## Week 1

## Conjugacy classes.

Throughout, $\mathbb{C}$ will be the field of complex numbers. Recall that $\mathbb{C}$ is algebraically closed and is equipped with a norm $|a|=a \bar{a}$ (here, $\bar{a}$ is the complex conjugate of the complex number a) making $\mathbb{C}$ into a topological space.

Let $V$ be a finite dimensional $\mathbb{C}$-vectorspace, say of dimension $d$, then after choosing a basis we can identify $V$ with the space of $d$-tuples $\mathbb{C}^{d}$. As such we can equip $V$ with the analytic topology induced by the metric $d(v, w)=|v-w|$ coming from the norm

$$
|v|=\max \left|v_{i}\right|
$$

for any $v=\left(v_{1}, \ldots, v_{d}\right) \in V=\mathbb{C}^{d}$. In this chapter we will be primarily interested in the analytic topology induced on closed subsets of some $V$.

With $G L_{n}$ we denote the group of all invertible $n \times n$ matrices $G L_{n}(\mathbb{C})$ with coefficients in $\mathbb{C}$. As an $n \times n$ matrix $A \in M_{n}(\mathbb{C})$ is invertible if and only if its determinant $\operatorname{det}(A)$ is non-zero, we see that $G L_{n}$ is a dense open subset of the $n^{2}$-dimensional vectorspace $M_{n}=M_{n}(\mathbb{C})$.

## 1a. Conjugacy classes of matrices.

An $n \times n$ matrix $A \in M_{n}$ is by left multiplication a linear operator on the $n$-dimensional vectorspace $V_{n}=\mathbb{C}^{n}$ of column vectors. If $g \in G L_{n}$ is the matrix describing the base change from the canonical basis of $V_{n}$ to a new basis, then the linear operator expressed in this new basis is represented by the matrix $g A g^{-1}$. For a given matrix $A$ we want to find an adapted basis such that the conjugated matrix $g A g^{-1}$ has a simple form.

That is, we consider the linear action of $G L_{n}$ on the $n^{2}$-dimensional vectorspace $M_{n}$ of $n \times n$ matrices determined by

$$
G L_{n} \times M_{n} \longrightarrow M_{n} \quad(g, A) \mapsto g \cdot A=g A g^{-1}
$$

The orbit $\mathcal{O}_{A}$ of $A$ under this action, that is the set of all matrices of the form $g A g^{-1}$ for some $g \in G L_{n}$, is called the conjugacy class of $A$. We look for a particularly nice representant in a given conjugacy class. That is, we want to solve the following orbit space problem.

## Problem 1.

Classify the conjugacy classes of $n \times n$ matrices.
With $\mathbb{1}_{n}$ we denote the identity matrix in $M_{n}$ and with $e_{i j}$ the matrix whose unique non-zero entry is 1 at entry $(i, j)$. Recall that the group $G L_{n}$ is generated by the following three classes of matrices :

- the permutation matrices $p_{i j}=\mathbb{1}_{n}+e_{i j}+e_{j i}-e_{i i}-e_{j j}$ for all $i \neq j$,
- the addition matrices $a_{i j}(\lambda)=\mathbb{1}_{n}+\lambda e_{i j}$ for all $i \neq j$ and $0 \neq \lambda$, and
- the multiplication matrices $m_{i}(\lambda)=\mathbb{1}_{n}+(\lambda-1) e_{i i}$ for all $i$ and $0 \neq \lambda$.

Conjugation by these matrices determine the three types of Jordan moves on $n \times n$ matrices, where the altered rows and columns are dashed :


Therefore, it suffices to consider sequences of these moves on a given $n \times n$ matrix $A \in M_{n}$. The characteristic polynomial of $A$ is defined to be the polynomial of degree $n$ in the variable $t$

$$
\chi_{A}(t)=\operatorname{det}\left(A-t \mathbb{1}_{n}\right) \in \mathbb{C}[t] .
$$

As $\mathbb{C}$ is algebraically closed, $\chi_{A}(t)$ decomposes as a product of linear terms $\prod_{i=1}^{e}\left(t-\lambda_{i}\right)^{d_{i}}$ where the $\left\{\lambda_{1}, \ldots, \lambda_{e}\right\}$ are called the eigenvalues of the matrix $A$. Observe that $\lambda_{i}$ is an eigenvalue of $A$ if and only if there is a non-zero eigenvector $v \in V_{n}=\mathbb{C}^{n}$ with eigenvalue $\lambda_{i}$, that is, $A . v=\lambda_{i} v$. In particular, the rank $r_{i}$ of the matrix $A_{i}=\lambda_{i} \rrbracket_{n}-A$ satisfies $n-d_{i} \leq r_{i}<n$.

We will apply the following reduction step to the matrices $A_{i}$. Let $B$ be a $n \times n$ matrix of rank $r$. Then, applying type p and type a Jordan moves we can conjugate $B$ to the following block form

where white blocks denote zero matrices, hence the rank of the (black) right hand side is equal to $r$. We separate two cases. First, assume that the square $r \times r$ bottom right matrix has rank $r$ (is invertible). Then, the rows of the upper right block are linear combinations of its rows. Then, we apply type a Jordan moves (which do not spoil the zero blocks) and conjugate the matrix to the block form


The top left zero block of size $n-r$ splits off and we stop the reduction. In the second case, assume the square $r \times r$ matrix has rank $s<r$ (hence the upper right block has rank $r-s \leq n-r$ ). Repeating the above reduction to the $r \times r$ block we obtain the situations :


The two bottom right blocks together have rank $s$, whence all rows of the two upper right blocks are linear combinations of them. Using type a Jordan moves (which preserve the zero blocks) we arrive at the middle block decomposition. Here, the two middle black blocks together have rank $r-s$. Using type p and type m Jordan moves (which preserve the other zero blocks) we can conjugate to obtain the rightmost block decomposition,
where the $\mathbb{1}$-block is the identity matrix $\mathbb{1}_{r-s} \in M_{r-s}$. That is, we obtain a matrix of block form

$$
\left[\begin{array}{cccc}
0_{n+s-2 r} & & & \\
& 0_{r-s} & \mathbb{1}_{r-s} & \\
& 0_{r-s} & 0_{r-s} & \operatorname{top}_{2} \\
& & & \operatorname{bot}_{2}
\end{array}\right]
$$

and the top left zero block of size $n+s-2 r$ splits off. If the bottom block bot $_{2}$ has rank $s$ we can conjugate to make top ${ }_{2}$ a zero block in which case also the Kronecker product block

$$
J_{2} \otimes \mathbb{1}_{r-s}=\left[\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right] \otimes \mathbb{1}_{r-s}=\left[\begin{array}{ll}
0_{r-s} & \mathbb{1}_{r-s} \\
0_{r-s} & 0_{r-s}
\end{array}\right]
$$

splits off and we stop the reduction. Otherwise, the square $s \times s$ block $^{\text {bot }}{ }_{2}$ has rank $t<s$ and we repeat the above reduction

$$
\left[\begin{array}{ccc}
0_{n-r} & \text { top }_{1} & \text { r-s } \\
0 & \text { bot }_{1} & \mathrm{~s}
\end{array}\right] \xrightarrow{\text { moves }}\left[\begin{array}{ccccc}
0_{n+s-2 r} & & & & \\
& 0_{r-s} & \mathbb{1}_{r-s} & & \\
& 0_{r-s} & 0_{r-s} & \operatorname{top}_{2} & \text { s.t } \\
& & 0 & \operatorname{bot}_{2} & \mathrm{t}
\end{array}\right]
$$

to the new top/bottom block (the tiny integers give the ranks of the top $p_{i}$ and bot $_{i}$ blocks). This gives us the following block matrix

$$
\left[\begin{array}{ccccccc}
0_{\mathrm{n}+\mathrm{s}-2 \mathrm{r}} & & & & & & \\
& 0_{\mathrm{r}+\mathrm{t}-2 \mathrm{~s}} & 0 & \mathbb{T}_{\mathrm{r}+\mathrm{t}-2 \mathrm{~s}} & 0 & & \\
& 0 & 0_{\mathrm{s}-\mathrm{t}} & 0 & \tau_{\mathrm{s}-\mathrm{t}} & & \\
& 0_{\mathrm{r}+\mathrm{t}-2 \mathrm{~s}} & 0 & 0_{\mathrm{r}+\mathrm{t}-2 \mathrm{~s}} & 0 & & \\
& 0 & 0_{\mathrm{s}-\mathrm{t}} & 0 & 0_{\mathrm{s}-\mathrm{t}} & \mathbb{1}_{\mathrm{s} \cdot \mathrm{t}} & \\
& & & & 0_{\mathrm{s}-\mathrm{t}} & 0_{\mathrm{s}-\mathrm{t}} & \operatorname{top}_{3} \\
& & & & & 0 & \operatorname{bot}_{3}
\end{array}\right]
$$

which after some permutation Jordan moves can be brought into the form

$$
\left[\begin{array}{ccccccccc}
0_{\mathrm{n}+\mathrm{s}-2 \mathrm{r}} & & & & & & & \\
& 0_{\mathrm{rtt-2s}} & \mathbb{1}_{\mathrm{r}+\mathrm{t}-2 \mathrm{~s}} & & & & & \\
& 0_{\mathrm{rrt-2s}} & 0_{\mathrm{r}+t-2 \mathrm{~s}} & & & & & \\
& & & 0_{\mathrm{s}-\mathrm{t}} & \mathbb{1}_{-\mathrm{s}} & 0_{\mathrm{s}-\mathrm{t}} & & \\
& & & 0_{\mathrm{s}-\mathrm{t}} & 0_{\mathrm{s}-\mathrm{t}} & \mathbb{1}_{\mathrm{s}-\mathrm{t}} & & \\
& & & 0_{\mathrm{ss-t}} & 0_{\mathrm{s}-\mathrm{t}} & 0_{\mathrm{s}-\mathrm{t}} & \text { top }_{3} & \mathrm{t}-\mathrm{u} \\
& & & & & 0 & \text { bot }_{3} & \mathrm{u}
\end{array}\right]
$$

Hence, the block $J_{1} \otimes \mathbb{1}_{r+t-2 s}$ splits off and if bot $_{3}$ is of rank $t$, so does the newly created Kronecker product

$$
J_{3} \otimes \mathbb{1}_{s-t}=\left[\begin{array}{ccc}
0 & 1 & 0 \\
0 & 0 & 1 \\
0 & 0 & 0
\end{array}\right] \otimes \mathbb{1}_{s-t}
$$

at which stage we stop the reduction. If the rank of $\operatorname{bot}_{3}$ is $u<t$, then after the next cycle we will be able to split off the Kronecker product $J_{3} \otimes \mathbb{1}_{s+u-2 t}$ and create a new 'almost'-split Kronecker product $J_{4} \otimes 1_{t-u}$.

It is now clear that after a finite number of cycles of this reduction process we will conjugate our $n \times n$ matrix $B$ to a matrix in diagonal block form

$$
\left(J_{1} \otimes \mathbb{1}_{\mathrm{n}+\mathrm{s}-2 \mathrm{r}}\right) \oplus\left(J_{2} \otimes \mathbb{1}_{r+t-2 \mathrm{~s}}\right) \oplus\left(J_{3} \otimes \mathbb{1}_{\mathrm{s}+\mathrm{u}-2 \mathrm{t}}\right) \oplus \ldots \oplus\left(J_{m} \otimes \mathbb{1}_{\mathrm{y}-\mathrm{z}}\right) \oplus \operatorname{bot}_{m+1}
$$

with bot $_{m+1}$ invertible or of size zero. Here, $J_{k}$ is the Jordan block matrix of size $k$ with zeroes everywhere except ones on the next to main diagonal.

Finally, again by permutation moves we can conjugate $B$ to the block diagonal matrix


Here, the bottom right corner bot is invertible, hence has all its eigenvalues $\left\{\mu_{1}, \mu_{2}, \ldots\right\}$ nonzero, and all the diagonal blocks in the upper left $d \times d$ corner are Jordan blocks $J_{k}$ (there are $n+s-2 r$ blocks $J_{1}, r+t-2 s$ blocks $J_{2}$ etc.). The integer $d$ is determined as the maximal power such that $t^{d}$ divides the characteristic polynomial $\chi_{B}(t)$. Hence, the sizes of these Jordan blocks

$$
p=(\underbrace{m, \ldots, m}_{y-z}, \ldots, \underbrace{3, \ldots, 3}_{s+4-2 \mathrm{t}}, \underbrace{2, \ldots, 2}_{\mathrm{r}+\mathrm{t}-2 \mathrm{~s}}, \underbrace{1, \ldots, 1}_{\mathrm{n}+\mathrm{s}-2 \mathrm{r}})
$$

form a partition of $d$.
Recall that a partition $p=\left(a_{1}, a_{2}, \ldots, a_{k}\right)$ of $d$ is a decompositions in natural numbers

$$
d=a_{1}+a_{2}+\ldots+a_{k} \quad \text { with } \quad a_{1} \geq a_{2} \geq \ldots \geq a_{k} \geq 1
$$

It is traditional to assign to a partition $p=\left(a_{1}, \ldots, a_{k}\right)$ a Young diagram with $a_{i}$ boxes in the $i$-th row, the rows of boxes lined up to the left.

The dual partition $p^{*}=\left(a_{1}^{*}, \ldots, a_{r}^{*}\right)$ to $p$ is defined by interchanging rows and columns in the Young diagram of $p$. For example, to the partition $p=$
$(3,2,1,1)$ of 7 we assign the Young diagram

with dual partition $p^{*}=(4,2,1)$.

## 1b. The Jordan-Weierstrass theorem.

Let us return to an arbitrary $n \times n$ matrix $A$ with characteristic polynomial $\chi_{A}(t)=\prod_{i=1}^{e}\left(t-\lambda_{i}\right)^{d_{i}}$. Apply the above reduction to the matrix $B=A_{i}=$ $\lambda_{i} \mathbb{1}-A$. Then, $A$ itself is conjugated to a block diagonal matrix of the form

with all the blocks in the $d_{i} \times d_{i}$ upper left corner Jordan blocks with eigenvalue $\lambda_{i}$, that is of the form

$$
J_{k}\left(\lambda_{i}\right)=\left[\begin{array}{cccc}
\lambda_{i} & 1 & & \\
& \lambda_{i} & \ddots & \\
& & \ddots & 1 \\
& & & \lambda_{i}
\end{array}\right] \in M_{k}(\mathbb{C})
$$

and the remaining block bot has eigenvalues $\left\{\lambda_{i}, \ldots, \check{\lambda}_{i}, \ldots, \lambda_{e}\right\}$. Repeating this procedure for the other eigenvalues we obtain the Jordan-Weierstrass theorem.

Theorem 1.1. Let $A \in M_{n}(\mathbb{C})$ with characteristic polynomial $\chi_{A}(t)=$ $\prod_{i=1}^{e}\left(t-\lambda_{i}\right)^{d_{i}}$. Then, $A$ determines unique partitions

$$
p_{i}=\left(a_{i 1}, a_{i 2}, \ldots, a_{i m_{i}}\right) \text { of } d_{i}
$$

for $1 \leq i \leq e$ such that $A$ is conjugated to a unique (up to permutation of the blocks) block-diagonal matrix
with $m=m_{1}+\ldots+m_{e}$ and exactly one block $B_{l}$ of the form $J_{a_{i j}}\left(\lambda_{i}\right)$ for all $1 \leq i \leq e$ and $1 \leq j \leq m_{i}$.

