ is the set

$$
\partial K:=\bar{K} \cap \overline{K^{\prime}},
$$

the intersection of the closure of $K$ and the closure of the complement of $K$.

Theorem A.11. (Change of variables.) Let $K \subset \mathbb{R}^{n}$ be a compact and connected set having boundary with measure zero. Let $U$ be an open set containing K. Suppose that $\phi: A \rightarrow \mathbb{R}^{n}$ is a $C^{1}$-mapping (i.e., with continuous first partials) and such that $\phi$ is injective on the interior $K^{\circ}$ and $\operatorname{det}(J \phi) \neq 0$ on $K^{\circ}$. Then if $f: \phi(K) \rightarrow$ $\mathbb{R}^{n}$ is continous, it follows that

$$
\left.\int_{\phi(K)} f=\int_{K}(f \circ \phi) \mid \operatorname{det}(J \phi)\right) \mid .
$$

For the definitions of boundary and interior is Definition B.1. In the change of variables theorem, we have a sequence of mappings

$$
K \xrightarrow{\phi} \phi(K) \xrightarrow{f} \mathbb{R}^{n} .
$$

The mapping $\phi$ is the change of variables. For instance, if we are integrating a function over a 2 -sphere, $K$ might be a rectangle and $\phi$ spherical coordinates.

## A.2.2. Curve integrals.

Definition A.12. Let $C:[a, b] \rightarrow \mathbb{R}^{n}$ be a differentiable curve in $n$-space, and let $f$ be a real-valued function defined on the image of $C$. The path integral of $f$ along $C$ is

$$
\int_{C} f:=\int_{a}^{b}(f \circ C)\left|C^{\prime}\right|
$$

If $f=1$, we get the length of $C$ :

$$
\operatorname{length}(C):=\int_{C} 1=\int_{a}^{b}\left|C^{\prime}\right|
$$

Remark A.13.
(1) In the above displayed formulas, the stuff on the left of the equals sign is just notation for the stuff on the right-hand side. In other words, the lefthand side has no independent meaning. The same holds for the formulas appearing in the definitions below.
(2) The path integral is sometimes called the line integral or the curve integral. Another notation for the path integral is $\int_{C} f d C$.
(3) The above integrals may not exist. They will exist if, for example, $f$ is continous on the image of $C$ and $C^{\prime}$ is continous.
(4) Think of the factor of $\left|C^{\prime}(t)\right|$ as a "stretching" factor, recording how much the domain, $[a, b]$ is stretched by $C$ as it is placed in $\mathbb{R}^{n}$.

Geometric Motivation. The integral of $f$ along $C$ is meant to measure the weighted (by $f$ ) length of $C$. To see where the definition comes from, imagine how one would go about calculating the length of $C$. One might approximate the image of $C$ by line segments, and in order to do this, it is natural to take a partition of the domain, $[a, b]$, say $a=t_{0} \leq t_{1} \leq \cdots \leq t_{k}=b$. The image of these partition points can be used to break up the image of $C$ into line segments as illustrated below:


Approximating a curve with matchsticks.

We want to add up the lengths of these line segments, weighting each one using $f$, which we might as well evaluate at an endpoint of each segment. Thus, we get

$$
\text { weighted length of } \begin{aligned}
C & \approx \sum_{i=1}^{k} f\left(C\left(t_{i}\right)\right)\left|C\left(t_{i}\right)-C\left(t_{i-1}\right)\right| \\
& =\sum_{i=1}^{k} f\left(C\left(t_{i}\right)\right)\left|\frac{C\left(t_{i}\right)-C\left(t_{i-1}\right)}{t_{i}-t_{i-1}}\right|\left(t_{i}-t_{i-1}\right) \\
& \approx \int_{a}^{b}(f \circ C)\left|C^{\prime}(t)\right|
\end{aligned}
$$

Definition A.14. Let $C:[a, b] \rightarrow \mathbb{R}^{n}$ be a curve in $n$-space, and let $F=\left(F_{1}, \ldots, F_{n}\right)$ be a vector field in $\mathbb{R}^{n}$ defined on the image of $C$. The flow of $F$ along $C$ is

$$
\int_{C} F \cdot d C:=\int_{C} w_{F}
$$

where $\omega_{F}:=F_{1} d x_{1}+\cdots+F_{n} d x_{n}$ is the 1-form associated with $F$.
Remark A. 15.
(1) Another notation for the flow is $\int_{C} F \cdot \vec{t}$. If $C$ is a closed curve, meaning $C(a)=C(b)$, then one often sees the notation $\oint_{C} F \cdot d C$.
(2) Working out the definition given above, one sees that

$$
\int_{C} F \cdot d C=\int_{a}^{b} F(C(t)) \cdot C^{\prime}(t)
$$

Geometric Motivation. As claimed above, the flow can be expressed as $\int_{a}^{b} F(C(t))$. $C^{\prime}(t)$, which can be re-written as

$$
\int_{a}^{b}\left[F(C(t)) \cdot \frac{C^{\prime}(t)}{\left|C^{\prime}(t)\right|}\right]\left|C^{\prime}(t)\right|
$$

Notice the unit tangent vector appearing inside the square brackets. The quantity inside the brackets is thus the component of $F$ in the direction in which $C$ is moving. The factor of $\left|C^{\prime}(t)\right|$ is the stretching factor which appears when taking a path integral of a function. Comparing with the geometric motivation given for the path integral, we see that the flow is measuring the length of $C$, weighted by the component of $F$ along $C$. Hence, it makes sense to call this integral the flow of $F$ along $C$.

## A.2.3. SURFACE INTEGRALS.

Definition A.16. Let $S: D \rightarrow \mathbb{R}^{3}$ with $D \subseteq \mathbb{R}^{2}$ be a differentiable surface in 3space, and let $f$ be a real-valued function defined on the image of $S$. The surface integral of $f$ along $S$ is

$$
\int_{S} f:=\int_{D}(f \circ S)\left|S_{u} \times S_{v}\right|
$$

where $S_{u}$ and $S_{v}$ are the partial derivatives of $S$ with respect to $u$ and $v$, respectively. If $f=1$, we get the surface area of $S$ :

$$
\operatorname{area}(S)=\int_{S} 1=\int_{D}\left|S_{u} \times S_{v}\right|
$$

Remark A. 17.
(1) The cross product of two vectors $v=\left(v_{1}, v_{2}, v_{3}\right)$ and $w=\left(w_{1}, w_{2}, w_{3}\right)$ in $\mathbb{R}^{3}$ is given by

$$
v \times w:=\left(v_{2} w_{3}-v_{3} w_{2}, v_{3} w_{1}-v_{1} w_{3}, v_{1} w_{2}-v_{2} w_{1}\right)
$$

which may be easier to remember with one of the following mnemonic devices:

$$
v \times w=\operatorname{det}\left[\begin{array}{ccc}
\mathbf{i} & \mathbf{j} & \mathbf{k} \\
v_{1} & v_{2} & v_{3} \\
w_{1} & w_{2} & w_{3}
\end{array}\right]=([2,3],[3,1],[1,2]),
$$

thinking of reading off the numbers 1,2 , and 3 arranged around a circle.
The important things to remember about the cross product are: (i) it is a vector perpendicular to the plane spanned by $v$ and $w$, (ii) its direction is given by the "right-hand rule", and (iii) its length is the area of the parallelogram spanned by $v$ and $w$.
(2) The factor $\left|S_{u} \times S_{v}\right|$ is the area of the parallelogram spanned by $S_{u}$ and $S_{v}$ and should be thought of as the factor by which space is locally stretched by $f$ (analogous to $\left|C^{\prime}(t)\right|$ for curve integrals).

Geometric Motivation. The integral of $f$ along $S$ is meant to measure the weighted (by $f$ ) surface area of $S$. To estimate this value, one might reasonably partition the domain of $S$. The image of the subrectangles would be warped parallelograms lying on the image of $S$ :


Warped parallelograms on a surface.