Remains only to prove the unicity of the partitions $p_{i}$ of $d_{i}$ corresponding to the eigenvalue $\lambda_{i}$ of $A$. Assume $A$ is conjugated to another Jordan block matrix $J_{\left(q_{1}, \ldots, q_{e}\right)}$, necessarily with partitions $q_{i}=\left(b_{i 1}, \ldots, b_{i m_{i}^{\prime}}\right)$ of $d_{i}$. To begin, observe that for a Jordan block of size $k$ we have that

$$
r k J_{k}(0)^{l}=k-l \quad \text { for all } l \leq k \text { and if } \mu \neq 0 \text { then } \quad r k J_{k}(\mu)^{l}=k
$$

for all $l$. As $J_{\left(p_{1}, \ldots, p_{e}\right)}$ is conjugated to $J_{\left(q_{1}, \ldots, q_{e}\right)}$ we have for all $\lambda \in \mathbb{C}$ and all $l$

$$
r k\left(\lambda \mathbb{1}_{n}-J_{\left(p_{1}, \ldots, p_{e}\right)}\right)^{l}=r k\left(\lambda \mathbb{1}_{n}-J_{\left(q_{1}, \ldots, q_{e}\right)}\right)^{l}
$$

Now, take $\lambda=\lambda_{i}$ then only on the Jordan blocks with eigenvalue $\lambda_{i}$ are important in the calculation and one obtains for the ranks

$$
\begin{equation*}
n-\sum_{h=1}^{l} \#\left\{j \mid a_{i j} \geq h\right\} \quad \text { respectively } \quad n-\sum_{h=1}^{l} \#\left\{j \mid b_{i j} \geq h\right\} \tag{1.1}
\end{equation*}
$$

Now, for any partition $p=\left(c_{1}, \ldots, c_{u}\right)$ and any natural number $h$ we see that the number $z=\#\left\{j \mid c_{j} \geq h\right\}$

is the number of blocks in the $h$-th row of the dual partition $p^{*}$. Therefore, the above rank equality implies that $p_{i}^{*}=q_{i}^{*}$ and hence that $p_{i}=q_{i}$. As we
can repeat this argument for the other eigenvalues we have the required unicity.

This completes the classification of the conjugacy classes of $n \times n$ matrices, or equivalently, the $G L_{n}$-orbits in $M_{n}$ which (for later reference) corresponds to the pattern :


We see that the classification consists of two parts : a discrete part choosing

- a partition $p=\left(d_{1}, d_{2}, \ldots, d_{e}\right)$ of $n$, and for each $d_{i}$,
- a partition $p_{i}=\left(a_{i 1}, a_{i 2}, \ldots, a_{i m_{i}}\right)$ of $d_{i}$,
determining the sizes of the Jordan blocks and a continuous part choosing
- an $e$-tuple of distinct complex numbers $\left(\lambda_{1}, \lambda_{2}, \ldots, \lambda_{e}\right)$.
fixing the eigenvalues. Moreover, this $e$-tuple $\left(\lambda_{1}, \ldots, \lambda_{e}\right)$ is determined only up to permutations of the subgroup $G$ of all permutations $S_{e}$ on $e$ letters where

$$
G=\left\{\pi \in S_{p} \mid p_{i}=p_{\pi(i)} \text { for all } 1 \leq i \leq e\right\}
$$

Whereas this gives a satisfactory set-theoretic description of the orbits, one might ask for a topological orbit space $C_{n}$ the points of which are in one-toone correspondence with the conjugacy classes, and a continuous surjection

$$
M_{n} \xrightarrow{c} C_{n}
$$

which is constant along $G L_{n}$-orbits and sends a matrix $A$ to the point of $C_{n}$ corresponding to the orbit $\mathcal{O}_{A}$. If we require that this space $C_{n}$ has at least the separation property that its points should be closed, then continuity of $c$ implies that for any matrix $A$ its conjugacy class $\mathcal{O}_{A}$ should be a closed subset of $M_{n}$.

However, this cannot be the case whenever $n \geq 2$. Consider the matrices

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

which by theorem 1.1. belong to distinct orbits. For any $\epsilon \neq 0$ we have that

$$
\left[\begin{array}{ll}
\epsilon & 0 \\
0 & 1
\end{array}\right] \cdot\left[\begin{array}{ll}
\lambda & 1 \\
0 & \lambda
\end{array}\right] \cdot\left[\begin{array}{cc}
\epsilon^{-1} & 0 \\
0 & 1
\end{array}\right]=\left[\begin{array}{cc}
\lambda & \epsilon \\
0 & \lambda
\end{array}\right]
$$

belongs to the orbit of $A$. Hence if $\epsilon \mapsto 0$, we see that $B$ lies in the closure of $\mathcal{O}_{A}$ whence $\mathcal{O}_{A}$ cannot be a closed orbit in $M_{2}$. As any matrix in $\mathcal{O}_{A}$ has trace $2 \lambda$, the orbit is contained in the 3 -dimensional subspace

$$
\left[\begin{array}{cc}
\lambda+x & y \\
z & \lambda-x
\end{array}\right] \longleftrightarrow M_{2}
$$

In this space, the orbit-closure $\overline{\mathcal{O}_{A}}$ is the set of points satisfying $x^{2}+y z=0$ (the determinant has to be $\lambda^{2}$ ), which is a cone having the origin as its top :


The orbit $\mathcal{O}_{B}$ is the top of the cone and the orbit $\mathcal{O}_{A}$ is the complement.

## Week 2

## The quotient space.

Last week we have seen that there is no Hausdorff topological space whose points are in one-to-one correspondence with the conjugacy classes of matrices. Still, we can try to solve :

## Problem 2.

Construct the best continuous approximation to the orbit space.
First, we construct a supply of complex valued continuous functions on $M_{n}$ that are constant along orbits.

## 2a. Invariant polynomial functions.

If two matrices are conjugated $A \sim B$, then $A$ and $B$ have the same unordered $n$-tuple of eigenvalues $\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}$ (occurring with multiplicities). Hence any symmetric function in the $\lambda_{i}$ will have the same values in $A$ as in $B$. In particular this is the case for the elementary symmetric functions $\sigma_{l}$

$$
\sigma_{l}\left(\lambda_{1}, \ldots, \lambda_{l}\right)=\sum_{i_{1}<i_{2}<\ldots<i_{l}} \lambda_{i_{1}} \lambda_{i_{2}} \ldots \lambda_{i_{l}} .
$$

Observe that for every $A \in M_{n}$ with eigenvalues $\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}$ we have

$$
\prod_{j=1}^{n}\left(t-\lambda_{j}\right)=\chi_{A}(t)=\operatorname{det}\left(t \mathbb{1}_{n}-A\right)=t^{n}+\sum_{i=1}^{n}(-1)^{i} \sigma_{i}(A) t^{n-i}
$$

Developing the determinant $\operatorname{det}\left(t \mathbb{1}_{n}-A\right)$ we see that each of the coefficients $\sigma_{i}(A)$ is in fact a polynomial function in the entries of $A$. A fortiori, $\sigma_{i}(A)$ is a complex valued continuous function on $M_{n}$. The above equality also
implies that the functions $\sigma_{i}: M_{n} \longrightarrow \mathbb{C}$ are constant along orbits. We now construct the continuous map

$$
M_{n} \xrightarrow{\pi} \mathbb{C}^{n}
$$

sending a matrix $A \in M_{n}$ to the point $\left(\sigma_{1}(A), \ldots, \sigma_{n}(A)\right)$ in $\mathbb{C}^{n}$. Clearly, if $A \sim B$ then they map to the same point in $\mathbb{C}^{n}$. We claim that $\pi$ is surjective. Take any point $\left(a_{1}, \ldots, a_{n}\right) \in \mathbb{C}^{n}$ and consider the matrix $A \in M_{n}$

$$
A=\left[\begin{array}{ccccc}
0 & & & & a_{n}  \tag{2.1}\\
-1 & 0 & & & a_{n-1} \\
& \ddots & \ddots & & \vdots \\
& & -1 & 0 & a_{2} \\
& & & -1 & a_{1}
\end{array}\right]
$$

then we will show that $\pi(A)=\left(a_{1}, \ldots, a_{n}\right)$, that is,

$$
\operatorname{det}\left(t \mathbb{1}_{n}-A\right)=t^{n}-a_{1} t^{n-1}+a_{2} t^{n-2}-\ldots+(-1)^{n} a_{n}
$$

Indeed, developing the determinant of $t \mathbb{T}_{n}-A$ along the first column we obtain

Here, the second determinant is equal to $(-1)^{n-1} a_{n}$ and by induction on $n$ the first determinant is equal to $t .\left(t^{n-1}-a_{1} t^{n-2}+\ldots+(-1)^{n-1} a_{n-1}\right)$, proving the claim.

Next, we will determine which $n \times n$ matrices can be conjugated to a matrix in the canonical form $A$ as above. We call a matrix $B \in M_{n}$ cyclic if there is a (column) vector $v \in \mathbb{C}^{n}$ such that $\mathbb{C}^{n}$ is spanned by the vectors $\left\{v, B . v, B^{2} . v, \ldots, B^{n-1} . v\right\}$. Let $g \in G L_{n}$ be the basechange transforming the standard basis to the ordered basis

$$
\left(v,-B \cdot v, B^{2} \cdot v,-B^{3} \cdot v, \ldots,(-1)^{n-1} B^{n-1} \cdot v\right) .
$$

In this new basis, the linear map determined by $B$ (or equivalently, $g \cdot B \cdot g^{-1}$ ) is equal to the matrix in canonical form

$$
\left[\begin{array}{ccccc}
0 & & & & b_{n} \\
-1 & 0 & & & b_{n-1} \\
& \ddots & \ddots & & \vdots \\
& & -1 & 0 & b_{2} \\
& & & -1 & b_{1}
\end{array}\right]
$$

where $B^{n} . v$ has coordinates $\left(b_{n}, \ldots, b_{2}, b_{1}\right)$ in the new basis.
Prove that a matrix $B \in M_{n}$ can be conjugated to one in standard form as above if and only if $B$ is a cyclic matrix.

We claim that the set of all cyclic matrices in $M_{n}$ is a dense open subset, that is, its closure is the whole of $M_{n}$. To see this take $v=\left(x_{1}, \ldots, x_{n}\right)^{\tau} \in \mathbb{C}^{n}$ and compute the determinant of the $n \times n$ matrix


This gives a polynomial of total degree $n$ in the $x_{i}$ with all its coefficients polynomial functions $c_{j}$ in the entries $b_{k l}$ of $B$. Now, $B$ is a cyclic matrix if and only if at least one of these coefficients is non-zero. That is, the set of non-cyclic matrices is exactly the intersection of the finitely many hypersurfaces

$$
V_{j}=\left\{B=\left(b_{k l}\right)_{k, l} \in M_{n} \mid c_{j}\left(b_{11}, b_{12}, \ldots, b_{n n}\right)=0\right\}
$$

in the vectorspace $M_{n}$. The claim follows because the complement of the hypersurface $V_{j}$ is a dense open subset (being equal to the inverse image of the dense open set $\mathbb{C}-\{0\}$ in $\mathbb{C}$ under the continuous surjection $c_{j}$ : $M_{n} \longrightarrow \mathbb{C}$ ).

Theorem 2.1. The best continuous approximation to the orbit space is given by the surjection

$$
M_{n} \xrightarrow{\pi} \mathbb{C}^{n}
$$

mapping a matrix $A \in M_{n}(\mathbb{C})$ to the $n$-tuple $\left(\sigma_{1}(A), \ldots, \sigma_{n}(A)\right)$.

Let $f: M_{n} \longrightarrow \mathbb{C}$ be a continuous function which is constant along conjugacy classes. We will show that $f$ factors through $\pi$, that is, $f$ is really a continuous function in the $\sigma_{i}(A)$. Consider the diagram

where $s$ is the section of $\pi$ (that is, $\pi \circ s=i d_{\mathbb{C}^{n}}$ ) determined by sending a point $\left(a_{1}, \ldots, a_{n}\right)$ to the cyclic matrix in canonical form $A$ as in equation (2.1). Clearly, $s$ is continuous, hence so is $f^{\prime}=f \circ s$. The approximation property follows if we prove that $f=f^{\prime} \circ \pi$. By continuity, it suffices to check equality on the dense open set of cyclic matrices in $M_{n}$.

There it is a consequence of the following three facts we have proved before : (1) : any cyclic matrix lies in the same orbit as one in standard form, (2) $: s$ is a section of $\pi$ and (3) : $f$ is constant along orbits.

## 2b. Some examples.

## Miniature 1. Orbits in $M_{2}$.

A $2 \times 2$ matrix $A$ can be conjugated to an upper triangular matrix with diagonal entries the eigenvalues $\lambda_{1}, \lambda_{2}$ of $A$. As the trace and determinant of both matrices are equal we have

$$
\sigma_{1}(A)=\operatorname{tr}(A) \text { and } \sigma_{2}(A)=\operatorname{det}(A) .
$$

The best approximation to the orbitspace is therefore given by the surjective map

$$
M_{2} \xrightarrow{\pi} \mathbb{C}^{2} \quad\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right] \mapsto(a+d, a d-b c)
$$

The matrix $A$ has two equal eigenvalues if and only if the discriminant of the characteristic polynomial $t^{2}-\sigma_{1}(A) t+\sigma_{2}(A)$ is zero, that is when $\sigma_{1}(A)^{2}-4 \sigma_{2}(A)=0$. This condition determines a closed curve $C$ in $\mathbb{C}^{2}$ where

$$
C=\left\{(x, y) \in \mathbb{C}^{2} \mid x^{2}-4 y=0\right\} .
$$



Observe that $C$ is a smooth 1 -dimensional submanifold of $\mathbb{C}^{2}$. We will describe the fibers (that is, the inverse images of points) of the surjective map $\pi$.

If $p=(x, y) \in \mathbb{C}^{2}-C$, then $\pi^{-1}(p)$ consists of precisely one orbit (which is then necessarily closed in $M_{2}$ ) namely that of the diagonal matrix

$$
\left[\begin{array}{cc}
\lambda_{1} & 0 \\
0 & \lambda_{2}
\end{array}\right] \quad \text { where } \quad \lambda_{1,2}=\frac{-x \pm \sqrt{x^{2}-4 y}}{2}
$$

If $p=(x, y) \in C$ then $\pi^{-1}(p)$ consists of two orbits,

$$
\mathcal{O}^{\lambda}\left[\begin{array}{ll}
\lambda & 1 \\
0 & \lambda
\end{array}\right] \quad \text { and } \quad \mathcal{O} \quad\left[\begin{array}{cc}
\lambda & 0 \\
0 & \lambda
\end{array}\right]
$$

where $\lambda=\frac{1}{2} x$. We have seen that the second orbit lies in the closure of the first. Observe that the second orbit reduces to one point in $M_{2}$ and hence is closed. Hence, also $\pi^{-1}(p)$ contains a unique closed orbit.

To describe the fibers of $\pi$ as closed subsets of $M_{2}$ it is convenient to write any matrix $A$ as a linear combination

$$
A=u(A)\left[\begin{array}{cc}
\frac{1}{2} & 0 \\
0 & \frac{1}{2}
\end{array}\right]+v(A)\left[\begin{array}{cc}
\frac{1}{2} & 0 \\
0 & -\frac{1}{2}
\end{array}\right]+w(A)\left[\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right]+z(A)\left[\begin{array}{ll}
0 & 0 \\
1 & 0
\end{array}\right] .
$$

Expressed in the coordinate functions $u, v, w$ and $z$ the fibers $\pi^{-1}(p)$ of a point $p=(x, y) \in$ $\mathbb{C}^{2}$ are the common zeroes of

$$
\begin{cases}u & =x \\ v^{2}+4 w z & =x^{2}-4 y\end{cases}
$$

The first equation determines a three dimensional affine subspace of $M_{2}$ in which the second equation determines a quadric.


If $p \notin C$ this quadric is non-degenerate and thus $\pi^{-1}(p)$ is a smooth 2-dimensional submanifold of $M_{2}$. If $p \in C$, the quadric is a cone with top lying in the point $\frac{x}{2} \mathbb{1}_{2}$. Under the $G L_{2}$-action, the unique singular point of the cone must be clearly fixed giving us the closed orbit of dimension 0 corresponding to the diagonal matrix. The other orbit is the complement of the top and hence is a smooth 2-dimensional (non-closed) submanifold of $M_{2}$. The graphs represent the orbit-closures and the dimensions of the orbits.

## Miniature 2. Orbits in $M_{3}$.

We will describe the fibers of the surjective map $M_{3} \xrightarrow{\pi} \mathbb{C}^{3}$. If a $3 \times 3$ matrix has multiple eigenvalues then the discriminant $d=\left(\lambda_{1}-\lambda_{2}\right)^{2}\left(\lambda_{2}-\lambda_{3}\right)^{2}\left(\lambda_{3}-\lambda_{1}\right)^{2}$ is zero. Clearly, $d$ is a symmetric polynomial and hence can be expressed in terms of $\sigma_{1}, \sigma_{2}$ and $\sigma_{3}$. More precisely,

$$
d=4 \sigma_{1}^{3} \sigma_{3}+4 \sigma_{2}^{3}+27 \sigma_{3}^{2}-\sigma_{1}^{2} \sigma_{2}^{2}-18 \sigma_{1} \sigma_{2} \sigma_{3}
$$

The set of points in $\mathbb{C}^{3}$ where $d$ vanishes is a surface $S$ with singularities.


These singularities are the common zeroes of the $\frac{\partial d}{\partial \sigma_{i}}$ for $1 \leq i \leq 3$. One computes that these singularities form a twisted cubic curve $C$ in $\mathbb{C}^{3}$, that is,

$$
C=\left\{\left(3 c, 3 c^{2}, c^{3}\right) \mid c \in \mathbb{C}\right\}
$$

The description of the fibers $\pi^{-1}(p)$ for $p=(x, y, z) \in \mathbb{C}^{3}$ is as follows. When $p \notin S$, then $\pi^{-1}(p)$ consists of a unique orbit (which is therefore closed in $M_{3}$ ), the conjugacy class of a matrix with pairwise distinct eigenvalues. If $p \in S-C$, then $\pi^{-1}(p)$ consists of the orbits of

$$
A_{1}=\left[\begin{array}{ccc}
\lambda & 1 & 0 \\
0 & \lambda & 0 \\
0 & 0 & \mu
\end{array}\right] \text { and } A_{2}=\left[\begin{array}{ccc}
\lambda & 0 & 0 \\
0 & \lambda & 0 \\
0 & 0 & \mu
\end{array}\right]
$$

Finally, if $p \in C$, then the matrices in the fiber $\pi^{-1}(p)$ have a single eigenvalue $\lambda=\frac{1}{3} x$ and the fiber consists of the orbits of the matrices

$$
B_{1}=\left[\begin{array}{ccc}
\lambda & 1 & 0 \\
0 & \lambda & 1 \\
0 & 0 & \lambda
\end{array}\right] \quad B_{2}=\left[\begin{array}{ccc}
\lambda & 1 & 0 \\
0 & \lambda & 0 \\
0 & 0 & \lambda
\end{array}\right] \quad B_{3}=\left[\begin{array}{ccc}
\lambda & 0 & 0 \\
0 & \lambda & 0 \\
0 & 0 & \lambda
\end{array}\right]
$$

We observe that the strata with distinct fiber behavior (that is, $\mathbb{C}^{3}-S, S-C$ and $C$ ) are all submanifolds of $\mathbb{C}^{3}$.

The dimension of an orbit $\mathcal{O}_{A}$ in $M_{n}$ is computed as follows. Let $C_{A}$ be the subspace of all matrices in $M_{n}$ commuting with $A$. Then, the stabilizer subgroup of $A$ is a dense open subset of $C_{A}$ whence the dimension of $\mathcal{O}_{A}$ is equal to $n^{2}-\operatorname{dim} C_{A}$.

Performing these calculations for the matrices given above, we obtain the following graphs representing orbit-closures and the dimensions of orbits


## Week 3

## The orbit closures.

This week we will prove the Gerstenhaber-Hesselink theorem which gives an answer to :

## Problem 3.

Describe the orbit-closures for general $n$.

Consider the quotient map $M_{n} \xrightarrow{\pi} \mathbb{C}^{n}$ and a point $x=\left(x_{1}, \ldots, x_{n}\right) \in \mathbb{C}^{n}$. All matrices $A$ in the fiber $\pi^{-1}(x)$ have the same eigenvalues $\left\{\lambda_{1}, \ldots, \lambda_{e}\right\}$ and multiplicities $\left\{d_{1}, \ldots, d_{e}\right\}$ because

$$
\prod_{i=1}^{e}\left(t-\lambda_{i}\right)^{d_{i}}=\chi_{A}(t)=t^{n}-x_{1} t^{n+1}+x_{2} t^{n-2}-\ldots+(-1)^{n} x_{n}
$$

We have seen in example ?? that the orbit corresponding to the diagonal matrix with these eigenvalues and multiplicities is contained in the closure of $\mathcal{O}_{A}$ for any $A \in \pi^{-1}(x)$. Moreover, for any $\lambda \in \mathbb{C}$, the linear automorphism

$$
M_{n} \xrightarrow{\phi_{\lambda}} M_{n} \quad \text { defined by } \quad A \mapsto A-\lambda \mathbb{1}_{n}
$$

commutes with the action of $G L_{n}$. Hence, in studying the fibers $\pi^{-1}(x)$ we may assume that one of the eigenvalues $\lambda_{i}$ is zero. An important subproblem is therefore to study the orbit-closures in the nullcone, $N=\pi^{-1}(\underline{0})$, that is the set of all nilpotent $n \times n$ matrices. We recall that a matrix $A$ is said to be nilpotent if $A^{k}=0$ for some power $k$. Clearly, $A \in M_{n}$ is nilpotent if and only if $A^{n}=0$.

A nilpotent matrix $A$ has 0 as its unique eigenvalue (occurring with multiplicity $n$ ). Therefore, by theorem ?? the orbits of nilpotent matrices are in one-to-one correspondence with partitions of $n$. We will first introduce a dominance ordering on all partitions of $n$ and consequently show that this ordering determines the orbit closures of nilpotent matrices.

## 3a. The Gerstenhaber-Hesselink theorem.

It is sometimes convenient to relax our definition of partitions to include zeroes at its tail. That is, a partition $p$ of $n$ is an integral $n$-tuple $\left(a_{1}, a_{2}, \ldots, a_{n}\right)$ with $a_{1} \geq a_{2} \geq \ldots \geq a_{n} \geq 0$ with $\sum_{i=1}^{n} a_{i}=n$. As before, we represent a partition by a Young diagram by omitting rows corresponding to zeroes.

If $q=\left(b_{1}, \ldots, b_{n}\right)$ is another partition of $n$ we say that $p$ dominates $q$ and write

$$
p>q \quad \text { if and only if } \quad \sum_{i=1}^{r} a_{i} \geq \sum_{i=1}^{r} b_{i} \quad \text { for all } 1 \leq r \leq n
$$

For example, the partitions of 4 are ordered as indicated below


Note however that the dominance relation is not a total ordering whenever $n \geq 6$. For example, the following two partition of 6

are not comparable. The dominance order is induced by the Young move of throwing a row-ending box down the diagram. Indeed, let $p$ and $q$ be partitions of $n$ such that $p>q$ and assume there is no partition $r$ such that $p>r$ and $r>q$. Let $i$ be the minimal number such that $a_{i}>b_{i}$. Then by the assumption $a_{i}=b_{i}+1$. Let $j>i$ be minimal such that $a_{j} \neq b_{j}$, then we have $b_{j}=a_{j}+1$ because $p$ dominates $q$. But then, the remaining rows of $p$ and $q$ must be equal. That is, a Young move can be depicted as


For example, the Young moves between the partitions of 4 given above are as indicated


A Young $p$-tableau is the Young diagram of $p$ with the boxes labeled by integers from $\{1,2, \ldots, s\}$ for some $s$ such that each label appears at least ones. A Young $p$-tableau is said to be of type $q$ for some partition $q=\left(b_{1}, \ldots, b_{n}\right)$ of $n$ if the following conditions are met:

- the labels are non-decreasing along rows,
- the labels are strictly increasing along columns, and
- the label $i$ appears exactly $b_{i}$ times.

For example, if $p=(3,2,1,1)$ and $q=(2,2,2,1)$ then the $p$-tableau below

| 1 | 1 | 3 |
| :--- | :--- | :--- |
| 2 | 2 |  |
| 3 |  |  |
| 4 |  |  |
|  |  |  |

is of type $q$ (observe that $p>q$ and even $p \rightarrow q$ ). In general, let $p=\left(a_{1}, \ldots, a_{n}\right)$ and $q=\left(b_{1}, \ldots, b_{n}\right)$ be partitions of $n$ and assume that $p \rightarrow q$. Then, there is a Young $p$-tableau of type $q$. For, fill the Young diagram of $q$ by putting 1's in the first row, 2's in the second and so on. Then, upgrade the fallen box together with its label to get a Young $p$-tableau of type $q$. In the example above


Conversely, assume there is a Young $p$-tableau of type $q$. The number of boxes labeled with a number $\leq i$ is equal to $b_{1}+\ldots+b_{i}$. Further, any box with label $\leq i$ must lie in the first $i$ rows (because the labels strictly increase along a column). There are $a_{1}+\ldots+a_{i}$ boxes available in the first $i$ rows whence

$$
\sum_{j=1}^{i} b_{i} \leq \sum_{j=1}^{i} a_{i} \quad \text { for all } \quad 1 \leq i \leq n
$$

and therefore $p>q$. After these preliminaries on partitions, let us return to nilpotent matrices.

Let $A$ be a nilpotent matrix of type $p=\left(a_{1}, \ldots, a_{n}\right)$, that is, conjugated to a matrix with Jordan blocks (all with eigenvalue zero) of sizes $a_{i}$. It follows
from equation ?? that the subspace $V_{l}$ of column vectors $v \in \mathbb{C}^{n}$ such that $A^{l} \cdot v=0$ has dimension

$$
\sum_{h=1}^{l} \#\left\{j \mid a_{j} \geq h\right\}=\sum_{h=1}^{l} a_{h}^{*}
$$

where $p^{*}=\left(a_{1}^{*}, \ldots, a_{n}^{*}\right)$ is the dual partition of $p$. Choose a basis $\left\{v_{1}, \ldots, v_{n}\right\}$ of $\mathbb{C}^{n}$ such that for all $l$ the first $a_{1}^{*}+\ldots+a_{l}^{*}$ base vectors span the subspace $V_{l}$. For example, if $A$ is in Jordan normal form of type $p=(3,2,1,1)$

$$
\left[\begin{array}{lllllll}
0 & 1 & 0 & & & & \\
0 & 0 & 1 & & & & \\
0 & 0 & 0 & & & & \\
& & & 0 & 1 & & \\
& & & 0 & 0 & & \\
& & & & & 0 & \\
& & & & & & 0
\end{array}\right]
$$

then $p^{*}=(4,2,1)$ and we can choose the standard base vectors ordered as follows


Take a partition $q=\left(b_{1}, \ldots, b_{n}\right)$ with $p \rightarrow q$ (in particular, $p>q$ ), then for the dual partitions we have $q^{*} \rightarrow p^{*}$ (and thus $q^{*}>p^{*}$ ). But then there is a Young $q^{*}$-tableau of type $p^{*}$. In the example with $q=(2,2,2,1)$ we have $q^{*}=(4,3)$ and $p^{*}=(4,2,1)$ and we can take the Young $q^{*}$-tableau of type $p^{*}$

| 1 | 1 | 1 | 1 |
| :--- | :--- | :--- | :--- |
| 2 | 2 | 3 |  |
|  |  |  |  |

Now label the boxes of this tableau by the base vectors $\left\{v_{1}, \ldots, v_{n}\right\}$ such that the boxes labeled $i$ in the Young $q^{*}$-tableau of type $p^{*}$ are filled with the base vectors from $V_{i}-V_{i-1}$. Call this tableau $T$. In the example, we can take

$$
T=
$$

Define a linear operator $F$ on $\mathbb{C}^{n}$ by the rule that $F\left(v_{i}\right)=v_{j}$ if $v_{j}$ is the label of the box in $T$ just above the box labeled $v_{i}$. In case $v_{i}$ is a label of a box in the first row of $T$ we take $F\left(v_{i}\right)=0$. Obviously, $F$ is a nilpotent $n \times n$ matrix and by construction we have that

$$
r k F^{l}=n-\left(b_{1}^{*}+\ldots+b_{l}^{*}\right)
$$

That is, $F$ is nilpotent of type $q=\left(b_{1}, \ldots, b_{n}\right)$. Moreover, $F$ satisfies $F\left(V_{i}\right) \subset$ $V_{i-1}$ for all $i$ by the way we have labeled the tableau $T$ and defined $F$.

In the example above, we have $F\left(e_{2}\right)=e_{1}, F\left(e_{5}\right)=e_{4}, F\left(e_{3}\right)=e_{6}$ and all other $F\left(e_{i}\right)=0$. That is, $F$ is the matrix

which is seen to be of type $(2,2,2,1)$ after performing a few Jordan moves.
Returning to the general case, consider for all $\epsilon \in \mathbb{C}$ the $n \times n$ matrix :

$$
F_{\epsilon}=(1-\epsilon) F+\epsilon A .
$$

We claim that for all but finitely many values of $\epsilon$ we have $F_{\epsilon} \in \mathcal{O}_{A}$. Indeed, we have seen that $F\left(V_{i}\right) \subset V_{i-1}$ where $V_{i}$ is defined as the subspace such that $A^{i}\left(V_{i}\right)=0$. Hence, $F\left(V_{1}\right)=0$ and therefore

$$
F_{\epsilon}\left(V_{1}\right)=(1-\epsilon) F+\epsilon A\left(V_{1}\right)=0 .
$$

Assume by induction that $F_{\epsilon}^{i}\left(V_{i}\right)=0$ holds for all $i<l$, then we have that

$$
\begin{aligned}
F_{\epsilon}^{l}\left(V_{l}\right) & =F_{\epsilon}^{l-1}((1-\epsilon) F+\epsilon A)\left(V_{l}\right) \\
& \subset F_{\epsilon}^{l-1}\left(V_{l-1}\right)=0
\end{aligned}
$$

because $A\left(V_{l}\right) \subset V_{l-1}$ and $F\left(V_{l}\right) \subset V_{l-1}$. But then we have for all $l$ that

$$
r k F_{\epsilon}^{l} \leq \operatorname{dim} V_{l}=n-\left(a_{1}^{*}+\ldots+a_{l}^{*}\right)=r k A^{l} \stackrel{\text { def }}{=} r_{l} .
$$

Then for at least one $r_{l} \times r_{l}$ submatrix of $F_{\epsilon}^{l}$ its determinant considered it as a polynomial of degree $r_{l}$ in $\epsilon$ is not identically zero (as it is nonzero for $\epsilon=1$ ). But then this determinant is non-zero for all but finitely many $\epsilon$. Hence, $r k F_{\epsilon}^{l}=r k A^{l}$ for all $l$ for all but finitely many $\epsilon$. As these numbers determine the dual partition $p^{*}$ of the type of $A, F_{\epsilon}$ is a nilpotent $n \times n$ matrix of type $p$ for all but finitely many values of $\epsilon$, proving the claim. But then, $F_{0}=F$ which we have proved to be a nilpotent matrix of type $q$ belongs to the closure of the orbit $\mathcal{O}_{A}$. That is, we have proved the difficult part of the Gerstenhaber-Hesselink theorem.

Theorem 3.1. Let A be a nilpotent $n \times n$ matrix of type $p=\left(a_{1}, \ldots, a_{n}\right)$ and $B$ nilpotent of type $q=\left(b_{1}, \ldots, b_{n}\right)$. Then, $B$ belongs to the closure of the orbit $\mathcal{O}_{A}$, that is,

$$
B \in \overline{\mathcal{O}_{A}} \quad \text { if and only if } \quad p>q
$$

in the domination order on partitions of $n$.
To prove the theorem we only have to observe that if $B$ is contained in the closure of $A$, then $B^{l}$ is contained in the closure of $A^{l}$ and hence $r k A^{l} \geq$ $r k B^{l}$ (because $r k A^{l}<k$ is equivalent to vanishing of all determinants of $k \times k$ minors which is a closed condition). But then,

$$
n-\sum_{i=1}^{l} a_{i}^{*} \geq n-\sum_{i=1}^{l} b_{i}^{*}
$$

for all $l$, that is, $q^{*}>p^{*}$ and hence $p>q$. The other implication was proved above if we remember that the domination order was induced by the Young moves and clearly we have that if $B \in \overline{\mathcal{O}_{C}}$ and $C \in \overline{\mathcal{O}_{A}}$ then also $B \in \overline{\mathcal{O}_{A}}$.