The warped parallelograms can be approximated by scaling the parallelograms spanned by the partial derivative vectors, $S_{u}$ and $S_{v}$. Hence, if $J$ is a subrectangle of the partition, then $S(J)$ is a warped parallelogram whose area is approximately that of the parallelogram spanned by $S_{u}$ and $S_{v}$, scaled by $\operatorname{vol}(J)$. Recalling that the area of the parallelogram spanned by $S_{u}$ and $S_{v}$ is the length of their cross product, we get

$$
\begin{aligned}
\text { weighted surface area } & \approx \sum_{J} f\left(S\left(x_{J}\right)\right) \text { area }(S(J)) \quad\left(x_{J} \text { any point in } J\right) \\
& \approx \sum_{J} f\left(S\left(x_{J}\right)\right) \operatorname{area}(J)\left|S_{u}\left(x_{J}\right) \times S_{v}\left(x_{J}\right)\right| \\
& \approx \int_{D}(f \circ S)\left|S_{u} \times S_{v}\right|
\end{aligned}
$$

Definition A.18. Let $D \subseteq \mathbb{R}^{2}$, and let $S: D \rightarrow \mathbb{R}^{3}$ be a differentiable surface in 3 -space, and let $F=\left(F_{1}, F_{2}, F_{3}\right)$ be a vector field defined on the image of $S$. The flux of $F$ through $S$ is

$$
\int_{S} F \cdot \vec{n}:=\int_{S} \omega_{F}
$$

where

$$
\begin{aligned}
\omega_{F} & :=F_{1} d y \wedge d z-F_{2} d x \wedge d z+F_{3} d x \wedge d y \\
& =F_{1} d y \wedge d z+F_{2} d z \wedge d x+F_{3} d x \wedge d y
\end{aligned}
$$

is the flux 2-form for $F$.
Remark A.19. (1) From the definition, it follows that

$$
\int_{S} F \cdot \vec{n}=\int_{D}(F \circ S) \cdot\left(S_{u} \times S_{v}\right)
$$

(2) To calculate the flux of a vector field $F$ through a hypersurface, $S$, in $\mathbb{R}^{n}$, integrate $\int_{S} \omega_{F}$ where $\omega_{F}:=\sum_{i=1}^{n}(-1)^{i-1} d x_{1} \wedge \cdots \wedge \widehat{d x_{i}} \wedge \cdots \wedge d x_{n}$. The notation signifies omitting $d x_{i}$ and wedging all of the remaining $d x_{j}$ 's in the $i$-term. By hypersurface, we mean that the domain of $S$ sits in $\mathbb{R}^{n-1}$. Letting $n=3$ recovers the usual notion of flux through a (2-dimensional) surface.

Geometric Motivation. To see the geometric motivation behind the definition of flux, one first needs to check that our definition for flux is equivalent to the integral $\int_{D}(F \circ S) \cdot\left(S_{u} \times S_{v}\right)$. Then, re-writing the intergral gives:

$$
\begin{aligned}
\text { flux } & =\int_{D}(F \circ S) \cdot\left(S_{u} \times S_{v}\right) \\
& =\int_{D}\left[(F \circ S) \cdot \frac{S_{u} \times S_{v}}{\left|S_{u} \times S_{v}\right|}\right]\left|S_{u} \times S_{v}\right|
\end{aligned}
$$

Note the unit normal appearing inside the square brackets. Thus, the quantity inside the brackets is the component of $F$ in the unit normal direction. The factor of $\left|S_{u} \times S_{v}\right|$ appearing outside the brackets is the "stretching" factor which appears in our earlier definition of the integral of a function along a surface. In light of the geometric motivation given in that situation (above), we see that the integral is measuring the surface area of $S$, weighted by the normal component of $F$ normal to the surface.

## A.2.4. SOLID INTEGRALS.

Definition A.20. Let $V: D \rightarrow \mathbb{R}^{n}$ with $D \subset \mathbb{R}^{n}$ be a differentiable mapping, and let $f$ be a real-valued function defined on the image of $V$. The integral of $f$ over $V$ is

$$
\int_{V} f:=\int_{D}(f \circ V)\left|\operatorname{det} V^{\prime}\right|
$$

where $V^{\prime}$ is the $n \times n$ Jacobian matrix for $V$. If $f=1$, we get the volume of $V$ :

$$
\operatorname{vol}(V)=\int_{V} 1=\int_{D}\left|\operatorname{det} V^{\prime}\right|
$$

Remark A.21.
(1) The distinguishing feature for this integral is that both the domain and codomain are subsets of $\mathbb{R}^{n}$.
(2) One could leave off the absolute value signs about the determinant. In that case, we would be taking orientation into account, and some parts of the integral could cancel with others.

Geometric Motivation. The integral is supposed to measure the volume of $V$ weighted by $f$. To this end, partition the domain, $D$. Let $J$ be a subrectangle of the partition. Then $V(J)$ is a warped rectangle which can be approximated by the parallelepiped spanned by the partials of $V$, i.e., the columns of the Jacobian matrix, $V^{\prime}$, scaled by $\operatorname{vol}(J)$. Recall that the volume of the parallelepiped spanned by the columns of a square matrix is given by the absolute value of the determinant of the matrix. Hence,

$$
\begin{aligned}
\text { weighted volume } & \approx \sum_{J} f\left(V\left(x_{J}\right)\right) \operatorname{vol}(V(J)) \quad\left(x_{J} \text { any point in } J\right) \\
& \approx \sum_{J} f\left(V\left(x_{J}\right)\right)\left|\operatorname{det} V^{\prime}\right| \operatorname{vol}(\mathrm{J}) \\
& \approx \int_{D}(f \circ V)\left|\operatorname{det} V^{\prime}\right|
\end{aligned}
$$

A.2.5. Grad, curl, and div. Stokes' theorem says that

$$
\int_{C} d \omega=\int_{\partial C} \omega
$$

for any $k$-chain $C$ in $\mathbb{R}^{n}$ and $(k-1)$-form $\omega$ in $\mathbb{R}^{n}$. In this section, we would like to consider the special cases: $k=1,2,3$.

$$
\text { Case } k=1
$$

To apply Stokes' theorem in the case $k=1$, we start with a 1-chain $C$ and a 0 -form $\omega$. We will consider the case where $C:[0,1] \rightarrow \mathbb{R}^{n}$ is a curve in $\mathbb{R}^{n}$. Recall that a 0 -form is a function: $\omega=f: \mathbb{R}^{n} \rightarrow \mathbb{R}$.

Definition A.22. Let $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$ be a differentiable function. The vector field corresponding to the 1 -form $d f$ is called the gradient of $f$ :

$$
\operatorname{grad}(f):=\nabla f:=\left(D_{1} f, \ldots, D_{n} f\right)
$$

The function $f$ (or sometimes $-f$ ) is called a potential function for the vector field.
In this case, Stokes' theorem says

$$
\int_{C} d f=\int_{\partial C} f .
$$

The left-hand side is by definition the flow of $\nabla f$ along $C$, and the right-hand side is $f(C(1))-f(C(0))$, the change in potential. We get the following classical result:

Theorem A.23. The flow of the gradient vector field $\nabla f$ along $C$ is given by the change in potential:

$$
\int_{C} \nabla f \cdot d C=f(C(1))-f(C(0))
$$

Geometric interpretation of the gradient. To understand the gradient, let $p \in \mathbb{R}^{n}$, and let $v \in \mathbb{R}^{n}$ be a unit vector. Define

$$
\begin{aligned}
C_{\varepsilon}:[0,1] & \rightarrow \mathbb{R}^{n} \\
t & \mapsto p+t \varepsilon v
\end{aligned}
$$

Let $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$. If $\varepsilon$ is small, then $\nabla f$ is approximately $\nabla f(p)$ along $C_{\varepsilon}$. Hence, the flow of $\nabla f$ along $C_{\varepsilon}$ is approximately $\varepsilon \nabla f(p) \cdot v$ (noting that $v$ is the unit tangent for $\left.C_{\varepsilon}\right)$. By Stokes' theorem, the flow is given by $f(C(1))-f(C(0))$. Hence,

$$
\nabla f(p) \cdot v \approx \frac{f(p+\varepsilon v)-f(p)}{\varepsilon}
$$

It turns out that taking the limit as $\varepsilon \rightarrow 0$ gives $\nabla f(p) \cdot v$ exactly. Therefore, the component of the gradient in any particular direction gives the rate of change of the function in that direction. Of course, the component is maximized when the direction points the same way as the gradient; so the gradient points in the direction of quickest increase of the function. In this way, the gradient can be thought of as "change density," and Stokes' theorem says, roughly, that the integral of change density gives the total change, i.e., the change in potential.

Fundamental Theorem in one variable. Specializing further, let $a, b$ be real numbers with $a<b$, and define $C(t)=a+t(b-a)$. Let $\omega=f: \mathbb{R} \rightarrow \mathbb{R}$ be a 0 -form. Stokes' theorem says that $\int_{C} d f=\int_{\partial C} f$. For the left-hand side, we get

$$
\int_{C} d f=\int_{C} f^{\prime} d x=\int_{0}^{1}\left(f^{\prime} \circ C\right) C^{\prime}=\int_{a}^{b} f^{\prime}
$$

using the change of variables theorem. Thus, Stokes' says

$$
\int_{a}^{b} f^{\prime}=\int_{\partial C} f=f(C(1))-f(C(0))=f(b)-f(a)
$$

the fundamental theorem.

$$
\text { Case } k=2
$$

To apply Stokes' theorem in the case $k=2$, we start with a 2 -chain and a 1-form. For the 2-chain, we will take a surface $S: D \rightarrow \mathbb{R}^{3}$, where $D=[0,1] \times[0,1]$. The 1form looks like $\omega=F_{1} d x+F_{2} d y+F_{3} d z$. Let $F:=\left(F_{1}, F_{2}, F_{3}\right)$ be the corresponding vector field. For Stokes' theorem, we need to consider $d \omega$. A straightforward calculation (do it!) yields
$d \omega=\left(D_{2} F_{3}-D_{3} F_{2}\right) d y \wedge d z-\left(D_{3} F_{1}-D_{1} F_{3}\right) d x \wedge d z+\left(D_{1} F_{2}-D_{2} F_{1}\right) d x \wedge d y$,
which corresponds to the vector field called the curl of $F$.