## 3b. Some examples.

## Miniature 3. Nilpotent matrices for small $n$.

We will apply theorem 1 to describe the orbit-closures of nilpotent matrices of $8 \times 8$ matrices. The following table lists all partitions (and their dual in the other column)

The partitions of 8 .

|  |  |  |  |
| :--- | :--- | :---: | :--- |
| a | $(8)$ | v | $(1,1,1,1,1,1,1,1)$ |
| b | $(7,1)$ | u | $(2,1,1,1,1,1,1)$ |
| c | $(6,2)$ | t | $(2,2,1,1,1,1)$ |
| d | $(6,1,1)$ | s | $(3,1,1,1,1,1)$ |
| e | $(5,3)$ | r | $(2,2,2,1,1)$ |
| f | $(5,2,1)$ | q | $(3,2,1,1,1)$ |
| g | $(5,1,1,1)$ | p | $(4,1,1,1,1)$ |
| h | $(4,4)$ | o | $(2,2,2,2)$ |
| i | $(4,3,1)$ | n | $(3,2,2,1)$ |
| j | $(4,2,2)$ | m | $(3,3,1,1)$ |
| k | $(3,3,2)$ | k | $(3,3,2)$ |
| l | $(4,2,1,1)$ | l | $(4,2,1,1)$ |
|  |  |  |  |

The domination order between these partitions can be depicted as follows where all
the Young moves are from left to right


Of course, from this graph we can read off the dominance order graphs for partitions of $n \leq 8$. The trick is to identify a partition of $n$ with that of 8 by throwing in a tail of ones and to look at the relative position of both partitions in the above picture. Using these conventions we get the following graph for partitions of 7

and for partitions of 6 the dominance order is depicted as follows


We have already mentioned that the dominance order on partitions of $n \leq 5$ is a total ordering.

We will prove later that knowledge of the orbit closures of nilpotent matrices for $m \times m$ matrices for all $m \leq n$ is enough to understand the orbit closures in all the fibers $\pi^{-1}(x)$ for the quotient map $M_{n} \xrightarrow{\pi} \mathbb{C}^{n}$.

## Week 4

## Dynamical systems.

In this section we will consider linear time-invariant dynamical systems. Whereas this is a gross simplification of actual processes, often one can reduce to such a situation (as a first approximation), for example near an equilibrium state of the system. A linear time invariant dynamical system $\Sigma$ is governed by the following system of differential equations

$$
\begin{cases}\frac{d x}{d t} & =B x+A u  \tag{4.1}\\ y & =C x\end{cases}
$$

Here, $u(t) \in \mathbb{C}^{m}$ is the input or control of the system at tome $t, x(t) \in \mathbb{C}^{n}$ the state of the system and $y(t) \in \mathbb{C}^{p}$ the output of the system $\Sigma$. Time invariance of $\Sigma$ means that the matrices $A \in M_{n \times m}(\mathbb{C}), B \in M_{n}(\mathbb{C})$ and $C \in M_{p \times n}(\mathbb{C})$ are constant. The system $\Sigma$ can be represented as a black box

which is in a certain state $x(t)$ that we can try to change by using the input controls $u(t)$. By reading the output signals $y(t)$ we can try to determine the state of the system. We briefly recall how one solves a linear dynamical system.

## 4a. Solving linear systems.

Recall that the matrix exponential $e^{B}$ of any $n \times n$ matrix $B$ is defined by the infinite series

$$
e^{B}=\mathbb{1}_{n}+B+\frac{B^{2}}{2!}+\ldots+\frac{B^{m}}{m!}+\ldots
$$

Observe that this series converges to a matrix in $M_{n}(\mathbb{C})$ as the norm $\left|B^{n}\right|$ is bounded by $|B|^{m}$ for any $m$. The importance of this construction is clear from the fact that $e^{B t}$ is the fundamental matrix for the homogeneous differential equation $\frac{d x}{d t}=B x$. That is, the columns of $e^{B t}$ are a basis for the $n$-dimensional space of solutions of the equation $\frac{d x}{d t}=B x$.

Motivated by this, let us look for a solution to equation (4.1) as the form $x(t)=e^{B t} g(t)$ for some function $g(t)$. Substitution gives the condition

$$
\frac{d g}{d t}=e^{-B t} A u \quad \text { whence } \quad g(\tau)=g\left(\tau_{0}\right)+\int_{\tau_{0}}^{\tau} e^{-B t} A u(t) d t
$$

Observe that $x\left(\tau_{0}\right)=e^{B \tau_{0}} g\left(\tau_{0}\right)$ and we obtain the solution of the linear dynamical system $\Sigma=(A, B, C)$ :

$$
\left\{\begin{aligned}
x(\tau) & =e^{\left(\tau-\tau_{0}\right) B} x\left(\tau_{0}\right)+\int_{\tau_{0}}^{\tau} e^{(\tau-t) B} A u(t) d t \\
y(\tau) & =C e^{B\left(\tau-\tau_{0}\right)} x\left(\tau_{0}\right)+\int_{\tau_{0}}^{\tau} C e^{(\tau-t) B} A u(t) d t .
\end{aligned}\right.
$$

Differentiating we see that this is indeed a solution and it is the unique one having a prescribed starting state $x\left(\tau_{0}\right)$. Indeed, given another solution $x_{1}(\tau)$ we have that $x_{1}(\tau)-x(\tau)$ is a solution to the homogeneous system $\frac{d x}{d t}=B t$, but then

$$
x_{1}(\tau)=x(\tau)+e^{\tau B} e^{-\tau_{0} B}\left(x_{1}\left(\tau_{0}\right)-x\left(\tau_{0}\right)\right)
$$

## 4b. Observable and controllable systems.

We will recall some important system-theoretic notions describing the level of control or observation a given system allows.

We call the system $\Sigma$ completely controllable if we can steer any starting state $x\left(\tau_{0}\right)$ to the zero state by some control function $u(t)$ in a finite time span $\left[\tau_{0}, \tau\right]$. That is, the equation

$$
0=x\left(\tau_{0}\right)+\int_{\tau_{0}}^{\tau} e^{\left(\tau_{0}-t\right) B} A u(t) d t
$$

has a solution in $\tau$ and $u(t)$. As the system is time-invariant we may always eqcontrol assume that $\tau_{0}=0$ and have to satisfy the equation

$$
\begin{equation*}
0=x_{0}+\int_{0}^{\tau} e^{t B} A u(t) d t \quad \text { for every } \quad x_{0} \in \mathbb{C}^{n} \tag{4.2}
\end{equation*}
$$

Consider the control matrix $c(\Sigma)$ which is the $n \times m n$ matrix

$c(\Sigma)=$| $A$ | $B A$ | $B^{2} A$ | $\cdots$ | $B^{\mathrm{n}-1} A$ |
| :--- | :--- | :--- | :--- | :--- |

Assume that $r k c(\Sigma)<n$ then there is a non-zero state $s \in \mathbb{C}^{n}$ such that $s^{t r} c(\Sigma)=0$, where $s^{t r}$ denotes the transpose (row column) of $s$. Because $B$ satisfies the characteristic polynomial $\chi_{B}(t), B^{n}$ and all higher powers $B^{m}$ are linear combinations of $\left\{\tau_{n}, B, B^{2}, \ldots, B^{n-1}\right\}$. Hence, $s^{\tau} B^{m} A=0$ for all $m$. Writing out the power series expansion of $e^{t B}$ in equation (4.2) this leads to the contradiction that $0=s^{\tau} x_{0}$ for all $x_{0} \in \mathbb{C}^{n}$. Hence, if $r k c(\Sigma)<n$, then $\Sigma$ is not completely controllable.

Conversely, let $r k c(\Sigma)=n$ and assume that $\Sigma$ is not completely controllable. That is, the space of all states

$$
s(\tau, u)=\int_{0}^{\tau} e^{-t B} A u(t) d t
$$

is a proper subspace of $\mathbb{C}^{n}$. But then, there is a non-zero state $s \in \mathbb{C}^{n}$ such that $s^{t r} s(\tau, u)=0$ for all $\tau$ and all functions $u(t)$. Differentiating this with respect to $\tau$ we obtain

$$
\begin{equation*}
s^{t r} e^{-\tau B} A u(\tau)=0 \quad \text { whence } \quad s^{t r} e^{-\tau B} A=0 \tag{4.3}
\end{equation*}
$$

for any $\tau$ as $u(\tau)$ can take on any vector. For $\tau=0$ this gives $s^{t r} A=0$. If we differentiate (4.3) with respect to $\tau$ we get $s^{t r} B e^{-\tau B} A=0$ for all $\tau$ and for $\tau=0$ this gives $s^{t r} B A=0$. Iterating this process we show that $s^{t r} B^{m} A=0$ for any $m$, whence

$$
s^{\operatorname{tr}}\left[\begin{array}{lllll}
A & B A & B^{2} A & \ldots & B^{n-1} A
\end{array}\right]=0
$$

contradicting the assumption that $r k c(\Sigma)=n$. That is, we have proved :
Proposition 4.1. A linear time-invariant dynamical system $\Sigma$ determined by the matrices $(A, B, C)$ is completely controllable if and only if $r k c(\Sigma)$ is maximal.


Next, we turn to the problem to what degree we can obtain information about the system by reading off its output displays. We say that a state $x(\tau)$ at time $\tau$ is unobservable if $C e^{(\tau-t) B} x(\tau)=0$ for all $t$. Intuitively this means that the state $x(\tau)$ cannot be detected uniquely from the output of the system $\Sigma$. Again, if we differentiate this condition a number of times and evaluate at $t=\tau$ we obtain the conditions

$$
C x(\tau)=C B x(\tau)=\ldots=C B^{n-1} x(\tau)=0
$$

We say that $\Sigma$ is completely observable if the zero state is the only unobservable state at any time $\tau$. Consider the observation matrix $o(\Sigma)$ of the system $\Sigma$ which is the $p n \times n$ matrix

$$
o(\Sigma)=\left[\begin{array}{llll}
C^{t r} & (C B)^{t r} & \ldots & \left(C B^{n-1}\right)^{t r}
\end{array}\right]^{t r}
$$

An analogous argument as in the proof of proposition 4.1. gives us :
Proposition 4.2. A linear time-invariant dynamical system $\Sigma$ determined by the matrices $(A, B, C)$ is completely observable if and only if $r k o(\Sigma)$ is maximal.

## 4c. Some examples.

## Miniature 4. A physical dynamical system.

We consider a simple magnetic-ball suspension system. The objective of the system is to control the position of the steel ball by adjusting the current in the electromagnet through the input voltage $e(t)$.


The differential equations determining this system are :

$$
\begin{cases}M \frac{d^{2} y(t)}{d t^{2}} & =M g-\frac{i^{2}(t)}{y(t)} \\ e(t) & =R i(t)+L \frac{d i(t)}{d t}\end{cases}
$$

Here, $y(t)$ is the distance of teh ball from the magnet, $M$ the mass of the ball and $g$ the gravitational constant. The electromagnet has winding inductance $L$, winding resistance $R$ and $i(t)$ is the winding current. If we want to approximate this system by a linear system we have to solve two problems : (1) to replace the higher order differential term in the left hand side by first order terms, and (2) to linearize the non-linear terms in teh right hand side.

The first problem is solved by teh standard trick that a single $n$-th order differential equation

$$
\frac{d^{n} y}{d t^{n}}+a_{1} \frac{d^{n-1} y}{d t^{n-1}}+\ldots+a_{n} y=u(t)
$$

is equivalent to a linear system

$$
\frac{d x}{d t}=F x+u
$$

where we take $x=\left(x_{1}, \ldots, x_{n}\right)$ with $x_{1}=y, x_{i}=\frac{d^{i} y}{d t^{i}}$ for $i \geq 2$ and where $F$ is the $n \times n$ matrix

$$
F=\left[\begin{array}{ccccc}
0 & 0 & \ldots & 0 & -a_{n} \\
1 & 0 & \ldots & 0 & -a_{n-1} \\
0 & 1 & & 0 & -a_{n-2} \\
& & \ddots & & \vdots \\
0 & 0 & & 1 & -a_{1}
\end{array}\right]
$$

Hence, in the above suspension system we take as state vaiables $x_{1}(t)=y(t), x_{2}(t)=\frac{d y}{d t}$ and $x_{3}(t)=i(t)$. Then, the defining equations of teh system become

$$
\begin{cases}\frac{d x_{1}}{d t} & =x_{2} \\ \frac{d x_{2}}{d t} & =g-\frac{1}{M} \frac{x_{3}^{2}}{x_{1}} \\ \frac{d x_{3}}{d t} & =-\frac{R}{L} x_{3}+\frac{1}{L} e\end{cases}
$$

To remove the non-linear terms in teh right hand side we consider an equilibrium state with $y_{0}(t)=a$ is constant. Then, $\frac{d^{i} y_{0}(t)}{d t^{2}}=0$ and usbstitution in the differential equations gives that the equilibrium state determined by $a$ is

$$
x_{0}(t)=(a, 0, \sqrt{M g a})
$$

Expanding the nonlinear terms into Taylor series about $x_{0}(t)$ and neglecting higher order terms we can approximate the system by the linear time-invariant system $\Sigma: \frac{d x^{\prime}}{d t}=$ $B x^{\prime}+A$ where $x_{i}^{\prime}(t)=x_{i}(t)-x_{0 i}(t)$ and

$$
B=\left[\begin{array}{ccc}
0 & 1 & 0 \\
\frac{g}{a} & 0 & -2 \sqrt{\frac{g}{M a}} \\
0 & 0 & -\frac{R}{L}
\end{array}\right] \quad \text { and } \quad A=\left[\begin{array}{c}
0 \\
0 \\
\frac{1}{L}
\end{array}\right]
$$

The control matrix $c(\Sigma)$ is teh $3 \times 3$ matrix

$$
\left[A, B A, B^{2} A\right]=\left[\begin{array}{ccc}
0 & 0 & -\frac{2}{L} \\
0 & -\frac{2}{L} \sqrt{\frac{g}{M a}} & \frac{2 R}{L^{2}} \sqrt{\frac{g}{M a}} \\
\frac{1}{L} & -\frac{R}{L^{2}} & \frac{R^{2}}{L^{3}}
\end{array}\right]
$$

which is of rank 3 hence $\Sigma$ is completely controllable. Observability of the system depends on which variable we define as the output. First, assume that the output signal is the distance $x_{1}$ from teh ball to the magnet, that is, $C=[1,0,0]$. Then, teh observation matrix $o(\Sigma)$ is equal to

$$
\left[\begin{array}{c}
C \\
C B \\
C^{2} B
\end{array}\right]=\left[\begin{array}{ccc}
1 & 0 & 0 \\
0 & 1 & 0 \\
\frac{g}{a} & 0 & -2 \sqrt{\frac{g}{M a}}
\end{array}\right]
$$

and the system is completely controllable. Similarly, if the output is teh speed $x_{2}$ of the ball then one verifies complete controllability. However, if our output signal is the current $i$, that is, $C=[0,0,1]$ then the controllability matrix is

$$
\left[\begin{array}{c}
C \\
C B \\
C^{2} B
\end{array}\right]=\left[\begin{array}{ccc}
0 & 0 & 1 \\
0 & 0 & -\frac{R}{L} \\
0 & 0 & 2 \frac{R}{L} \sqrt{\frac{g}{M a}}
\end{array}\right]
$$

Thus, the system is not completely controllable. That is, observing only the current $i(t)$ we are not always able to reconstruct the state of teh system. If however, we would observe the distance $y(t)$ (or the speed $\frac{d y}{d t}$ ) we are able to reconstruct the state of the system.

## Week 5

## The orbit space.

Usually, a system is a black box, that is, its inner workings are unknown to us and we can only detect its input/output behavior. Let us restrict to linear time-invariant dynamical systems which are both completely controllable and completely observable and call such systems Schurian. Of fundamental importance in system theory is the solution to :

## Problem 4.

Classify Schurian dynamical systems with the same input/output behavior.