Definition A.24. Let $F=\left(F_{1}, F_{2}, F_{3}\right)$ be a vector field on $\mathbb{R}^{3}$. The curl of $F$ is the vector field
$\operatorname{curl}(F):=\nabla \times F:=\operatorname{det}\left[\begin{array}{ccc}\mathbf{i} & \mathbf{j} & \mathbf{k} \\ D_{1} & D_{2} & D_{3} \\ F_{1} & F_{2} & F_{3}\end{array}\right]:=\left(D_{2} F_{3}-D_{3} F_{2}, D_{3} F_{1}-D_{1} F_{3}, D_{1} F_{2}-D_{2} F_{1}\right)$.
where $D_{i}$ denotes the derivative with respect to the $i$-th variable.
Stokes' theorem says

$$
\int_{S} d \omega=\int_{\partial S} \omega
$$

By definition, the left-hand side is the flux of the $\operatorname{curl}(F)$ through $S$ and the righthand side is the flow of $F$ along the boundary. We get the classical result:

Theorem A.25. The flux of the curl of $F$ through the surface $S$ is equal to the flow of $F$ along the boundary of the surface:

$$
\int_{S} \operatorname{curl}(F) \cdot \vec{n}=\int_{\partial S} F \cdot d C
$$

where $C=\partial S$.
Geometric interpretation of the curl. To get a physical understanding of the curl, pick a point $p \in \mathbb{R}^{3}$ and a unit vector $v$. Let $D_{\varepsilon}$ be a parametrized disk of radius $\varepsilon$ centered at $p$ and lying in the plane normal to $v$. For $\varepsilon$ small, the curl of $F$ is approximately constant at $\operatorname{curl}(F)(p)$ over $D_{\varepsilon}$. The flux of the curl is,
thus, approximately the normal component, $\operatorname{curl}(F)(p) \cdot v$ times the area of $D_{\varepsilon}$. By Stokes' theorem, the flux is the circulation about the boundary. Thus,

$$
\operatorname{curl}(F)(p) \cdot v \approx \frac{1}{\operatorname{area}\left(D_{\varepsilon}\right)} \int_{C} F \cdot d C
$$

where $C=\partial D_{\varepsilon}$. It turns out that we get equality if we take the limit as $\varepsilon \rightarrow 0$. Hence, the component of the curl in the direction of $v$ measures the circulation of the original vector field about a point in the plane perpendicular to $v$. In this sense, the curl measures "circulation density". So roughly, Stokes' theorem says that integrating circulation density gives the total circulation.

## Case $k=3$

In this case, we are concerned with a 3 -chain and a 2 -form. We will consider the case where the chain is a solid $V: D \rightarrow \mathbb{R}^{3}$, where $D=[0,1] \times[0,1] \times[0,1]$. We will assume that $\operatorname{det} V^{\prime} \geq 0$ at all points. The 2 -form can be written

$$
\omega=F_{1} d y \wedge d z-F_{2} d x \wedge d z+F_{3} d x \wedge d y
$$

A simple calculation (do it!) yields

$$
d \omega=\left(D_{1} F_{1}+D_{2} F_{2}+D_{3} F_{3}\right) d x \wedge d y \wedge d z
$$

Definition A.26. If $F$ is a vector field on $\mathbb{R}^{n}$, the divergence of $F$ is the scalar function

$$
\operatorname{div}(F):=\nabla \cdot F:=\sum_{i=1}^{n} D_{i} F_{i}
$$

Stokes' theorem says that $\int_{V} d \omega=\int_{\partial V} \omega$. The left-hand side is

$$
\begin{aligned}
\int_{V} d \omega & =\int_{V} \operatorname{div}(F) d x \wedge d y \wedge d z \\
& =\int_{D} \operatorname{div}(F) \circ V \operatorname{det}\left(V^{\prime}\right) \\
& \left.=\int_{V} \operatorname{div}(F) \quad \text { (since we've assumed } \operatorname{det}\left(V^{\prime}\right) \geq 0\right)
\end{aligned}
$$

(The first equality follows from the definition of $\operatorname{div}(F)$, the second from the definition of integration of a differential form, and the third from the definition of a solid integral, given earlier in this handout.) The right-hand side of Stokes' is by definition the flux of $F$ through the boundary of $V$. The classical result is:

Definition A.27. The integral of the divergence of $F$ over $V$ is equal to the flux of $F$ through the boundary of $V$ :

$$
\int_{V} \operatorname{div}(F)=\int_{\partial V} F \cdot \vec{n}
$$

Geometric interpretation of the divergence. Pick a point $p \in \mathbb{R}^{3}$, and let $V_{\varepsilon}$ be a solid ball of radius $\varepsilon$ centered at $p$. If $\varepsilon$ is small, the divergence of $F$ will not
change much from $\operatorname{div}(F)(p)$ on $V_{\varepsilon}$. Hence, the integral of the divergence will be approximately $\operatorname{div}(F)(p)$ times the volume of $V_{\varepsilon}$. By Stokes' we get

$$
\operatorname{div}(F)(p) \approx \frac{1}{\operatorname{vol}\left(V_{\varepsilon}\right)} \int_{S} F \cdot \vec{n}
$$

where $S=\partial V$. Taking a limit gives an equality. Thus, the divergence measures "flux density": the amount of flux per unit volume diverging from a given point. So Stokes' theorem in this case is saying that the integral of flux density gives the total flux.

## Appendix B. Topology

B.1. Topological spaces. This appendix is an extraction of the salient parts, for our notes, of An outline summary of basic point set topology, by Peter May. (Note: this reference's definition of a neighborhood differs slightly from ours-we do not insist that a neighborhood is open. See below.)

Definition B.1. A topology on a set $X$ is a collection of subsets $\tau$ of $X$ such that
(1) $\emptyset \in \tau$.
(2) $X \in \tau$.
(3) $\tau$ is closed under arbitrary unions.
(4) $\tau$ is closed under finite intersections.

The elements of $\tau$ are called the open sets for the topology and $(X, \tau)$ (or just $X$, if $\tau$ is understood from context) is called a topological space. A neighborhood of a point $x \in X$ is any set $N$ (not necessarily open) that contains an open set containing $x$. A subset of a topological space is closed if its complement is open. The closure, $\bar{A}$, of a subset $A \subseteq X$ is the intersection of all closed set containing $A$. The interior of $A$, denoted $A^{\circ}$ is the union of all open sets (in $X$ ) that are contained in $A$, and the boundary of $A$ is the set $\partial A:=A \backslash A^{\circ}$.

Definition B.2. A collection of subsets $\mathcal{B}$ of a set $X$ is a basis for a topology on $X$ if
(1) $\mathcal{B}$ covers $X$ : for each $x \in X$, there exists an element of $B \in \mathcal{B}$ containing $x$;
(2) if $x \in B^{\prime} \cap B^{\prime \prime}$ for some $B^{\prime}, B^{\prime \prime} \in \mathcal{B}$, then there exists $B \in \mathcal{B}$ such that $x \in$ $B \subseteq B^{\prime} \cap B^{\prime \prime}$.
The topology generated by a basis $\mathcal{B}$ consists of all unions of sets in $\mathcal{B}$.
Exercise B.3. Let $\tau$ be the topology on $X$ generated by a basis $\mathcal{B}$. Show that $U \in \tau$ if and only if for all $x \in U$, there exists $B \in \mathcal{B}$ such that $x \in B \subseteq U$.

Example B.4. The open ball in $\mathbb{R}^{n}$ of radius $r \in \mathbb{R}_{\geq 0}$ centered at $x \in \mathbb{R}^{n}$ is the set

$$
B_{r}(x):=\left\{y \in \mathbb{R}^{n}:|x-y|<r\right\} .
$$

Note that the empty set is an open ball (taking $r=0$ ). The open balls generate the standard topology on $\mathbb{R}^{n}$.

In order to do integration on manifolds, the following concept is important:

Definition B.5. A topological space $X$ is second countable if it has a countable basis, i.e., if it has a basis whose elements can be put in bijection with $\mathbb{N}$.

Definition B.6. A topology on a set $X$ is Hausdorff if for all points $x \neq y$ in $X$, there exist neighborhoods $U$ of $x$ and $V$ of $y$ such that $U \cap V=\emptyset$.

Exercise B.7. Let $X$ be a Hausdorff topological space. Show that $\{x\}$ is closed for each point $x \in X$.

Definition B.8. Let $X$ and $Y$ be topological spaces. The product topology on $X \times Y$ has basis consisting of the sets $U \times V$ where $U$ is open in $X$ and $V$ is open in $Y$. Unless otherwise indicated, we will always assume the product topology on $X \times Y$.

Exercise B.9. Show that a topological space $X$ is Hausdorff if and only if the diagonal $\Delta:=\{(x, x) \in X \times X: x \in X\}$ is closed.

Definition B.10. The subspace topology on a subset $A$ of a topological space $X$ is the set of intersections $A \cap U$ such that $U$ is open in $X$. A subset of $X$ endowed with the subspace topology is called a (topological) subspace of $X$.

Exercise B.11. Let $X$ and $Y$ be topological spaces with subsets $A \subseteq X$ and $B \subseteq$ $Y$. Define two topologies on $A \times B$ : on one hand, endow $A$ and $B$ with their respective subspace topologies, and then form the product topology on $A \times B$; and on the other hand, give $A \times B$ the subspace topology as a subset of $X \times Y$. Show these topologies are the same.

Definition B.12. Let $X$ be a topological space, and let $\pi: X \rightarrow Y$ be a surjective function. The quotient topology on $Y$ is the set of subsets $U \subseteq Y$ such that $\pi^{-1}(U)$ is open in $X$.

Exercise B.13. Show that a subspace of a Hausdorff space is Hausdorff and that the product of Hausdorff spaces is Hausdorff. Show that the quotient of a Hausdorff space need not be Hausdorff.
B.2. Continuous functions and homeomorphisms. Let $f: X \rightarrow Y$ be a function between topological spaces. Then $f: X \rightarrow Y$ is continuous if $f^{-1}(U)$ is open in $X$ for each open set $U$ in $Y$. To check continuity, it suffices to check that $f^{-1}(U)$ is open for each set $U$ in a basis for $Y$, or to check that $f^{-1}(A)$ for each closed subset $A$ of $Y$.

## Exercise B.14.

(1) Show that $f: X \rightarrow Y$ is continuous if and only if for each $x \in X$ and neighborhood $V$ of $f(x)$, there exists a neighborhood $U$ of $x$ such that $f(U) \subseteq V$.
(2) Show that a function $f: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ is continuous if and only if it satisfies the usual $\epsilon-\delta$ definition of continuity.

If $f: X \rightarrow Y$ is bijective and both it and its inverse, $f^{-1}$ are continuous, then $f$ is a homeomorphism. In that case, we think of $X$ and $Y$ as being "isomorphic" or "the same" as topological spaces.
B.3. Connectedness. Let $I=[0,1]$ be the unit interval in $\mathbb{R}$, and let $X$ be a topological space. A path in $X$ is a continuous function $f: I \rightarrow X$ connecting the points $f(0)$ and $f(1)$.

## Definition B.15.

(1) $X$ is connected if the only subsets of $X$ that are both open and closed are $\emptyset$ and $X$, itself.
(2) $X$ is path connected if every two points of $X$ can be connected by a path.

Exercise B.16. Let $h: X \rightarrow Y$ be a continuous mapping, and let $A \subseteq X$ be a connected subspace of $X$. Show that $h(A)$ is connected.

Definition B.17. Define two equivalence relations on a topological space $X$. First, say $x \sim y$ for two points in $X$ if $x$ and $y$ are both elements in some connected subspace of $X$. Second, say $x \approx y$ if there is a path in $X$ from $x$ to $y$. An equivalence class for $\sim$ is a connected component of $X$, and an equivalence class for $\approx$ is a path component of $X$.

If $x, y$ are in some path component of a topological space $X$, let $f$ be a path in $X$ joining $x$ to $y$. Then the image of $f$, with the subspace topology, is connected. Hence, each connected component is the union of path components. It is not necessarily true that the set of connected components is the same as the set of path components, but we will see below (Theorem B.19) that in the case of relevance to our study-locally Euclidean spaces - the two notions coincide.

Definition B.18. Let $X$ be a topological space.
(1) $X$ is locally connected if for each $x \in X$ and each neighborhood $U$ of $x$, there exists a connected neighborhood $V$ of $x$ contained in $U$.
(2) $X$ is locally path connected if for each $x \in X$ and each neighborhood $U$ of $x$, there exists a path connected neighborhood $V$ of $x$ contained in $U$.

Theorem B.19. If a topological space is locally path connected, then its components and path components coincide.

## B.4. Compactness.

Definition B.20. A topological space $X$ is compact if every open cover of $X$ has a finite subcover, that is: if $\left\{U_{\alpha}\right\}_{\alpha}$ is a family of open subsets of $X$ such that $X=\cup_{\alpha} U_{\alpha}$, then there are finitely many indices $\alpha_{1}, \ldots, \alpha_{k}$ such that $X=\cup_{i=1}^{k} U_{\alpha_{i}}$.

Here is a list of first properties of compactness:

- A subspace $A$ of a topological space $X$ is compact if and only if every open cover of $A$ by open subsets of $X$ has a finite subcover.
- A subspace of $\mathbb{R}^{n}$ is compact if and only if it is closed and bounded.
- A compact subspace of a compact Hausdorff space is closed.
- The continuous image of a compact space is compact.
- If $f: X \rightarrow Y$ is a bijective continuous mapping, $X$ is compact, and $Y$ is Hausdorff, then $f$ is a homeomorphism, i.e., it has a continuous inverse.
B.5. Partition of unity. Let $M$ be a manifold that is second countable and Hausdorff, and let $\mathcal{U}$ be an open cover of $M$. Then there exists a collection $\Xi$ of smooth functions $M \rightarrow[0,1] \subset \mathbb{R}$ such that:
- For each $\lambda \in \Xi$ there exist $U \in \mathcal{U}$ containing $\operatorname{supp}(\lambda)$. [Note: The support of $\lambda$ is defined by

$$
\operatorname{supp}(\lambda):=\{x \in M: \lambda(x) \neq 0\} .]
$$

- For each $x \in M$, there exists an open neigborhood $V$ of $x$ such that $\lambda \mid V=0$ for all but finitely many $\lambda \in \Xi$, and

$$
\sum_{\lambda \in \Xi} \lambda(x)=1
$$

[Note: the sum makes sense since all but finitely terms are zero.]
Partitions of unity are important to glue together locally-defined objects. For instance, integration of forms on manifolds is at first defined locally, then glued together for a global definition. For more on the existence of partitions of unity, see the Wikipedia page on paracompact spaces.

## Appendix C. Measure Theory

For more details on measure theory, see The Elements of Integration and Lebesgue Measure, by Bartle, [1], or Math 321, Real Analysis, by Perkinson.

To put a "measure" on a set, we first need to divvy up the set into measurable pieces:

Definition C.1. A $\sigma$-algebra (sigma algebra) on a set $X$ is a collection of subsets $\Sigma \subseteq 2^{X}$ satisfying:
(1) $\emptyset \in \Sigma$ and $X \in \Sigma$,
(2) (closed under complementation) $A \in \Sigma$ implies $A^{c} \in \Sigma$,
(3) (closed under countable unions) $\left\{A_{i}\right\}_{i \in N} \subseteq \Sigma$ implies $\cup_{i \in N} A_{i} \in \Sigma$.

Next, we decide the size of each piece:
Definition C.2. Let $X$ be a set with $\sigma$-algebra $\Sigma$. A measure on $(X, \sigma)$ is a function

$$
\mu: \Sigma \rightarrow \mathbb{R}_{\geq 0} \cup\{\infty\}
$$

such that
(1) $\mu(\emptyset)=0$,
(2) (additive) $\left.\mu\left(\cup_{i \in N} A_{i}\right)=\sum_{i \in \mathbb{N}} \mu\left(A_{i}\right)\right)$ for pairwise disjoint $A_{i} \in \Sigma$.

The triple $(X, \sigma, \mu)$ is called a measure space.
We can then define integration of nice functions on the set. We first define "nice":
Definition C.3. Let $(X, \sigma, \mu)$ be a measure space. A function $f: X \rightarrow \mathbb{R}$ is measurable if

$$
f^{-1}((a, \infty)):=\{x \in X: f(x)>a\} \in \Sigma
$$

for all $a \in \mathbb{R}$.

Then define the integral of a simple function:
Definition C.4. The characteristic function of a subset $A \subseteq X$ is the function defined by

$$
\chi_{A}(x):= \begin{cases}1 & \text { if } x \in A \\ 0 & \text { if } x \in X \backslash A\end{cases}
$$

If $A$ is measurable and $c \in \mathbb{R}$, then

$$
\int c \chi_{A}:=c \mu(A)
$$

A function $\phi: X \rightarrow \mathbb{R}$ is simple if it is a linear combination of characteristic functions: $\phi=\sum_{i=1}^{k} c_{i} \chi_{E_{i}}$ for some measurable sets $E_{1}, \ldots, E_{k}$. In that case, its integral is

$$
\int \phi:=\sum_{i=1}^{k} \int c_{i} \chi_{E_{i}}=\sum_{i=1}^{k} c_{i} \mu\left(E_{i}\right)
$$

Finally, we define the integral of a measurable function by approximating it with simple functions:

Definition C.5. Let $f$ be a measurable function on a measure space $(X, \sigma, \mu)$. If $f(x) \geq 0$ for all $x \in X$, define

$$
\int f:=\sup _{\phi}\left\{\int \phi\right\}
$$

where the sup is over all simple functions $\phi$ such that $0 \leq \phi(x) \leq f(x)$ for all $x \in X$ (see Figure 35).