First, we will reduce this problem to the study of $G L_{n}$-orbits in an open subset of a certain vectorspace. Assume we have two systems $\Sigma$ and $\Sigma^{\prime}$, determined by matrix triples from Sys $=M_{n \times m}(\mathbb{C}) \times M_{n}(\mathbb{C}) \times M_{p \times n}(\mathbb{C})$

producing the same output $y(t)$ when given the same input $u(t)$, for all possible input functions $u(t)$. We recall that the output function $y$ for a system $\Sigma=(A, B, C)$ is determined by

$$
y(\tau)=C e^{B\left(\tau-\tau_{0}\right)} x\left(\tau_{0}\right)+\int_{\tau_{0}}^{\tau} C e^{(\tau-t) B} A u(t) d t
$$

Differentiating this a number of times and evaluating at $\tau=\tau_{0}$ as in the proof of proposition ?? equality of input/output for $\Sigma$ and $\Sigma^{\prime}$ gives the conditions

$$
C B^{i} A=C^{\prime} B^{\prime i} A^{\prime} \quad \text { for all } i
$$

## 5a. The Kalman code.

As a consequence the systems $\Sigma$ and $\Sigma^{\prime}$ have the same Hankel matrix which by definition is the product of the observation matrix with the control matrix of the system :

Alternatively, we can express this condition in terms of linear maps. Consider the two compositions

$$
\left\{\begin{array}{l}
\mathbb{C}^{m n} \xrightarrow{c(\Sigma)} \stackrel{\mathbb{C}^{n} \xlongequal{o(\Sigma)} \mathbb{C}^{p n}}{\mathbb{C}^{m n} \xrightarrow{c\left(\Sigma^{\prime}\right)}} \mathbb{C}^{n} \xrightarrow{o\left(\Sigma^{\prime}\right)} \mathbb{C}^{p n}
\end{array}\right.
$$

Here, the control maps are onto by complete controllability and the observation maps are into by complete observability. Equality of input/output implies equality of the Hankel matrices and so the composed linear maps $\mathbb{C}^{m n} \longrightarrow \mathbb{C}^{p n}$ are equal.

But then, we have for any $v \in \mathbb{C}^{m n}$ that $c(\Sigma)(v)=0 \Leftrightarrow c\left(\Sigma^{\prime}\right)(v)$ and we can decompose $\mathbb{C}^{p n}=V \oplus W$ such that the restriction of $c(\Sigma)$ and $c\left(\Sigma^{\prime}\right)$ to $V$ are the zero map and the restrictions to $W$ give isomorphisms with $\mathbb{C}^{n}$. Hence, there is an invertible matrix $g \in G L_{n}$ such that $c\left(\Sigma^{\prime}\right)=g c(\Sigma)$ and from the commutative diagram

we obtain that also $o\left(\Sigma^{\prime}\right)=o(\Sigma) g^{-1}$.
Consider the system $\Sigma_{1}=\left(A_{1}, B_{1}, C_{1}\right)$ equivalent with $\Sigma$ under the basechange matrix $g$. That is, $\Sigma_{1}=g . \Sigma=\left(g A, g B g^{-1}, C g^{-1}\right)$. Then,

$$
\left[A_{1}, B_{1} A_{1}, \ldots, B_{1}^{n-1} A_{1}\right]=g c(\Sigma)=c\left(\Sigma^{\prime}\right)=\left[A^{\prime}, B^{\prime} A^{\prime}, \ldots, B^{\prime n-1} A^{\prime}\right]
$$

and so $A_{1}=A^{\prime}$. Further, as $B_{1}^{i+1} A_{1}=B^{\prime i+1} A^{\prime}$ we have by induction on $i$ that the restriction of $B_{1}$ on the subspace of $B^{\prime i} \operatorname{Im}\left(A^{\prime}\right)$ is equal to the
restriction of $B^{\prime}$ on this space. Moreover, as $\sum_{i=0}^{n-1} B^{\prime i} \operatorname{Im}\left(A^{\prime}\right)=\mathbb{C}^{n}$ it follows that $B_{1}=B^{\prime}$. Because $o\left(\Sigma^{\prime}\right)=o(\Sigma) g^{-1}$ we also have $C_{1}=C^{\prime}$. This finishes the proof of :

Proposition 5.1. Let $\Sigma=(A, B, C)$ and $\Sigma^{\prime}=\left(A^{\prime}, B^{\prime}, C^{\prime}\right)$ be two Schurian dynamical systems. The following are equivalent

1. The input / output behavior of $\Sigma$ and $\Sigma^{\prime}$ are equal.
2. The systems $\Sigma$ and $\Sigma^{\prime}$ are equivalent, that is, there exists an invertible matrix $g \in G L_{n}$ such that

$$
A^{\prime}=g A, \quad B^{\prime}=g B g^{-1} \quad \text { and } \quad C^{\prime}=C g^{-1} .
$$

This proposition reduces the system theoretic problem to a fabric setting. We consider the linear action of $G L_{n}$ on the vectorspace of matrix triples defining linear time-invariant dynamical systems

$$
\text { Sys }=M_{n \times m}(\mathbb{C}) \times M_{n}(\mathbb{C}) \times M_{p \times n}(\mathbb{C})
$$

defined for all $g \in G L_{n}$ by

$$
g \cdot(A, B, C)=\left(g A, g B g^{-1}, C g^{-1}\right)
$$

For later reference we depict this action by the pattern


By definition, a dynamical system $\Sigma=(A, B, C)$ is Schurian if (and only if) the determinant of at least one $n \times n$ minor of $c(\Sigma)$ and $o(\Sigma)$ is non-zero. That is, the subset Sys ${ }^{s}$ of Schurian dynamical systems is open in Sys and is stable under the $G L_{n}$-action. Our next job is to classify the orbits under this action.

We introduce a combinatorial gadget : the Kalman code. It is an array consisting of $(n+1) \times m$ boxes each having a position label $(i, j)$ where $0 \leq i \leq n$ and $1 \leq j \leq m$. These boxes are ordered lexicographically that is $\left(i^{\prime}, j^{\prime}\right)<(i, j)$ if and only if either $i^{\prime}<i$ or $i^{\prime}=i$ and $j^{\prime}<j$. Exactly $n$ of these boxes are painted black subject to the rule that if box $(i, j)$ is black,
then so is box $\left(i^{\prime}, j\right)$ for all $i^{\prime}<i$. That is, a Kalman code looks like


We assign to a completely controllable system $\Sigma=(A, B, C)$ its Kalman code $K(\Sigma)$ as follows : let $A=\left[\begin{array}{llll}A_{1} & A_{2} & \ldots & A_{m}\end{array}\right]$, that is $A_{i}$ is the $i$-th column of $A$. Paint the box $(i, j)$ black if and only if the column vector $B^{i} A_{j}$ is linearly independent of the column vectors $B^{k} A_{l}$ for all $(k, l)<(i, j)$. The painted array $K(\Sigma)$ is indeed a Kalman code. Assume that box $(i, j)$ is black but box $\left(i^{\prime}, j\right)$ white for $i^{\prime}<i$, then

$$
B^{i^{\prime}} A_{j}=\sum_{(k, l)<\left(i^{\prime}, j\right)} \alpha_{k l} B^{k} A_{l} \quad \text { but then, } \quad B^{i} A_{j}=\sum_{(k, l)<\left(i^{\prime}, j\right)} \alpha_{k l} B^{k+i-i^{\prime}} A_{l}
$$

and all $\left(k+i-i^{\prime}, l\right)<(i, l)$, a contradiction. Moreover, $K(\Sigma)$ has exactly $n$ black boxes as there are $n$ linearly independent columns of the control matrix $c(\Sigma)$ when $\Sigma$ is completely controllable.

The Kalman code is a discrete invariant of the orbit $\mathcal{O}_{\Sigma}$ under the action of $G L_{n}$. This follows from the fact that $B^{i} A_{j}$ is linearly independent of the $B^{k} A_{l}$ for all $(k, l)<(i, j)$ if and only if $g B^{i} A_{j}$ is linearly independent of the $g B^{k} A_{l}$ for any $g \in G L_{n}$ and the observation that $g B^{k} A_{l}=\left(g B g^{-1}\right)^{k}(g A)_{l}$. Next, we will clarify the geometric significance of the Kalman code.

As the Kalman code depends only on the input part $(A, B)$ of the system $\Sigma=(A, B, C)$ we consider the linear action of $G L_{n}$ on the vectorspace of matrixpairs $V=M_{n \times m}(\mathbb{C}) \times M_{n}(\mathbb{C})$ defined by $g \cdot(A, B)=\left(g A, g B g^{-1}\right)$.

With $V_{c}$ we will denote the open subset of all completely controllable pairs $(A, B)$ that is, those for which the rank of the $n \times n m$ matrix $\left[\begin{array}{lllll}A & B A & B^{2} A & \ldots & B^{n-1} A\end{array}\right]$ is maximal. We consider the map

$$
\begin{aligned}
V=M_{n \times m}(\mathbb{C}) \times M_{n}(\mathbb{C}) & \xrightarrow{\psi} \\
(A, B) & \mapsto\left[\begin{array}{llllll}
A & B A & B^{2} A & \ldots & B^{n-1} A & B^{n} A
\end{array}\right]
\end{aligned}
$$

The matrix $\psi(A, B)$ determines a linear map $\psi_{(A, B)}: \mathbb{C}^{(n+1) m} \longrightarrow \mathbb{C}^{n}$ and $(A, B)$ is a completely controllable pair if and only if the corresponding
linear map $\psi_{(A, B)}$ is surjective. Moreover, there is a linear action of $G L_{n}$ on $M_{n \times(n+1) m}(\mathbb{C})$ by left multiplication and the map $\psi$ is $G L_{n}$-equivariant meaning that $\psi(g .(A, B))=g \psi(A, B)$.

The Kalman code allows us to find a canonical pair in the orbit $\mathcal{O}_{(A, B)}$ when $(A, B)$ is a completely controllable pair. There is a natural one-toone correspondence between the boxes in the Kalman code array and the columns of $\psi(A, B)$ by sending box $(i, j)$ to the $j$-th column of the submatrix $B^{i} A$.

## 5b. Grassman manifolds.

The Kalman code induces a barcode on $\psi(A, B)$, that is the $n \times n$ minor of $\psi(A, B)$ determined by the columns corresponding to black boxes in the Kalman code.


By construction this minor is an invertible matrix $g^{-1} \in G L_{n}$. The canonical element in the orbit $\mathcal{O}_{(A, B)}$ we have in mind is the pair $g \cdot(A, B)$. It has the characteristic property that the $n \times n$ minor of its image under $\psi$, determined by the Kalman code is the identity matrix $\mathbb{1}_{n}$. The matrix $\psi(g .(A, B))$ will be denoted by $b(A, B)$ and is called barcode of the pair $(A, B)$. We now claim that the barcode determines the orbit uniquely.

In fact, the map $\psi$ is injective on the open set $V_{c}$ of completely controllable pairs. Indeed, if

$$
\left[\begin{array}{llll}
A & B A & \ldots & B^{n} A
\end{array}\right]=\left[\begin{array}{llll}
A^{\prime} & B^{\prime} A^{\prime} & \ldots & B^{\prime n} A^{\prime}
\end{array}\right]
$$

then $A=A^{\prime}, B\left|\operatorname{Im}(A)=B^{\prime}\right| \operatorname{Im}(A)$ and hence by induction also

$$
B\left|B^{i} \operatorname{Im}(A)=B^{\prime}\right| B^{\prime i} \operatorname{Im}\left(A^{\prime}\right) \quad \text { for all } i \leq n-1
$$

But then, $B=B^{\prime}$ as both pairs $(A, B)$ and $\left(A^{\prime}, B^{\prime}\right)$ are completely controllable, that is, $\sum_{i=0}^{n-1} B^{i} \operatorname{Im}(A)=\mathbb{C}^{n}=\sum_{i=0}^{n-1} B^{\prime i} \operatorname{Im}\left(A^{\prime}\right)$. Hence, the barcode $b(A, B)$ determines the orbit $\mathcal{O}_{(A, B)}$ and is a point in the Grassman manifold $\operatorname{Gras}_{n}\left(\mathbb{C}^{m(n+1)}\right)$.

We recall briefly the definition of the Grassman manifolds. Let $k \leq l$ be integers, then the points of Grassman manifold $\operatorname{Gras}_{k}\left(\mathbb{C}^{l}\right)$ are in one-to-one
correspondence with $k$-dimensional subspaces of $\mathbb{C}^{l}$. For example, if $k=1$ then $\operatorname{Gras}_{1}\left(\mathbb{C}^{l}\right)$ is the projective $l-1$-space $\mathbb{P}^{l-1}$. We know that projective space can be covered by affine spaces defining a manifold structure on it. Also Grassman manifold admit a cover by affine spaces.

Let $W$ be a $k$-dimensional subspace of $\mathbb{C}^{l}$ then fixing a basis $\left\{w_{1}, \ldots, w_{k}\right\}$ of $W$ determines an $k \times l$ matrix $M$ having as $i$-th row the coordinates of $w_{i}$ with respect to the standard basis of $\mathbb{C}^{l}$. Linear independence of the vectors $w_{i}$ means that there is a barcode design $I$ on $M$

where $I=1 \leq i_{1}<i_{2}<\ldots<i_{k} \leq l$ such that the corresponding $k \times k$ minor $M_{I}$ of $M$ is invertible. Observe that $M$ can have several such designs.

Conversely, given a $k \times l$ matrix $M$ of rank $k$ determines a $k$-dimensional subspace of $l$ spanned by the transposed rows. Two $k \times l M$ and $M^{\prime}$ matrices of rank $k$ determine the same subspace provided there is a basechange matrix $g \in G L_{k}$ such that $g M=M^{\prime}$. That is, we can identify $\operatorname{Grass}_{k}\left(\mathbb{C}^{l}\right)$ with the orbit space of the linear action of $G L_{k}$ by left multiplication on the open set $M_{k \times l}^{\max }(\mathbb{C})$ of $M_{k \times l}(\mathbb{C})$ of matrices of maximal rank. Let $I$ be a barcode design and consider the subset of $\operatorname{Grass}_{k}\left(\mathbb{C}^{l}\right)(I)$ of subspaces having a matrix representation $M$ having $I$ as barcode design. Multiplying on the left with $M_{I}^{-1}$ the $G L_{k}$-orbit $\mathcal{O}_{M}$ has a unique representant $N$ with $N_{I}=\mathbb{1}_{k}$. Conversely, any matrix $N$ with $N_{I}=1_{k}$ determines a point in $\operatorname{Grass}_{k}\left(\mathbb{C}^{l}\right)(I)$. Thus, $\operatorname{Grass}_{k}\left(\mathbb{C}^{l}\right)(I)$ depends on $k(l-k)$ free parameters (the entries of the negative of the barcode)

and we have an identification $\operatorname{Grass}_{k}\left(\mathbb{C}^{l}\right) \xrightarrow{\pi_{I}} \mathbb{C}^{k(l-k)}$. For a different barcode design $I^{\prime}$ the image $\pi_{I}\left(\operatorname{Grass}_{k}\left(\mathbb{C}^{l}\right)(I) \cap \operatorname{Grass}_{k}\left(\mathbb{C}^{l}\right)\left(I^{\prime}\right)\right)$ is an open subset of $\mathbb{C}^{k(l-k)}$ (one extra nonsingular minor condition) and $\pi_{I^{\prime}} \circ \pi_{I}^{-1}$ is a diffeomorphism on this set. That is, the maps $\pi_{I}$ provide us with an atlas and determine a manifold structure on $\operatorname{Grass}_{k}\left(\mathbb{C}^{l}\right)$.

Applying the foregoing construction, the barcode $b(A, B)$ determined by the Kalman code determines a unique point in $\operatorname{Grass}_{n}\left(\mathbb{C}^{m(n+1)}\right)$. We have
the following diagram

where $\psi$ is a $G L_{n}$-equivariant embedding and $\chi$ the orbit map. Observe that both $\psi$ and 'chi are clearly continuous maps, hence so is the orbit map $b$. Observe that the barcode matrix $b(A, B)$ shows that the stabilizer of $(A, B)$ is trivial. Indeed, the minor of $g . b(A, B)$ determined by the Kalman code is equal to $g$. Moreover, continuity of $b$ implies that the orbit $\mathcal{O}_{(A, B)}$ is closed in $V_{c}$.