If $f$ is not necessarily nonnegative, then write $f=f^{+}-f^{-}$where

$$
f^{+}(x):=\left\{\begin{array}{ll}
f(x) & \text { if } f(x) \geq 0 \\
0 & \text { otherwise }
\end{array} \quad \text { and } \quad f^{-}(x):= \begin{cases}-f(x) & \text { if } f(x) \leq 0 \\
0 & \text { otherwise }\end{cases}\right.
$$

Then $f$ is integrable if $\int f^{+}$and $\int f^{-}$are finite, in which case

$$
\int f:=\int f^{+}-\int f^{-}
$$



Figure 35. Approximating a function with a simple function.
Lebesgue measure. We now turn to the case of interest for this text: the standard measure on $\mathbb{R}^{n}$. To approximate the size of a set $X \subset \mathbb{R}^{n}$, we cover it with rectangles:

Definition C.6. If $I \in \mathbb{R}^{n}$ is a rectangle, define $\mu(I)$ to be the product of the lengths of its sides. The outer measure of $X \subset \mathbb{R}^{n}$ is

$$
\operatorname{outer}(X)=\inf \left\{\sum_{i=1} \mu\left(I_{i}\right)\right\}
$$

where the inf is over sequence of rectangles $I_{1}, I_{2}, \ldots$ covering $X$, i.e., such that $X \subseteq \cup_{i=1}^{k} I_{i}$. The set $X$ is Lebesgue measurable if it "splits additively in measure":

$$
\text { outer } X=\operatorname{outer}(X \cap A)+\operatorname{outer}\left(X \cap A^{c}\right)
$$

for all $A \subseteq \mathbb{R}^{n}$.
Theorem C.7. The collection of Lebesgue measurable sets $\mathcal{L}$ in $\mathbb{R}^{n}$ forms a $\sigma$ algebra. It contains all open sets (and, hence, all closed sets). Outer measure restricted to $\mathcal{L}$ forms is a measure, i.e., defining $\mu(A)=\operatorname{outer}(A)$ for all $A \in \mathcal{L}$, it follows that $\left(\mathbb{R}^{n}, \mathcal{L}, \mu\right)$ is a measure space.

## Appendix D. Simplicial homology

D.1. Exact sequences. A sequence of linear mappings of vector spaces

$$
V^{\prime} \xrightarrow{f} V \xrightarrow{g} V^{\prime \prime}
$$

is exact (or exact at V ) if $\operatorname{im}(f)=\operatorname{ker}(g)$. A short exact sequence of vector spaces is a sequence of linear mappings

$$
0 \rightarrow V^{\prime} \xrightarrow{f} V \xrightarrow{g} V^{\prime \prime} \rightarrow 0
$$

exact at $V^{\prime}, V$, and $V^{\prime \prime}$.
Exercise D.1. For a short exact sequence of $R$-modules as above, show that
(1) $f$ is injective;
(2) $g$ is surjective;
(3) $V^{\prime \prime}$ is isomorphic to coker $(f)$;
(4) $\operatorname{dim} V=\operatorname{dim} V^{\prime}+\operatorname{dim} V^{\prime \prime}$.

In general, a sequence of linear mappings

$$
\cdots \rightarrow V_{i} \rightarrow V_{i+1} \rightarrow \ldots
$$

is exact if it is exact at each $V_{i}$ (except the first and last, if they exist).
Consider a commutative diagram of linear mappings with exact rows

(By commutative, we mean $\phi \circ f=h \circ \phi^{\prime}$ and $\phi^{\prime \prime} \circ g=k \circ \phi$.)
The snake lemma says there is an exact sequence

$$
\operatorname{ker} \phi^{\prime} \rightarrow \operatorname{ker} \phi \rightarrow \operatorname{ker} \phi^{\prime \prime} \rightarrow \operatorname{coker} \phi^{\prime} \rightarrow \operatorname{coker} \phi \rightarrow \operatorname{coker} \phi^{\prime \prime}
$$

If $f$ is injective, then so is $\operatorname{ker} \phi^{\prime} \rightarrow \operatorname{ker} \phi$, and if $k$ is surjective, so is coker $\phi \rightarrow$ coker $\phi^{\prime \prime}$.

Exercise D.2. Prove the snake lemma.
A chain complex $C$ of $R$-modules is a collection of $R$-modules $\left\{C_{n}\right\}_{n}$ together with $R$-module maps $\left\{d_{n}: C_{n} \rightarrow C_{n-1}\right\}_{n}$ such that $d_{n-1} \circ d_{n}=0$, in other words, $\operatorname{im}\left(d_{n}\right) \subseteq \operatorname{ker}\left(d_{n-1}\right)$. Sometimes for convenience, subscripts for $d$ are often omitted and we simply write $d^{2}=0$.

Given chain complexes $C$ and $D$, a chain map $f$ from $C$ to $D$ is a collection of $R$-module maps $\left\{f_{n}: C_{n} \rightarrow D_{n}\right\}$ satisfying $f d=d f$, namely, the following diagram commutes:


A short exact sequence of chain complexes of $R$-modules is a sequence of chain maps $0 \rightarrow A \xrightarrow{f} B \xrightarrow{g} C \rightarrow 0$ such that at each level $n$, we have a short exact sequence $0 \rightarrow A_{n} \xrightarrow{f_{n}} B_{n} \xrightarrow{g_{n}} C_{n} \rightarrow 0$.

The $n$-th homology group is defined to be the quotient $H_{n}(C):=\operatorname{ker}\left(d_{n-1}\right) / \operatorname{im}\left(d_{n}\right)$. The homology groups measures how far away the chain complex is from being exact. If all of the homology groups are zero, then we say that the chain complex is exact.

Using tools from homological algbera, it can be shown that a short exact sequence $0 \rightarrow A \xrightarrow{f} B \xrightarrow{g} C \rightarrow 0$ of chain complexes induces the connecting homomorphisms $\partial: H_{n}(C) \rightarrow H_{n-1}(A)$ and a long exact sequence of homology groups:

$$
\cdots \xrightarrow{g} H_{n+1}(C) \xrightarrow{\partial} H_{n}(A) \xrightarrow{f} H_{n}(B) \xrightarrow{g} H_{n}(C) \xrightarrow{\partial} H_{n-1}(A) \xrightarrow{f} \cdots .
$$

## D.2. Simplicial complexes.

D.2.1. First definitions. An (abstract) simplicial complex $\Delta$ on a finite set $S$ is a collection of subsets of $S$, closed under the operation of taking subsets. The elements of a simplicial complex $\Delta$ are called faces. An element $\sigma \in \Delta$ of cardinality $i+1$ is called an $i$-dimensional face or an $i$-face of $\Delta$. The empty set, $\emptyset$, is the unique face of dimension -1 . Faces of dimension 0, i.e., elements of $S$, are vertices and faces of dimension 1 are edges.

The maximal faces under inclusion are called facets. To describe a simplicial complex, it is often convenient to simply list its facets - the other faces are exactly determined as subsets. The dimension of $\Delta$, denoted $\operatorname{dim}(\Delta)$, is defined to be the maximum of the dimensions of its faces. A simplicial complex is pure if each of its facets has dimension $\operatorname{dim}(\Delta)$.

Example D.3. If $G=(V, E)$ is a simple connected graph (undirected with no multiple edges or loops), then $G$ is the pure one-dimensional simplicial complex on $V$ with $E$ as its set of facets.

Example D.4. Figure 36 pictures a simplicial complex $\Delta$ on the set $[5]:=\{1,2,3,4,5\}$ :

$$
\Delta:=\{\emptyset, \overline{1}, \overline{2}, \overline{3}, \overline{4}, \overline{5}, \overline{12}, \overline{13}, \overline{23}, \overline{24}, \overline{34}, \overline{123}\}
$$

writing, for instance, $\overline{23}$ to represent the set $\{2,3\}$.


Figure 36. A 2-dimensional simplicial complex, $\Delta$.

The sets of faces of each dimension are:

$$
\begin{array}{ll}
F_{-1}=\{\emptyset\} & F_{0}=\{\overline{1}, \overline{2}, \overline{3}, \overline{4}, \overline{5}\} \\
F_{1}=\{\overline{12}, \overline{13}, \overline{23}, \overline{24}, \overline{34}\} & F_{2}=\{\overline{123}\}
\end{array}
$$

Its facets are $\overline{5}, \overline{24}, \overline{34}$, and $\overline{123}$. The dimension of $\Delta$ is 2 , as determined by the facet $\overline{123}$. Since not all of the facets have the same dimension, $\Delta$ is not pure.
D.2.2. Simplicial homology. Let $\Delta$ be an arbitrary simplicial complex. By relabeling, if necessary, assume its vertices are $[n]:=\{1, \ldots, n\}$. For each $i$, let $F_{i}(\Delta)$ be the set of faces of dimension $i$, and define the group of $i$-chains to be the $\mathbb{R}$-vector space with basis $F_{i}(\Delta)$ :

$$
C_{i}=C_{i}(\Delta):=\mathbb{R} F_{i}(\Delta):=\left\{\sum_{\sigma \in F_{i}(\Delta)} a_{\sigma} \sigma: a_{\sigma} \in \mathbb{R}\right\}
$$

The boundary of $\sigma \in F_{i}(\Delta)$ is

$$
\partial_{i}(\sigma):=\sum_{j \in \sigma} \operatorname{sign}(j, \sigma)(\sigma \backslash j),
$$

where $\operatorname{sign}(j, \sigma)=(-1)^{k-1}$ if $j$ is the $k$-th element of $\sigma$ when the elements of $\sigma$ are listed in order, and $\sigma \backslash j:=\sigma \backslash\{j\}$. Extending linearly gives the $i$-th boundary mapping,

$$
\partial_{i}: C_{i}(\Delta) \rightarrow C_{i-1}(\Delta)
$$

If $i>n-1$ or $i<-1$, then $C_{i}(\Delta):=0$, and we define $\partial_{i}:=0$. We sometimes simply write $\partial$ for $\partial_{i}$ if the dimension $i$ is clear from context.