Our final aim is to prove that $\psi$ is a diffeomorphism to a locally closed submanifold of $M_{n \times m(n+1)}(\mathbb{C})$. To prove this we have to consider the differential of $\psi$. We recall briefly the definition of a differential. Consider a map

$$
f=\left(f_{1}, \ldots, f_{l}\right): \mathbb{C}^{k} \longrightarrow \mathbb{C}^{l}
$$

with all the $f_{i}$ differentiable complex valued maps in the coordinate functions $x_{i}$ of $\mathbb{C}^{k}$. For a point $p \in \mathbb{C}^{k}$ the differential of $f$ at $p$ is a linear map

$$
d f_{p}: T_{p} \mathbb{C}^{k} \simeq \mathbb{C}^{k} \longrightarrow \mathbb{C}^{l}=T_{f(p)} \mathbb{C}^{l}
$$

between the tangent space to $\mathbb{C}^{k}$ in $p$ and the tangent space to $\mathbb{C}^{l}$ in the image $f(p)$. This linear map is determined by the matrix

$$
\left[\begin{array}{ccc}
\frac{\partial f_{1}}{\partial x_{1}}(p) & \cdots & \frac{\partial f_{1}}{\partial x_{k}}(p) \\
\vdots & \ddots & \vdots \\
\frac{\partial f_{l}}{\partial x_{1}}(p) & \cdots & \frac{\partial f_{l}}{\partial x_{k}}(p)
\end{array}\right] \in M_{l \times k}(\mathbb{C})
$$

When all the $f_{i}$ are polynomials in the variables $x_{i}$ we can compute the differential map by the $\epsilon$-method : compute $f_{i}(p+\epsilon v)$ bearing in mind that $\epsilon^{2}=0$, then one has

$$
f(p+\epsilon v)=f(p)+\epsilon d f_{p}(v) \quad \text { for all } v \in T_{p} \mathbb{C}^{k} .
$$

If $d f_{p}$ is injective, then the implicit function theorem implies that the image of $f$ is locally around $f(p)$ a closed submanifold of dimension $n$ in $\mathbb{C}^{l}$.

All the coordinate functions of $\psi$ are polynomials in the coordinates of $W$. For all $(A, B) \in W$ and $(X, Y) \in T_{(A, B)}(W) \simeq W$ we have

$$
(B+\epsilon Y)^{j}(A+\epsilon X)=B^{n} A+\epsilon\left(B^{n} X+\sum_{i=0}^{j-1} B^{i} Y B^{n-1-i} A\right)
$$

Therefore the differential of $\psi$ in $(A, B) \in W, d \psi_{(A, B)}(X, Y)$ is equal to

$$
\left[\begin{array}{lllll}
X & B X+Y A & B^{2} X+B Y A+Y B A & \ldots & B^{n} X+\sum_{i=0}^{n-1} B^{i} Y B^{n-1-i} A
\end{array}\right]
$$

Assume $d \psi_{(A, B)}(X, Y)$ is the zero matrix, then $X=0$ and substituting in the next term also $Y A=0$. Substituting in the third gives $Y B A=0$, then in the fourth $Y B^{2} A=0$ and so on until $Y B^{n-1} A=0$. But then,

$$
Y\left[\begin{array}{lllll}
A & B A & B^{2} A & \ldots & B^{n-1} A
\end{array}\right]=0 .
$$

If $(A, B)$ is a completely controllable pair, this implies that $Y=0$ and hence shows that $d \psi_{(A, B)}$ is injective for all $(A, B) \in V_{c}$.

By the implicit function theorem, $\psi$ induces a $G L_{n}$-equivariant diffeomorphism between the open subset $V_{c}$ of completely controllable pairs and a locally closed submanifold of $M_{n \times(n+1) m}(\mathbb{C})^{\max }$. The image of this submanifold under the orbit map $\chi$ is again a manifold as all fibers are equal to $G L_{n}$. This concludes the difficult part of the Kalman theorem :

Theorem 5.2. The orbit space $O_{c}=V_{c} / G L_{n}$ of equivalence classes of completely controllable pairs is a locally closed submanifold of dimension m.n of the Grassman manifold $\operatorname{Grass}_{n}\left(\mathbb{C}^{m(n+1)}\right)$. In fact $V_{c} \xrightarrow{b} O_{c}$ is a principal $G L_{n}$-bundle.

To prove the dimension statement, consider $V_{c}(K)$ the set of completely controllable pairs $(A, B)$ having Kalman code $K$ and let $O_{c}(K)$ be the image under the orbit map. After identifying $V_{c}(K)$ with its image under $\psi$, the barcode matrix $b(A, B)$ gives a section $O_{c}(K) \stackrel{s}{\hookrightarrow} V_{c}(K)$. In fact,

$$
\begin{gathered}
G L_{n} \times O_{c}(K) \longrightarrow V_{c}(K) \\
(g, x) \mapsto g \cdot s(x)
\end{gathered}
$$

is a $G L_{n}$-equivariant diffeomorphism because the $n \times n$ minor determined by $K$ of $g . b(A, B)$ is $g$. Apply the local product decomposition to the generic Kalman code $K^{g}$

obtained by painting the top boxes black from left to right until one has $n$ black boxes. Clearly $V_{c}\left(K^{g}\right)$ is open in $V_{c}$ and one deduces

$$
\operatorname{dim} O_{c}=\operatorname{dim} O_{c}\left(K^{g}\right)=\operatorname{dim} V_{c}\left(K^{g}\right)-\operatorname{dim} G L_{n}=m n+n^{2}-n^{2}=m n
$$

The Kalman theorem 5.2. implies immediately the existence of an orbit space for completely controllable and Schurian systems. Indeed, let $\Sigma=(A, B, C)$ completely controllable and let $g=g_{(A, B)} \in G L_{n}$ be the uniquely determined basechange such that $g \cdot(A, B)=b(A, B)$, then we have a canonical representant $\left(g A, g B g^{-1}, C g^{-1}\right)$ in the orbit $\mathcal{O}_{\Sigma}$. As the stabilizer $\operatorname{Stab}(A, B)$ is trivial the orbits of $(A, B, C)$ and $\left(A, B, C^{\prime}\right)$ are distinct if $C=C^{\prime}$. That is the natural projection $p r_{3}$

descends to define an orbit space which is an $M_{p \times n}(\mathbb{C})$ - bundle over $O_{c}$ and hence is a manifold. The Schurian systems Sys $_{s}$ form a $G L_{n}$-stable open subset of Sys ${ }_{c}$ and hence their orbit space is an open submanifold of $S y s_{c} / G L_{n}$. This concludes the solution to problem 5:

Theorem 5.3. Let $S y s_{c}\left(\right.$ resp. $\left.S y s_{s}\right)$ the the open subset of $S y s=M_{n \times m}(\mathbb{C}) \times$ $M_{n}(\mathbb{C}) \times M_{p \times n}(\mathbb{C})$ determined by the completely controllable (resp. Schurian) linear dynamical systems.

1. The orbit space for the $G L_{n}$ action on $S y s_{c}$ exists and is a vectorbundle of rank pn over $O_{c}$.
2. The orbit space for the $G L_{n}$-action on $S y s_{s}$ exists and is a manifold of dimension $m n^{2} p$.

## Week 6

## The slice space.

Another important problem in system theory is to determine how a system can change under small perturbations. That is, given a completely controllable pair $(A, B)$ or system $\Sigma=(A, B, C)$, we want to construct a slice giving unique representants in nearby orbits.

## Problem 5.

> Construct a slice for completely controllable systems.

Consider first the case of a controllable pair $p=(A, B)$. By a slice we mean a submanifold $S$ of $V=M_{n \times m}(\mathbb{C}) \times M_{n}(\mathbb{C})$ passing through $(A, B)$ and such that we have a $G L_{n}$-equivariant diffeomorphism

$$
G L_{n} \times S \longrightarrow V \quad \text { given by } \quad(g, s) \mapsto g . s
$$

in a neighborhood of $(A, B)$. Hence, near $(A, B)$ an orbit intersects $S$ uniquely. We have the following situation


## 6a. Slice representation.

In order to determine a good candidate for $S$ let us compute the differential of the action map

$$
G L_{n} \times V \xrightarrow{a} V \quad \text { defined by } \quad\left(g,\left(A^{\prime}, B^{\prime}\right)\right) \mapsto\left(g A^{\prime}, g B g^{-1}\right)
$$

in the point $\left(\mathcal{L}_{n}, p\right)$ where $p=(A, B)$. Because $G L_{n}$ is an open submanifold of $M_{n}(\mathbb{C})$ we know that $T_{1_{n}} G L_{n}=M_{n}(\mathbb{C})=\left\{\tau_{n}+\epsilon m \mid m \in M_{n}(\mathbb{C})\right\}$. Similarly, the tangent space $T_{p} V=V$ and can be identified with $\{(A+\epsilon X, B+\epsilon Y) \mid$ $\left.X \in M_{n \times m}(\mathbb{C}), Y \in M_{n}(\mathbb{C})\right\}$. By the $\epsilon$-method to compute the differential we have to expand

$$
\left(\left(\mathbb{1}_{n}+\epsilon m\right)(A+\epsilon X),\left(\mathbb{1}_{n}+\epsilon m\right)(B+\epsilon Y)\left(\mathbb{1}_{n}-\epsilon m\right)\right)
$$

which is equal to

$$
(A, B)+\epsilon(X+m A, Y+[m, B]) \quad \text { where } \quad[m, B]=m B-B m
$$

That is, the differential of the action map in the point $\left(1_{n}, p\right)$ is given by

$$
d a_{\left(\mathbb{T}_{n}, p\right)}(m,(X, Y))=(X+m A, Y+[m, B]) .
$$

Likewise, we can compute that the differential of the orbit map $G L_{n} \xrightarrow{o} V$ of $\mathcal{O}_{(A, B)}$ defined by $g \mapsto g \cdot(A, B)$ in the point $\mathbb{1}_{n}$ is given by the linear map

$$
d o_{\chi_{n}}(m)=(m A,[m, B]) .
$$

Observe that $d o_{1_{n}}$ is injective. For if $m A=0$ and $m B=B m$ then for all $i$ we have that $m B^{i} A=B^{i} m A=0$ whence $m\left[\begin{array}{llll}A & B A & \ldots & B^{n-1} A\end{array}\right]=0$. But then $m=0$ by complete controllability of $(A, B)$.

The tangent space $T_{p} \mathcal{O}_{(A, B)}$ in $p=(A, B)$ to the orbit is the subset of pairs $(A, B)+\operatorname{Im} d o_{\mathbb{1}_{n}}=\left\{(A+m A, B+[m, B]) \mid m \in M_{n}(\mathbb{C})\right\}$. Take $S$ to be the normal space in $p$ to the orbit $\mathcal{O}_{(A, B)}$. That is, $S=(A, B)+\left(\operatorname{Im} d o_{\mathcal{\chi}_{n}}\right)^{\perp}$ where $\left(I m d o_{\rrbracket_{n}}\right)^{\perp}$ is subspace of $V$ orthogonal to $I m d o_{\rrbracket_{n}}$.


Here, orthogonality is meant with respect to the Hermitian inner product on $V=M_{n \times m}(\mathbb{C}) \times M_{n}(\mathbb{C})$ defined by

$$
\left\langle\left(A_{1}, B_{1}\right),\left(A_{2}, B_{2}\right)\right\rangle=\operatorname{Tr}\left(A_{1} \bar{A}_{2}^{t r}\right)+\operatorname{Tr}\left(B_{1} \bar{B}_{2}^{t r}\right)
$$

where $T r$ is the trace on $n \times n$ matrices and $\bar{X}^{t r}$ is the Hermitian transpose, that is if $U=\left(u_{i j}\right)_{i, j}$, then $\bar{U}^{t r}=\left(\bar{u}_{i j}\right)_{j, i}$. Now, $\left(\operatorname{Im~do}{\chi_{n}}\right)^{\perp}$ is the subspace of pairs $(X, Y)$ such that for all $m \in M_{n}(\mathbb{C})$ we have

$$
\begin{aligned}
0=\langle(m A,[m, B]),(X, Y)\rangle= & \operatorname{Tr}\left(m A \bar{X}^{t r}\right)+\operatorname{Tr}\left([m, B] \bar{Y}^{t r}\right) \\
= & \operatorname{Tr}\left(m A \bar{X}^{t r}\right)+\operatorname{Tr}\left(m\left[B, \bar{Y}^{t r}\right]\right) \\
& =\operatorname{Tr}\left(m\left(A \bar{X}^{\operatorname{tr}}+\left[B, \bar{Y}^{t r}\right]\right)\right)
\end{aligned}
$$

where we used that traces of products are the same for cyclic permutations of the factors. By the non-degeneracy of the trace on $M_{n}(\mathbb{C})$ (that is, $n=0$ if and only if $\operatorname{Tr}(m n)=0$ for all $m \in M_{n}(\mathbb{C})$ ) we conclude from this that

$$
\left(\operatorname{Im} d o_{\uparrow_{n}}\right)^{\perp}=\left\{(X, Y) \mid A \bar{X}^{t r}+\left[B, \bar{Y}^{t r}\right]=0\right\}
$$

Theorem 6.1. Let $p=(A, B)$ be a completely controllable pair. Then the manifold

$$
S=\left\{(A+X, B+Y) \mid A \bar{X}^{t r}+\left[B, \bar{Y}^{t r}\right]=0\right\}
$$

is a slice for the action of $G L_{n}$ on $V$ in $p=(A, B)$.
By the implicit function theorem the action map $G L_{n} \times S \xrightarrow{a} V$ is a diffeomorphism in $p=(A, B)$ if and only if the differential

$$
d a_{\left(\mathbb{1}_{n}, p\right)}: T_{\mathbb{1}_{n}} G L_{n} \oplus T_{p} S \longrightarrow T_{p} V=V
$$

is a linear isomorphism. We have seen that $T_{\uparrow_{n}} G L_{n}=M_{n}(\mathbb{C})=\operatorname{Im} d o_{\uparrow_{n}}$ and that $T_{p} S=\left(\operatorname{Im} d o_{\rrbracket_{n}}\right)^{\perp}$ so the dimensions of both left factors add up to $\operatorname{dim} V$. It therefore suffices to check injectivity of the differential. We have calculated that

$$
d a_{\left(\mathbb{1}_{n}, p\right)}(m,(X, Y))=(m A+X,[m, B]+Y)
$$

Assume that $X=-m A$ and $Y=-[m, B]$ then because $(X, Y) \in\left(I m d o_{\rrbracket_{n}}\right)^{\perp}$ we have $\langle(X, Y),(X, Y)\rangle=0$ whence $X=0$ and $Y=0$. But then as $m A=0$ and $m B=B m$ and $(A, B)$ is completely controllable we also have $m=0$ finishing the proof of the slice theorem.

## 6b. Control canonical form.

The determination of the slice $S$ is particularly simple in the point $\left(A^{\prime}, B^{\prime}\right)$ of the orbit $\mathcal{O}_{(A, B)}$ where $\left(A^{\prime}, B^{\prime}\right)$ is in control canonical form. This canonical form is similar to the Jordan normal form and is determined by the Kalman code $K$ associated to the completely controllable pair $(A, B)$. Assume $K$ has the following shape :


Here, $1 \leq j_{1}<\ldots<j_{t} \leq m$ form the set $J$ of integers $j$ such that the box $(0, j)$ is painted black, hence $t=\# J=\operatorname{rank} A$. For each $j_{k} \in J$ define $b_{k}$ to be the number of painted boxes in the $j$-th column of $K$.