Example D.5. Suppose $\sigma=\{1,3,4\}=\overline{134} \in \Delta$. Then $\sigma \in F_{2}(\Delta)$, and

$$
\operatorname{sign}(1, \sigma)=1, \quad \operatorname{sign}(3, \sigma)=-1, \quad \operatorname{sign}(4, \sigma)=1
$$

Therefore,

$$
\partial(\sigma)=\partial_{2}(\overline{134})=\overline{34}-\overline{14}+\overline{13}
$$

The (augmented) chain complex of $\Delta$ is the complex

$$
0 \longrightarrow C_{n-1}(\Delta) \xrightarrow{\partial_{n-1}} \cdots \stackrel{\partial_{2}}{\longrightarrow} C_{1}(\Delta) \xrightarrow{\partial_{1}} C_{0}(\Delta) \xrightarrow{\partial_{0}} C_{-1}(\Delta) \longrightarrow 0 .
$$

The word complex here refers to the fact that $\partial^{2}:=\partial \circ \partial=0$, i.e., for each $i$, we have $\partial_{i-1} \circ \partial_{i}=0$.


Figure 37. Two boundary mapping examples. Notation: if $i<j$, then we write $i \longleftrightarrow j$ for $\overline{i j}$ and $i \longleftrightarrow j$ for $-\overline{i j}$.

Figure 37 gives two examples of the application of a boundary mapping. Note that

$$
\partial^{2}(\overline{12})=\partial_{0}\left(\partial_{1}(\overline{12})\right)=\partial_{0}(\overline{2}-\overline{1})=\emptyset-\emptyset=0 .
$$

The reader is invited to verify $\partial^{2}(\overline{123})=0$.

$\overline{1234}$


$$
\longmapsto
$$


$\overline{234}-\overline{134}+\overline{124}-\overline{123}$

Figure 38. $\partial_{3}$ for a solid tetrahedron. Notation: if $i<j<k$,


Figure 38 shows the boundary of $\sigma=\overline{1234}$, the solid tetrahedron. Figure 39 helps to visualize the fact that $\partial^{2}(\sigma)=0$. The orientations of the triangles may be thought of as inducing a "flow" along the edges of the triangles. These flows cancel to give a net flow of 0 . This should remind you of Stokes' theorem from multivariable calculus.

Example D.6. Let $\Delta$ be the simplicial complex on [4] with facets $\overline{12}, \overline{3}$, and $\overline{4}$ pictured in Figure 40. The faces of each dimension are:

$$
F_{-1}(\Delta)=\{\emptyset\}, \quad F_{0}(\Delta)=\{\overline{1}, \overline{2}, \overline{3}, \overline{4}\}, \quad F_{1}(\Delta)=\{\overline{12}\} .
$$



Figure 39. As seen in Figure 38, the boundary of a solid tetrahedron consists of oriented triangular facets.


Figure 40. Simplicial complex for Example D.6.

Here is the chain complex for $\Delta$ :


In terms of matrices, the chain complex is given by


The sequence is not exact since $\operatorname{dim}\left(\operatorname{im} \partial_{1}\right)=\operatorname{dim} \partial_{1}=1$, whereas by rank-nullity, $\operatorname{dim}\left(\operatorname{ker}\left(\partial_{0}\right)\right)=4-\operatorname{dim} \partial_{0}=3$.

Definition D.7. For $i \in \mathbb{Z}$, the $i$-th (reduced) homology of $\Delta$ is the vector space

$$
\widetilde{H}_{i}(\Delta):=\operatorname{ker} \partial_{i} / \operatorname{im} \partial_{i+1}
$$

In particular, $\widetilde{H}_{n-1}(\Delta)=\operatorname{ker}\left(\partial_{n-1}\right)$, and $\widetilde{H}_{i}(\Delta)=0$ for $i>n-1$ or $i<0$. Elements of ker $\partial_{i}$ are called $i$-cycles and elements of im $\partial_{i+1}$ are called $i$-boundaries. The $i$-th (reduced) Betti number of $\Delta$ is the dimension of the $i$-th homology:

$$
\tilde{\beta}_{i}(\Delta):=\operatorname{dim} \widetilde{H}_{i}(\Delta)=\operatorname{dim}\left(\operatorname{ker} \partial_{i}\right)-\operatorname{dim}\left(\partial_{i+1}\right)
$$

Remark D.8. To define ordinary (non-reduced) homology groups, $H_{i}(\Delta)$, and Betti numbers $\beta_{i}(\Delta)$, modify the chain complex by replacing $C_{-1}(\Delta)$ with 0 and $\partial_{0}$ with
the zero mapping. The difference between homology and reduced homology is that $H_{0}(\Delta) \simeq \mathbb{R} \oplus \widetilde{H}_{0}(\Delta)$ and, thus, $\beta_{0}(\Delta)=\tilde{\beta}_{0}(\Delta)+1$. All other homology groups and Betti numbers coincide. From now on, we use "homology" to mean reduced homology.

In general, homology can be thought of as a measure of how close the chain complex is to being exact. In particular, $\widetilde{H}_{i}(\Delta)=0$ for all $i$ if and only if the chain complex for $\Delta$ is exact. For the next several examples, we will explore how exactness relates to the topology of $\Delta$.

The 0-th homology group measures "connectedness". Write $i \sim j$ for vertices $i$ and $j$ in a simplicial complex $\Delta$ if $\overline{i j} \in \Delta$. An equivalence class under the transitive closure of $\sim$ is a connected component of $\Delta$.

Exercise D.9. Show that $\tilde{\beta}_{0}(\Delta)$ is one less than the number of connected components of $\Delta$.

For instance, for the simplicial complex $\Delta$ in Example D.6,

$$
\tilde{\beta}_{0}(\Delta)=\operatorname{dim} \widetilde{H}_{0}(\Delta)=\operatorname{dim}\left(\operatorname{ker} \partial_{0}\right)-\operatorname{dim}\left(\partial_{1}\right)=3-1=2
$$

Example D.10. The hollow triangle,

$$
\Delta=\{\emptyset, \overline{1}, \overline{2}, \overline{3}, \overline{12}, \overline{13}, \overline{23}\}
$$


has chain complex

$$
\left.0 \longrightarrow \mathbb{R}^{3} \xrightarrow{\overline{3}\left(\begin{array}{ccc}
\overline{12} & \overline{13} & \overline{23} \\
\overline{1} & -1 & 0 \\
1 & 0 & -1 \\
0 & 1 & 1
\end{array}\right)} \mathbb{R}^{3} \xrightarrow{\bar{y}} \begin{array}{ccc}
\overline{1} & \overline{2} & \overline{3} \\
1 & 1 & 1
\end{array}\right)
$$

It is easy to see that $\operatorname{dim}\left(\partial_{1}\right)=\operatorname{dim}\left(\operatorname{ker} \partial_{0}\right)=2$. It follows that $\tilde{\beta}_{0}(\Delta)=0$, which could have been anticipated since $\Delta$ is connected. Since $\operatorname{dim}\left(\partial_{1}\right)=2$, rank-nullity says $\operatorname{dim}\left(\operatorname{ker} \partial_{1}\right)=1$, whereas $\partial_{2}=0$. Therefore, $\tilde{\beta}_{1}(\Delta)=\operatorname{dim}\left(\operatorname{ker} \partial_{1}\right)-\operatorname{dim}\left(\partial_{2}\right)=$ 1. In fact, $\widetilde{H}_{1}(\Delta)$ is generated by the 1-cycle


If we would add $\overline{123}$ to $\Delta$ to get a solid triangle, then the above cycle would be a boundary, and there would be no homology in any dimension. Similarly, a solid tetrahedron has no homology, and a hollow tetrahedron has homology only in dimension 2 (of rank 1 ).

Exercise D.11. Compute the Betti numbers for the simplicial complex formed by gluing two (hollow) triangles along an edge. Describe generators for the homology.

Example D.12. Consider the simplicial complex pictured in Figure 41 with facets $\overline{14}, \overline{24}, \overline{34}, \overline{123}$. It consists of a solid triangular base whose vertices are connected by edges to the vertex 4 . The three triangular walls incident on the base are hollow.


Figure 41. Simplicial complex for Example D.12.

What are the Betti numbers? The chain complex is:

By inspection, $\operatorname{dim}\left(\partial_{2}\right)=1$ and $\operatorname{dim}\left(\partial_{1}\right)=\operatorname{dim}\left(\right.$ ker $\left.\partial_{0}\right)=3$. Rank-nullity gives $\operatorname{dim}\left(\operatorname{ker} \partial_{1}\right)=6-3=3$. Therefore, $\tilde{\beta}_{0}=\tilde{\beta}_{2}=0$ and $\tilde{\beta}_{1}=2$. It is not surprising that $\tilde{\beta}_{0}=0$, since $\Delta$ is connected. Also, the fact that $\tilde{\beta}_{2}=0$ is easy to see since $\overline{123}$ is the only face of dimension 2 , and its boundary is not zero. Seeing that $\tilde{\beta}_{1}=2$ is a little harder. Given the cycles corresponding to the three hollow triangles incident on vertex 4 , one might suppose $\tilde{\beta}_{1}=3$. However, as conveyed in Figure 42, those cycles are not independent: if properly oriented their sum is the boundary of the solid triangle, $\overline{123}$; hence, their sum is 0 in the first homology group.
D.2.3. A quick aside on algebraic topology. Algebraic topology seeks an assignment of the form $X \mapsto \alpha(X)$ where $X$ is a topological space and $\alpha(X)$ is some algebraic invariant (a group, ring, etc.). If $X \simeq Y$ as topological spaces, i.e., if $X$ and $Y$ are homeomorphic, then we should have $\alpha(X) \simeq \alpha(Y)$ as algebraic objects-this is what it means to be invariant. The simplicial homology we have developed provides the tool for creating one such invariant.

Let $X$ be a 2-torus - the surface of a donut. Draw triangles on the surface so that neighboring triangles meet vertex-to-vertex or edge-to-edge. The triangulation is naturally interpreted as a simplicial complex $\Delta$. An amazing fact, of fundamental importance, is that the associated homology groups do not depend on the choice


Figure 42. A tetrahedron with solid base and hollow walls. Cycles around the walls sum to the boundary of the base, illustrating a dependence among the cycles in the first homology group.
of triangulation! In this way, we get an assignment

$$
X \mapsto \widetilde{H}_{i}(X):=\widetilde{H}_{i}(\Delta),
$$

and, hence, also $X \mapsto \tilde{\beta}_{i}(X):=\tilde{\beta}_{i}(\Delta)$, for all $i$.
In a course on algebraic topology, one learns that these homology groups do not see certain aspects of a space. For instance, they do not change under certain contraction operations. A line segment can be continuously morphed into a single point, and the same goes for a solid triangle or tetrahedron. So these spaces all have the homology of a point - in other words: none at all (all homology groups are trivial). A tree is similarly contractible to a point, so the addition of a tree to a space has no effect on homology. Imagine the tent with missing walls depicted in Figure 41. Contracting the base to a point leaves two vertices connected by three line segments. Contracting one of these line segments produces two loops meeting at a single vertex. No further significant contraction is possible - we are not allowed to contract around "holes" (of any dimension). These two loops account for $\tilde{\beta}_{1}=2$ in our previous calculation. As another example, imagine a hollow tetrahedron. Contracting a facet yields a surface that is essentially a sphere with three longitudinal lines connecting its poles, thus dividing the sphere into 3 regions. Contracting two of these regions results in a sphere - a bubble - with a single vertex drawn on it. No further collapse is possible. This bubble accounts for the fact that $\tilde{\beta}_{2}=1$ is the only nonzero Betti number for the sphere. (Exercise: verify that $\tilde{\beta}_{2}=1$ in this case.)
D.2.4. More examples. Let $\Delta \subset 2^{[n]}$ be a $d$-dimensional simplicial complex. For each $i \in \mathbb{Z}$ we have the space of $i$-dimensional chains $C_{i}:=\mathbb{R} F_{i}$, the vector space of formal sums of $i$-dimensional faces of $\Delta$. We have $C_{i}=0$ for $i<-1$ and $i>d$, and since $F_{-1}=\{\emptyset\}$, we have $C_{-1}=\mathbb{R}$. The boundary mapping $\partial_{i}: C_{i} \rightarrow C_{i-1}$
is defined as follows: if $\sigma=\overline{\sigma_{1} \ldots \sigma_{i+1}} \in F_{i}$ where the $\sigma_{k}$ are the vertices of $\sigma$ and $\sigma_{1}<\cdots<\sigma_{i+1}$, then

$$
\begin{aligned}
\partial_{i}(\sigma) & =\sum_{k=1}^{i+1}(-1)^{k-1} \overline{\sigma_{1} \ldots \widehat{\sigma_{k}} \ldots \sigma_{i+1}} \\
& =\overline{\sigma_{2} \sigma_{3} \sigma_{4} \ldots \sigma_{i+1}}-\overline{\sigma_{1} \sigma_{3} \sigma_{4} \ldots \sigma_{i+1}}+\overline{\sigma_{1} \sigma_{2} \sigma_{4} \ldots \sigma_{i+1}}+\cdots .
\end{aligned}
$$

for $-1 \leq 1 \leq d$ and $\partial_{i}=0$, otherwise. Recall the definitions of the reduced homology groups and Betti numbers:

$$
\begin{aligned}
\widetilde{H}_{i} & :=\widetilde{H}_{i}(\Delta):=\operatorname{ker} \partial_{i} / \operatorname{im} \partial_{i+1} \\
\tilde{\beta}_{i} & :=\tilde{\beta}_{i}(\Delta):=\operatorname{dim} \widetilde{H}_{i}=\operatorname{nullity}\left(\partial_{i}\right)-\operatorname{dim}\left(\partial_{i+1}\right)
\end{aligned}
$$

## Examples D.13.

I.1.

We have $\operatorname{dim}\left(\partial_{0}\right)=1$, and hence by the rank-nullity theorem, nullity $=4-1=3$. The only non-vanishing homology group is

$$
\begin{aligned}
\widetilde{H}_{0}(\Delta) & =\operatorname{ker} \partial_{0} / \operatorname{im} \partial_{1}=\operatorname{ker} \partial_{0}=\operatorname{Span}\{\overline{2}-\overline{1}, \overline{3}-\overline{1}, \overline{4}-\overline{1}\} \\
\tilde{\beta}_{0} & =3
\end{aligned}
$$

Question D.14. What are the homology groups and Betti numbers for $\Delta=\{\overline{1}, \ldots, \bar{n}\}$ for general $n \geq 1$ ?
I.2.

$$
\begin{aligned}
& \begin{array}{llll}
\overline{5} & \overline{6} & & \\
\circ & \bullet & & \\
& & & \\
\hline & \circ & \bullet & \bullet \\
\overline{1} & \overline{2} & \overline{3} & \overline{4}
\end{array} \\
& 0 \rightarrow \mathbb{R}^{2} \xrightarrow{\left(\begin{array}{rr}
-1 & 0 \\
0 & -1 \\
0 & 0 \\
0 & 0 \\
1 & 0 \\
0 & 1
\end{array}\right)} \mathbb{R}^{6} \xrightarrow{\left(\begin{array}{llllll}
1 & 1 & 1 & 1 & 1 & 1
\end{array}\right)} \mathbb{R} \rightarrow 0
\end{aligned}
$$

We have

$$
\begin{array}{r}
\operatorname{dim}\left(\partial_{0}\right)=1, \text { nullity }\left(\partial_{0}\right)=6-1=5 \\
\operatorname{dim}\left(\partial_{1}\right)=2, \text { nullity }\left(\partial_{1}\right)=0
\end{array}
$$

Therefore, $\tilde{\beta}_{0}=5-2=3$ and $\tilde{\beta}_{1}=0$. The same as in example I.
Homology:

$$
\begin{aligned}
\widetilde{H}_{0}=\operatorname{ker} \partial_{0} / \operatorname{im} \partial_{1} & =\operatorname{Span}\{\overline{2}-\overline{1}, \overline{3}-\overline{1}, \overline{4}-\overline{1}, \overline{5}-\overline{1}, \overline{6}-\overline{1}\} / \operatorname{Span}\{\overline{5}-\overline{1}, \overline{6}-\overline{2}\} \\
& \simeq \operatorname{Span}\{\overline{2}-\overline{1}, \overline{3}-\overline{1}, \overline{4}-\overline{1}\}
\end{aligned}
$$

In $\widetilde{H}_{0}$, we have $\overline{5}=\overline{1}$ and $\overline{6}=\overline{2}$, which means, $\overline{5}-\overline{1}=0$ and $\overline{6}-\overline{1}=\overline{2}-\overline{1}$.
Question D.15. How does this example generalize?
II. 1

$$
0 \rightarrow \mathbb{R}^{3} \xrightarrow{\left(\begin{array}{rrr}
-1 & -1 & 0 \\
1 & 0 & -1 \\
0 & 1 & 1
\end{array}\right)} \mathbb{R}^{3} \xrightarrow[\partial_{0}]{\left(\begin{array}{lll}
1 & 1 & 1
\end{array}\right)} \mathbb{R} \rightarrow 0
$$

We have

$$
\begin{gathered}
\operatorname{dim}\left(\partial_{0}\right)=1, \operatorname{nullity}\left(\partial_{0}\right)=3-1=2 \\
\operatorname{dim}\left(\partial_{1}\right)=2, \operatorname{nullity}\left(\partial_{1}\right)=3-2=1
\end{gathered}
$$