With notations as above, prove that there is a base-change matrix $g \in$ $G L_{n}$ such that $g \cdot(A, B)=\left(A^{\prime}, B^{\prime}\right)$ where the $n \times n$ matrix $B^{\prime}$ has the following block form

where the nonzero entries are in one of the identity matrix components $\prod_{b_{k}-1}$ or in the painted stripes which represent one row. The off-diagonal
block $B_{i j}$ is a $b_{i} \times b_{j}$ matrix of one of the following types

depending on whether $b_{i} \leq b_{j}$ or $b_{j} \leq b_{i}$. The diagonal block $B_{i i}$ is a $b_{i} \times b_{i}$ matrix of the form

$$
\mathbb{1}_{b_{i}-1}
$$

The $n \times m$ matrix $A$ has the following block structure :


## 6c. Another slice result.

Finally, we will generalize the slice construction to dynamical systems $\Sigma=$ $(A, B, C)$ which are completely controllable. We define an Hermitian inner product on Sys $=M_{n \times m}(\mathbb{C}) \times M_{n}(\mathbb{C}) \times M_{p \times n}(\mathbb{C})$ by the rule

$$
\left\langle\left(A_{1}, B_{1}, C_{1}\right),\left(A_{2}, B_{2}, C_{2}\right)\right\rangle=\operatorname{Tr}\left(A_{1} \bar{A}_{2}^{t r}\right)+\operatorname{Tr}\left(B_{1} \bar{B}_{2}^{t r}\right)+\operatorname{Tr}\left(\bar{C}_{2}^{t r} C_{1}\right)
$$

This time, the differential of the orbit map $G L_{n} \xrightarrow{o}$ Sys defined by $g \mapsto$ $g .(A, B, C)$ is computed by expanding the expression

$$
\left(\left(\mathbb{1}_{n}+\epsilon m\right)(A+\epsilon X),\left(\mathbb{T}_{n}+\epsilon m\right)(B+\epsilon Y)\left(\mathbb{T}_{n}-\epsilon m\right),(C+\epsilon Z)\left(\mathbb{T}_{n}-\epsilon m\right)\right)
$$

and we obtain that the image of the orbit map $d o_{\rrbracket_{n}}$ consists of the triples

$$
\operatorname{Im} d o_{{\iota_{n}}}=\left\{(m A,[m, B],-C m) \mid m \in M_{n}(\mathbb{C})\right\} .
$$

Again, using non-degeneracy of the trace map on $n \times n$ matrices one can identify the orthogonal complement of this space as the subspace

$$
\left(I m d o_{\varpi_{n}}\right)^{\perp}=\left\{(X, Y, Z) \mid A \bar{X}^{t r}+\left[B, \bar{Y}^{t r}\right]-\bar{Z}^{t r} C=0\right\}
$$

From this we again construct a slice as in the proof of theorem 6.1.
Theorem 6.2. Let $\Sigma=(A, B, C)$ be a completely controllable system. Then the manifold

$$
S=\left\{(A+X, B+Y, C+Z) \mid A \bar{X}^{t r}+\left[B, \bar{Y}^{t r}\right]-\bar{Z}^{t r} C=0\right\}
$$

is a slice for the $G L_{n}$-action on Sys in $\Sigma$.

## Week 7

## Hilbert schemes.

Consider completely controllable pairs of matrices $(A, B)$ with $m=1$. That is, $A=v$ is a column vector and we must have that

$$
B(v)=\left[\begin{array}{lllll}
v & B v & B^{2} v & \ldots & B^{n-1} v
\end{array}\right]
$$

has rank $n$, or equivalently, $v$ is a cyclic vector of $B$. There is only one Kalman code $K(v, B)$ possible, the generic one. The barcode $b(v, B)$ of the pair $(v, B)$ is the matrix $B(v)^{-1}\left[\begin{array}{ll}B(v) & B^{n} v\end{array}\right]$. That is, we have :


Hence, the orbit space $O_{c}(1, n) \simeq \mathbb{C}^{n}$ and the point $\left(a_{1}, \ldots, a_{n}\right) \in \mathbb{C}^{n}$ corresponding to the orbit of $(v, B)$ is determined by the characteristic polynomial of $B$ :

$$
\chi_{B}(t)=\operatorname{det}\left(t \mathbb{T}_{n}-B\right)=t^{n}-a_{n} t^{n-1}-\ldots-a_{2} t-a_{1} .
$$

Hence, we can identify the orbit space $O_{c}(1, n)$ with the space of all monic polynomials in $\mathbb{C}[t]$ of degree $n$. Such polynomials $f(t)$ are uniquely determined by their unordered $n$-tuple of roots $\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}$ as

$$
f(t)=\prod_{i=1}^{n}\left(t-\lambda_{i}\right)=t^{n}-a_{n} t^{n-1}-\ldots-a_{2} t-a_{1}
$$

where $a_{i+1}=(-1)^{n-1} \sigma_{n-i}\left(\lambda_{1}, \ldots, \lambda_{n}\right)$ with $\sigma_{j}$ the $j$-th elementary symmetric function in the $\lambda_{k}$.

That is, we can also identify $O_{c}(1, n)$ with $\operatorname{Hilb}_{n}\left(\mathbb{C}^{1}\right)$, the Hilbert scheme of $n$ points in $\mathbb{C}^{1}$. The points of $\operatorname{Hilb}_{n}\left(\mathbb{C}^{1}\right)$ parametrize the ideals $I \triangleleft \mathbb{C}[t]$ of codimension $n$ and as $\mathbb{C}[t]$ is a principal ideal domain such ideals have a unique monic generator of degree $n$.

The symmetric product $S^{n} \mathbb{C}^{1}$ of $n$ copies of $\mathbb{C}^{1}$ is defined to be

$$
S^{n} \mathbb{C}^{1}=\underbrace{\mathbb{C}^{1} \times \ldots \times \mathbb{C}^{1}}_{n} / S_{n}
$$

where the symmetric group on $n$-letters $S_{n}$ acts on the entries. A point of $S^{n} \mathbb{C}^{1}$ is represented as a formal sum $\sum_{i} n_{i}\left[\mu_{i}\right]$ where $n_{i} \in \mathbb{N}$ with $\sum_{i} n_{i}=n$ and $\mu_{i} \in \mathbb{C}^{1}$. In this case, the Hilbert-Chow map

$$
\operatorname{Hilb}_{n} \mathbb{C}^{1} \xrightarrow{\pi} S^{n} \mathbb{C}^{1}
$$

defined by sending a monic polynomial $f(t)$ of degree $n$ to $\sum_{i} n_{i}\left[\mu_{i}\right]$ where the $\mu_{i}$ are the roots of $f(t)$ occurring with multiplicity $n_{i}$, is a one-to-one correspondence. Observe that the numbers $n_{i}$ determine (when ordered) a partition of $n$ and that the projection

$$
\underbrace{\mathbb{C}^{1} \times \ldots \times \mathbb{C}^{1}}_{n} \longrightarrow S^{n} \mathbb{C}^{1}
$$

is defined by sending an $n$-tuple $\left(c_{1}, \ldots, c_{n}\right)$ to $\left(s_{1}, \ldots, s_{n}\right)$ where $s_{i}=$ $\sigma_{i}\left(c_{1}, \ldots, c_{n}\right)$. So, we still have another description of the orbit space $O_{c}(1, n)$ as $S^{n} \mathbb{C}^{1}$. However, all these identifications are particular to dimension 1. Let us consider :

## Problem 6.

Describe $H i l b_{n} \mathbb{C}^{2}$, the Hilbert scheme of $n$ points in the plane $\mathbb{C}^{2}$. That is, parametrize all ideals $I \triangleleft \mathbb{C}[x, y]$ of codimension $n$.

Let $I \triangleleft \mathbb{C}[x, y]$ be such that $V=\frac{\mathbb{C}[x, y]}{I}$ is an $n$-dimensional vectorspace and fix a basis $\left\{v_{1}, \ldots, v_{n}\right\}$ of $V$. Multiplication by $x$ (resp. $y$ ) on $\mathbb{C}[x, y]$ induces a linear operator on $V$ and hence determines a matrix $X \in M_{n}(\mathbb{C})$ (resp. $Y \in M_{n}(\mathbb{C})$ ). Clearly, $[X, Y]=0$ and they generate an $n$-dimensional subalgebra $\mathbb{C}[X, Y] \simeq \frac{\mathbb{C}[x, y]}{I}$ of $M_{n}(\mathbb{C})$. Further, the image of the unit element $1 \in \mathbb{C}[x, y]$ determines a column vector $v \in V=\mathbb{C}^{n}$ with the property that

$$
\mathbb{C}[X, Y] v=\mathbb{C}^{n}
$$

Note however that the triple $(v, X, Y) \in \mathbb{C}^{n} \oplus M_{n} \oplus M_{n}$ is not uniquely determined by the ideal $I$ as it depends on the choice of basis of $V$. If we choose a different basis $\left\{v_{1}^{\prime}, \ldots, v_{n}^{\prime}\right\}$ with basechange matrix $g \in G L_{n}$, then the corresponding triple is

$$
\left(v^{\prime}, X^{\prime}, Y^{\prime}\right)=\left(g v, g X g^{-1}, g Y g^{-1}\right)
$$

Consider the vectorspace of all triples

$$
H_{n}=\mathbb{C}^{n} \oplus M_{n} \oplus M_{n} \quad \text { with action } \quad g \cdot(v, X, Y)=\left(g v, g X g^{-1}, g Y g^{-1}\right)
$$

for all $g \in G L_{n}$. For later reference, we depict this action by the pattern


The above discussion shows that the ideal $I \triangleleft \mathbb{C}[x, y]$ of codimension $n$ determines an orbit $\mathcal{O}_{I}$ in $H_{n}$. Conversely, let $C_{n}^{c}$ be the subset of triples $(v, X, Y) \in H_{n}$ satisfying the additional conditions :

1. The matrix pair commutes : $[X, Y]=0$, and
2. $v$ is a cyclic vector for this pair : $\mathbb{C}[X, Y] v=\mathbb{C}^{n}$.

For $(v, X, Y) \in C_{n}^{c}$ we can define a $\operatorname{map} \mathbb{C}[x, y] \xrightarrow{\phi} \mathbb{C}^{n}$ by sending a polynomial $f=f(x, y)$ to the vector $\phi(f)=f(X, Y) v$. By the second condition, $\phi$ is surjective and therefore, its kernel $I=\{f \in \mathbb{C}[x, y] \mid \phi(f)=0\}$ is an ideal of codimension $n$. That is, the Hilbert scheme $H i l b_{n} \mathbb{C}^{2}$ of $n$ points in the plane $\mathbb{C}^{2}$ is the orbit space for the $G L_{n}$-action on the subset $C_{n}^{c}$.

## 7a. An example.

## Miniature 5. The Hilbert scheme $\mathrm{Hilb}_{2} \mathbb{C}^{2}$.

Let us first consider the Hilbert scheme $\mathrm{Hilb}_{1} C^{2}$ of one point in $\mathbb{C}^{2}$ which we expect to be $\mathbb{C}^{2}$. Indeed, $H_{1}=\{(v, X, Y) \mid v, X, Y \in \mathbb{C}\}$ and any pair $(X, Y)$ is commuting. Moreover, $v$ is cyclic for $(X, Y)$ if and only if $v \neq 0$. That is, $C_{1}^{c}=\mathbb{C}^{*} \times \mathbb{C} \times \mathbb{C}$. The group $G L_{1}=\mathbb{C}^{*}$ acts via $c .(v, X, Y)=(c v, X, Y)$ and hence the triples $\{(1, X, Y)\}=\mathbb{C}^{2}$ parametrize the orbits of $C_{1}^{c}$, that is, Hilb $_{1} \mathbb{C}^{2}=\mathbb{C}^{2}$ and the ideal $I$ of codimension one corresponding to the point $p=(X, Y) \in \mathbb{C}^{2}$ is the ideal of polynomials $f \in \mathbb{C}[x, y]$ vanishing in $p, f(X, Y)=0$.

Next, we consider the Hilbert scheme Hilb $_{2} \mathbb{C}^{2}$ of two points in $\mathbb{C}^{2}$. Let $(v, X, Y) \in C_{2}^{c}$ and assume that either $X$ or $Y$ has distinct eigenvalues (type a). As

$$
\left.\left[\begin{array}{cc}
\nu_{1} & 0 \\
0 & \nu_{2}
\end{array}\right],\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]\right]=\left[\begin{array}{cc}
0 & \left(\nu_{1}-\nu_{2}\right) b \\
\left(\nu_{2}-\nu_{1}\right) a & 0
\end{array}\right]
$$

we have a representant in the orbit of the form

$$
\left(\left[\begin{array}{l}
v_{1} \\
v_{2}
\end{array}\right],\left[\begin{array}{cc}
\lambda_{1} & 0 \\
0 & \lambda_{2}
\end{array}\right],\left[\begin{array}{cc}
\mu_{1} & 0 \\
0 & \mu_{2}
\end{array}\right]\right)
$$

where cyclicity of the column vector implies that $v_{1} v_{2} \neq 0$. The stabilizer subgroup of the matrix-pair is the group of diagonal matrices $\mathbb{C}^{*} \times \mathbb{C}^{*} \hookrightarrow G L_{2}$, hence the orbit has a unique representant with $v_{1}=v_{2}=1$. The corresponding ideal $I \triangleleft \mathbb{C}[x, y]$ is then

$$
I=\left\{f(x, y) \in \mathbb{C}[x, y] \mid f\left(\lambda_{1}, \mu_{1}\right)=0=f\left(\lambda_{2}, \mu_{2}\right)\right\}
$$

hence these orbits in $C_{2}^{c}$ correspond to sets of two distinct points in $\mathbb{C}^{2}$.
The situation is slightly more complicated when $X$ and $Y$ have only one eigenvalue (type b). If ( $v, X, Y) \in C_{2}^{c}$ then either $X$ or $Y$ is not diagonalizable. But then, as

$$
\left[\left[\begin{array}{ll}
\nu & 1 \\
0 & \nu
\end{array}\right],\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]\right]=\left[\begin{array}{cc}
c & d-a \\
0 & c
\end{array}\right]
$$

we have a representant in the orbit of the form

$$
\left(\left[\begin{array}{l}
v_{1} \\
v_{2}
\end{array}\right],\left[\begin{array}{cc}
\lambda & \alpha \\
0 & \lambda
\end{array}\right],\left[\begin{array}{cc}
\mu & \beta \\
0 & \mu
\end{array}\right]\right)
$$

with $[\alpha: \beta] \in \mathbb{P}^{1}$ and $v_{2} \neq 0$. The stabilizer of the matrixpair is the subgroup

$$
\left\{\left.\left[\begin{array}{ll}
c & d \\
0 & c
\end{array}\right] \right\rvert\, c \neq 0\right\} \longleftrightarrow G L_{2}
$$

and hence we have a unique representant with $v_{1}=0$ and $v_{2}=1$. The corresponding ideal $I \triangleleft \mathbb{C}[x, y]$ is

$$
I=\left\{f(x, y) \in \mathbb{C}[x, y] \mid f(\lambda, \mu)=0 \text { and } \alpha \frac{\partial f}{\partial x}(\lambda, \mu)+\beta \frac{\partial f}{\partial y}(\lambda, \mu)=0\right\}
$$

as one proves by verification on monomials because

$$
\left[\begin{array}{cc}
\lambda & \alpha \\
0 & \lambda
\end{array}\right]^{k}\left[\begin{array}{cc}
\mu & \beta \\
0 & \mu
\end{array}\right]^{l}\left[\begin{array}{l}
0 \\
1
\end{array}\right]=\left[\begin{array}{c}
k \alpha \lambda^{k-1} \mu^{l}+l \beta \lambda^{k} \mu^{l-1} \\
\lambda^{k} \mu^{l}
\end{array}\right]
$$

Therefore, $I$ corresponds to the set of two points at $(\lambda, \mu) \in \mathbb{C}^{2}$ infinitesimally attached to each other in the direction $\alpha \frac{\partial}{\partial x}+\beta \frac{\partial}{\partial y}$. For each point in $\mathbb{C}^{2}$ there is a $\mathbb{P}^{1}$ family of such fat points. Thus, points of $\mathrm{Hilb}_{2} \mathbb{C}^{2}$ correspond to either of the following two situations :

type a

type b

The Hilbert-Chow map $\mathrm{Hilb}_{2} \mathbb{C}^{2} \xrightarrow{\pi} S^{2} \mathbb{C}^{2}$ sends a point of type a to the formal sum $[p]+\left[p^{\prime}\right]$ and a point of type b to $2[p]$. Over the complement of (the image of) the diagonal,
this map is a one-to-one correspondence. However, over points on the diagonal the fibers are $\mathbb{P}^{1}$, so $\pi$ is not a one-to-one correspondence as in the case of $H i l b_{n} \mathbb{C}^{1}$. The situation is nicer for $\mathbb{C}^{1}$ because there points can only collide along one direction, whereas in $\mathbb{C}^{2}$ they can approach each other along a $\mathbb{P}^{1}$ family of lines leading to different ideals. In fact, the symmetric power $S^{2} \mathbb{C}^{2}$ has singularities and the Hilbert-Chow map $H i l b_{2} \mathbb{C}^{2} \xrightarrow{\pi} S^{2} \mathbb{C}^{2}$ is a resolution of singularities.