Therefore, $\tilde{\beta}_{0}=2-2=0$ and $\tilde{\beta}_{1}=1$.
Homology:

$$
\widetilde{H}_{1}=\operatorname{ker} \partial_{1} / \operatorname{im} \partial_{2}=\operatorname{ker} \partial_{1}=\operatorname{Span}\{\overline{23}-\overline{13}+\overline{12}\}
$$

A picture of the generator for $\widetilde{H}_{1}$ :

II.2.


$$
0 \rightarrow \mathbb{R}^{5} \xrightarrow{\left(\begin{array}{rrrrr}
-1 & 0 & 0 & 0 & -1 \\
1 & -1 & 0 & 0 & 0 \\
0 & 1 & -1 & 0 & 0 \\
0 & 0 & 1 & -1 & 0 \\
0 & 0 & 0 & 1 & 1
\end{array}\right)} \mathbb{R}^{3} \xrightarrow{\left(\begin{array}{lllll}
1 & 1 & 1 & 1 & 1
\end{array}\right)} \mathbb{R} \rightarrow 0
$$

We have

$$
\begin{gathered}
\operatorname{dim}\left(\partial_{0}\right)=1, \operatorname{nullity}\left(\partial_{0}\right)=5-1=4 \\
\operatorname{dim}\left(\partial_{1}\right)=4, \operatorname{nullity}\left(\partial_{1}\right)=5-4=1
\end{gathered}
$$

Therefore, $\tilde{\beta}_{0}=4-4=0$ and $\tilde{\beta}_{1}=1$.
Homology:

$$
\widetilde{H}_{1}=\operatorname{ker} \partial_{1} / \operatorname{im} \partial_{2}=\operatorname{ker} \partial_{1}=\operatorname{Span}\{\overline{12}+\overline{23}+\overline{34}+\overline{45}-\overline{15}\}
$$

The first homology is generated by a cycle of edges.
Question D.16. What happens in homology if we start with the triangle and subdivide its edges arbitrarily?

## III.1.



Let's compute the first homology.

$$
0 \underset{\partial_{2}}{\longrightarrow} \mathbb{R}^{7} \xrightarrow{\left(\begin{array}{rrrrrrr}
-1 & -1 & -1 & 0 & 0 & 0 & 0 \\
1 & 0 & 0 & -1 & -1 & 0 & 0 \\
0 & 1 & 0 & 1 & 0 & -1 & -1 \\
0 & 0 & 1 & 0 & 0 & 1 & 0 \\
0 & 0 & 0 & 0 & 1 & 0 & 1
\end{array}\right)} \mathbb{R}^{\left(\begin{array}{l}
1
\end{array}\right.}
$$

We have $\operatorname{dim} \partial_{1}=4$, and so nullity $\partial_{1}=7-4=3$. Therefore $\tilde{\beta}_{1}=3$. The homology is generated by the cycles surrounding the three bounded faces of the complex as drawn above:

$$
\widetilde{H}_{1}=\operatorname{ker} \partial_{1}=\operatorname{Span}\{\overline{23}-\overline{13}+\overline{12}, \overline{34}-\overline{14}+\overline{13}, \overline{35}-\overline{25}+\overline{23}\}
$$

III.2.


For first homology, note that $\partial_{2}(\overline{134})=\overline{34}-\overline{14}+\overline{13}$. We get

$$
\mathbb{R} \xrightarrow[\partial_{2}]{\left(\begin{array}{r}
0 \\
1 \\
-1 \\
0 \\
0 \\
1 \\
0
\end{array}\right)} \mathbb{R}^{7} \xrightarrow{\left(\begin{array}{rrrrrrr}
-1 & -1 & -1 & 0 & 0 & 0 & 0 \\
1 & 0 & 0 & -1 & -1 & 0 & 0 \\
0 & 1 & 0 & 1 & 0 & -1 & -1 \\
0 & 0 & 1 & 0 & 0 & 1 & 0 \\
0 & 0 & 0 & 0 & 1 & 0 & 1
\end{array}\right)} \mathbb{R}^{5}
$$

We have nullity $\left(\partial_{1}\right)=7-4=3$ and $\operatorname{dim} \partial_{2}=1$. Therefore, $\tilde{\beta}_{1}=3-1=2$. The homology is generated by the cycles surrounding the two unfilled bounded faces of the complex:

$$
\begin{aligned}
\widetilde{H}_{1}=\operatorname{ker} \partial_{1} \operatorname{im} \partial_{2} & =\operatorname{Span}\{\overline{23}-\overline{13}+\overline{12}, \overline{34}-\overline{14}+\overline{13}, \overline{35}-\overline{25}+\overline{23}\} / \operatorname{Span}\{\overline{34}-\overline{14}+\overline{13}\} \\
& \simeq \operatorname{Span}\{\overline{23}-\overline{13}+\overline{12}, \overline{35}-\overline{25}+\overline{23}\}
\end{aligned}
$$

The cycle $\overline{34}-\overline{14}+\overline{13}$ is the boundary of the shaded face, and thus has become 0 in the homology group.

Question D.17. How does this example generalize?

Exercise D.18. Draw the following simplicial complexes, determine their Betti numbers, and describe bases for their homology groups. Recall that a facet of a simplicial complex is a face that is maximal with respect to inclusion. So we can describe a simplicial complex by just listing its facets. The whole simplicial complex then consists of the facets and all of their subsets.
(1) $\Delta$ with facets $\overline{123}, \overline{24}, \overline{34}, \overline{45}, \overline{56}, \overline{57}, \overline{89}, \overline{10}$.
(2) $\Delta$ with facets $\overline{123}, \overline{14}, \overline{24}$, and $\overline{34}$.
(3) $\Delta$ with facets $\overline{123}, \overline{124}, \overline{134}, \overline{234}, \overline{125}, \overline{135}, \overline{235}$. (Two hollow tetrahedra glued along a face.)

## References

1. Robert G. Bartle, The elements of integration and Lebesgue measure, Wiley Classics Library, John Wiley \& Sons, Inc., New York, 1995, Containing a corrected reprint of the 1966 original [it The elements of integration, Wiley, New York; MR0200398 (34 \#293)], A Wiley-Interscience Publication. MR 1312157
2. Lawrence Conlon, Differentiable manifolds, second ed., Modern Birkhäuser Classics, Birkhäuser Boston, Inc., Boston, MA, 2008. MR 2413709
3. David A. Cox, The homogeneous coordinate ring of a toric variety, J. Algebraic Geom. 4 (1995), no. 1, 17-50.
4. V. I. Danilov, The geometry of toric varieties, Uspekhi Mat. Nauk 33 (1978), no. 2(200), 85-134, 247. MR 495499
5. William Fulton, Introduction to toric varieties, Annals of Mathematics Studies, vol. 131, Princeton University Press, Princeton, NJ, 1993, The William H. Roever Lectures in Geometry. MR 1234037
6. Allen Hatcher, Algebraic topology, Cambridge University Press, Cambridge, 2002.
7. David Hilbert, Mathematical problems, Bull. of the Amer. Math. Society 8 (1902), 437-479.
8. Klaus Jänich, Vector analysis, Undergraduate Texts in Mathematics, Springer-Verlag, New York, 2001, Translated from the second German (1993) edition by Leslie Kay. MR 1811820
9. S. L. Kleiman and Dan Laksov, Schubert calculus, Amer. Math. Monthly 79 (1972), 10611082.
10. John M. Lee, Introduction to smooth manifolds, second ed., Graduate Texts in Mathematics, vol. 218, Springer, New York, 2013. MR 2954043
11. Jerry Shurman, Calculus and analysis in Euclidean space, Undergraduate Texts in Mathematics, Springer, Cham, 2016. MR 3586606
12. Loring W. Tu, An introduction to manifolds, second ed., Universitext, Springer, New York, 2011. MR 2723362
13. Frank W. Warner, Foundations of differentiable manifolds and Lie groups, Graduate Texts in Mathematics, vol. 94, Springer-Verlag, New York-Berlin, 1983, Corrected reprint of the 1971 edition. MR 722297

[^0]:    Date: April 19, 2021.

[^1]:    ${ }^{1}$ We will typically refer to these simply as $n$-manifolds.

[^2]:    2 "Nice" would mean that the mapping and its derivative are injective on $U$.
    ${ }^{3}$ These names are ad hoc. There is no standard terminology.

[^3]:    ${ }^{4}$ In the following, as always, we identify vectors in $\mathbb{R}^{n}$ with column matrices.

[^4]:    ${ }^{5}$ A chain would be a sequence of such mappings between abelian groups in which the indices on the spaces went down instead of up.

[^5]:    ${ }^{6}$ The terminoloy of cocycles and coboundaries is motivated by singular/simplicial homology theory, where these concepts have simple geometric interpretations.

[^6]:    ${ }^{7} \mathrm{An} \mathbb{R}$-algebra is a ring that is also naturally a vector space over $\mathbb{R}$.

[^7]:    ${ }^{8}$ To learn more about this operation, read about interior products in another text on manifolds.

[^8]:    ${ }^{9}$ I did not make up the name of this theorem.

[^9]:    ${ }^{10}$ For background on exact sequences, see Appendix D.

[^10]:    ${ }^{11}$ By hypersurface, we mean codimension one in the ambient space.