## 7b. Hilbert stairs.

For the investigation of the $G L_{n}$-action on $H_{n}$ and on the subset $C_{n}^{c}$ we introduce a combinatorial gadget : the Hilbert $n$-stair. This is the lower triangular part of a square $n \times n$ array of boxes

filled with go-stones according to the following two rules :

- each row contains exactly one stone, and
- each column contains at most one stone of each color.

For example, the set of all possible Hilbert 3-stairs is given below.


Let $\mathbb{C}\langle x, y\rangle$ be the free associative algebra on the non-commuting variables $x$ and $y$. That is, $\mathbb{C}\langle x, y\rangle$ is the vectorspace with basis all words in $x$ and $y$ and with multiplication induced by concatenation of words. To every Hilbert stair we will now associate a sequence of words in $x$ and $y$.

At the top of the stairs we place the identity element 1 . Then, we descend the stairs according to the following rule. Every go-stone has a top word $T$ which we may assume we have constructed before and a side word
$S$ and they are related as indicated below


For example, for the Hilbert 3-stairs we have the following sequences of non-commutative words


Let $\sigma$ be a Hilbert $n$-stair with associated sequence of non-commutative words $W(\sigma)=\left\{1, w_{2}(x, y), \ldots, w_{n}(x, y)\right\}$. Let $(v, X, Y) \in H_{n}$ then replacing each occurrence of $x$ in the word $w_{i}(x, y)$ by $X$ and each occurrence of $y$ by $Y$ we obtain an $n \times n$ matrix $w_{i}(X, Y) \in M_{n}(\mathbb{C})$ and by left multiplication a column vector $w_{i}(X, Y) v$. We call the evaluation of $\sigma$ in $(v, X, Y)$ the determinant of the $n \times n$ matrix

$$
\sigma(v, X, Y)=\operatorname{det}\left[\begin{array}{lllll}
v & w_{1}(X, Y) v & w_{2}(X, Y) v & \ldots & w_{n}(X, Y) v
\end{array}\right]
$$

For a fixed Hilbert $n$-stair $\sigma$ we denote with $H_{n}(\sigma)$ the subset of triples $(v, X, Y) \in H_{n}$ with non-zero evaluation $\sigma(v, X, Y) \neq 0$. We claim that none of the $H_{n}(\sigma)$ is empty. Indeed, let $v$ be the basic column vector $e_{1}=\left[\begin{array}{llll}1 & 0 & \ldots & 0\end{array}\right]^{t r}$ and let every black stone in the Hilbert stair $\sigma$ fix a column of $X$ by the rule that if it lies in box $(i, j)$ the $j$-th column of $X$ is the basic column vector $e_{i}=\left[\begin{array}{lllllll}0 & \ldots & 0 & 1 & 0 & \ldots & 0\end{array}\right]^{\text {tr }}$ (a 1 at place $i$ )

and the same rule applies to white stones determining columns of $Y$. That is, one replaces each stone in $\sigma$ by 1 at the same spot in $X$ or $Y$ and fills the remaining spots in the same column by zeroes. We say that such a triple
$(v, X, Y)$ is in $\sigma$-standard form. With these conventions one easily verifies by induction that

$$
w_{i}(X, Y) e_{1}=e_{i} \quad \text { for all } 2 \leq i \leq n .
$$

Hence, filling up the remaining spots in $X$ and $Y$ arbitrarily one has that $\sigma(v, X, Y) \neq 0$ proving the claim. Hence, $H_{n}(\sigma)$ is an open subset of $H_{n}$ for every Hilbert $n$-stair $\sigma$. Further, for every word (monomial) $w(x, y)$ and every $g \in G L_{n}$ we have that

$$
w\left(g X g^{-1}, g Y g^{-1}\right) g v=g w(X, Y) v
$$

and therefore the open sets $H_{n}(\sigma)$ are stable under the $G L_{n}$-action on $H_{n}$. We will give representants of the orbits in $H_{n}(\sigma)$.

Let $W_{n}=\left\{1, x, y, x y, \ldots, y^{n}\right\}$ be the set of all words in the noncommuting variables $x$ and $y$ of length $\leq n$, ordered lexicographically. For every triple $(v, X, Y) \in H_{n}$ consider the $n \times m$ matrix

$$
\psi(v, X, Y)=\left[\begin{array}{llllll}
v & X v & Y v & X Y v & \ldots & Y^{n} v
\end{array}\right]
$$

where $m=2^{n+1}-1$ and the $j$-th column is the column vector $w(X, Y) v$ with $w(x, y)$ the $j$-th word in $W_{n}$. Hence, $(v, X, Y) \in H_{n}(\sigma)$ if and only if the $n \times n$ minor of $\psi(v, X, Y)$ determined by the word-sequence $\left\{1, w_{2}, \ldots, w_{n}\right\}$ of $\sigma$ is invertible. Moreover, as

$$
\psi\left(g v, g X g^{-1}, g Y g^{-1}\right)=g \psi(v, X, Y)
$$

we deduce that the $G L_{n}$-orbit of $(v, X, Y) \in H_{n}(\sigma)$ contains a unique triple ( $v^{\prime}, X^{\prime}, Y^{\prime}$ ) such that the corresponding minor of $\psi\left(v^{\prime}, X^{\prime}, Y^{\prime}\right)=\mathbb{1}_{n}$. Hence, each $G L_{n}$-orbit in $H_{n}(\sigma)$ contains a unique representant in $\sigma$-standard form. Therefore, the orbit space $O_{n}(\sigma)$ of $H_{n}(\sigma)$ is an affine space of dimension the number of non-forced entries in $X$ and $Y$. As we fixed $n-1$ columns in $X$ or $Y$ this dimension is equal to

$$
O_{c}(\sigma)=\mathbb{C}^{k} \quad \text { with } \quad k=2 n^{2}-(n-1) n=n^{2}+n .
$$

For example, representants for the orbits in $H_{3}(\sigma)$ are given by $(v, X, Y)$ with $v=\left[\begin{array}{lll}1 & 0 & 0\end{array}\right]^{\text {tr }}$ and

|  | $\bigcirc$ | $\bullet$ | 0 | $0$ | 0 | $\bigcirc$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $X$ | $\left[\begin{array}{lll}0 & a & b \\ 1 & c & d \\ 0 & e & f\end{array}\right]$ | $\left[\begin{array}{lll}0 & 0 & a \\ 1 & 0 & b \\ 0 & 1 & c\end{array}\right]$ | $\left[\begin{array}{lll}0 & a & b \\ 1 & c & d \\ 0 & e & f\end{array}\right]$ | $\left[\begin{array}{lll}0 & a & b \\ 0 & c & d \\ 1 & e & f\end{array}\right]$ | $\left[\begin{array}{lll}a & b & c \\ d & e & f \\ g & h & i\end{array}\right]$ | $\left[\begin{array}{lll}a & 0 & b \\ c & 0 & d \\ e & 1 & f\end{array}\right]$ |
|  | $\left[\begin{array}{lll}0 & g & h \\ 0 & i & j \\ 1 & k & l\end{array}\right]$ | $\left[\begin{array}{lll}d & e & f \\ g & h & i \\ j & k & l\end{array}\right]$ | $\left[\begin{array}{lll}g & 0 & h \\ i & 0 & j \\ k & 1 & l\end{array}\right]$ | $\left[\begin{array}{lll}0 & g & h \\ 1 & i & j \\ 0 & k & l\end{array}\right]$ | $\left[\begin{array}{lll}0 & 0 & j \\ 1 & 0 & k \\ 0 & 1 & l\end{array}\right]$ | $\left[\begin{array}{lll}0 & g & h \\ 1 & i & j \\ 0 & k & l\end{array}\right]$ |

We define the subset $H_{n}^{c}$ of cyclic triples, that is those $(v, X, Y) \in H_{n}$ such that there exists no proper subspace $W \subset \mathbb{C}^{n}$ containing $v$ and stable under the action of $X$ and $Y$, that is, $X W \subset W$ and $Y W \subset W$. If we denote with $\mathbb{C}\langle X, Y\rangle$ the (not necessarily commutative) $\mathbb{C}$-subalgebra of $M_{n}(\mathbb{C})$ generated by the matrix-pair $(X, Y)$, then $(v, X, Y)$ is a cyclic triple if and only if

$$
\mathbb{C}\langle X, Y\rangle v=\mathbb{C}^{n}
$$

Hence, clearly $H_{n}(\sigma) \subset H_{n}^{c}$ for any Hilbert $n$-stair $\sigma$. Conversely, we claim that a cyclic triple $(v, X, Y) \in H_{n}^{c}$ belongs to at least one of the open subsets $H_{n}(\sigma)$. Indeed, either $X v \notin \mathbb{C} v$ or $Y v \notin \mathbb{C} v$ as otherwise the subspace $W=\mathbb{C} v$ would contradict the cyclicity assumption. Fill the top box of the stairs with the corresponding stone and define the 2-dimensional subspace $V_{2}=\mathbb{C} v_{1}+\mathbb{C} v_{2}$ where $v_{1}=v$ and $v_{2}=w_{2}(X, Y) v$ with $w_{2}$ the corresponding word (either $x$ or $y$ ).

Assume by induction we have been able to fill the first $i$ rows of the stairs with stones leading to the sequence of words $\left\{1, w_{2}(x, y), \ldots, w_{i}(x, y)\right\}$ such that the subspace $V_{i}=\mathbb{C} v_{1}+\ldots+\mathbb{C} v_{i}$ with $v_{i}=w_{i}(X, Y) v$ has dimension $i$. Then, either $X v_{j} \notin V_{i}$ for some $j$ or $Y v_{j} \notin V_{i}$ (if not $V_{i}$ would contradict cyclicity). Then fill the $j$-th box in the $i+1$-th row of the stairs with the corresponding stone. Then, the top $i+1$ rows of the stairs form a Hilbert $i+1$-stair as there can be no stone of the same color lying in the same column. Define $w_{i+1}(x, y)=x w_{i}(x, y)$ (or $y w_{i}(x, y)$ ) and $v_{i+1}=w_{i+1}(X, Y) v$. Then, $V_{i+1}=\mathbb{C} v_{1}+\ldots+\mathbb{C} v_{i+1}$ has dimension $i+1$. Continuing we end up with a Hilbert $n$-stair $\sigma$ such that $(v, X, Y) \in H_{n}(\sigma)$. This concludes the proof of the following result.

Theorem 7.1. The orbit space $O_{n}^{c}$ for the $G L_{n}$-action on the open submanifold $H_{n}^{c}$ of cyclic triples is a manifold of dimension $n^{2}+n$.

## Week 8

## Hilbert manifolds.

Recall that $C_{n}^{c}$ is the subset of the affine space $H_{n}=\mathbb{C}^{n} \oplus M_{n} \oplus M_{n}$ consisting of those triples $(v, X, Y)$ such that $[X, Y]=0$ and $v$ is a cyclic vector for the pair $(X, Y)$. Clearly, $C_{n}^{c}$ is a subset of the open submanifold of cyclic triples $H_{n}^{c}$.

We aim to show that $C_{n}^{c}$ is a manifold of dimension $n^{2}+2 n$. This is slightly surprising as the closely related commuting variety

$$
C=\left\{(X, Y) \in M_{n} \oplus M_{n} \mid[X, Y]=0\right\}
$$

is known to have singularities. In fact, not much is known about this commuting variety except that it is connected.

## 8a. A slice argument.

Consider the vectorspace $Q_{n}=\mathbb{C}^{n} \oplus C^{n *} \oplus M_{n} \oplus M_{n}$ where $\mathbb{C}^{n *}$ denotes the $n$-dimensional vectorspace of row vectors. We define a linear action of $G L_{n}$ on $Q_{n}$ by the rule

$$
g \cdot(v, w, X, Y)=\left(g v, w g^{-1}, g X g^{-1}, g Y g^{-1}\right)
$$

for all $g \in G L_{n}$. For later reference, we depict this action by the pattern


Observe that modding out the row vectors gives a surjection $Q_{n} \xrightarrow{\gamma} H_{n}$ which is $G L_{n}$-equivariant. With $Q_{n}^{c}$ we will denote the open submanifold
$\gamma^{-1}\left(Q_{n}^{c}\right)$ of cyclic quartets. Consider the map

$$
\begin{array}{ccc}
Q_{n} \longrightarrow & \chi & M_{n} \\
(v, w, X, Y) & \mapsto & v . w+[X, Y]
\end{array}
$$

The differential $d \chi$ of this map is computed by expanding $(v+\epsilon a)(w+\epsilon b)$ $+[(X+\epsilon C),(Y+\epsilon D)]$ which shows

$$
d \chi_{(v, w, X, Y)}(a, b, C, D)=v . b+a . w+[X, D]+[C, Y]
$$

We claim that the differential is surjective whenever $(v, w, X, Y)$ is a cyclic quartet. Consider the Hermitian inner product $\langle M, N\rangle=\operatorname{Tr}\left(M \bar{N}^{t r}\right)$ on $M_{n}(\mathbb{C})$, then the space orthogonal to the image of $d \chi_{(v, w, X, Y)}$ is equal to

$$
\left\{M \in M_{n}(\mathbb{C}) \mid \operatorname{Tr}\left(v b \bar{M}^{t r}+a w \bar{M}^{t r}+[X, D] \bar{M}^{t r}+[C, Y] \bar{M}^{t r}\right)=0, \forall(a, b, C, D)\right\}
$$

Because $\operatorname{Tr}$ does not change under cyclic permutation and is non degenerate on $M_{n}(\mathbb{C})$, we see that this orthogonal space is equal to

$$
\left\{M \in M_{n}(\mathbb{C}) \mid \bar{M}^{t r} v=0 \quad w \bar{M}^{t r}=0 \quad\left[\bar{M}^{t r}, X\right]=0 \quad \text { and } \quad\left[Y, \bar{M}^{t r}\right]=0\right\}
$$

If $(v, w, X, Y)$ is a cyclic quartet, for such a matrix $M$ we have that the nullspace $\operatorname{Ker} \bar{M}^{\text {tr }}$ is a proper subspace of $\mathbb{C}^{n}$ containing $v$ and stable under $X$ and $Y$. By the cyclicity condition this implies that $\operatorname{Ker} \bar{M}^{t r}=\mathbb{C}^{n}$ or equivalently that $\bar{M}^{t r}=0$, proving the claim. By the implicit function theorem this implies that the fiber of any point in the image of $\chi$

is a submanifold of $Q_{n}^{c}$ of dimension $\operatorname{dim} Q_{n}-\operatorname{dim} M_{n}=n^{2}+2 n$. In particular, $\chi^{-1}(0)$ is a submanifold of dimension $n^{2}+2 n$. We will now identify the fiber $\chi^{-1}(0)$ with the subset $C_{n}^{c}$ and hence prove :

Proposition 8.1. The subset $C_{n}^{c}$ of cyclic triples $(v, X, Y)$ with $[X, Y]=0$ is a submanifold of dimension $n^{2}+2 n$ of $H_{n}$.

Indeed, let $(v, w, X, Y)$ be a cyclic quartet satisfying $v . w+[X, Y]=0$ and let $m(x, y)$ be any word in the non-commuting variables $x$ and $y$. We claim that $w m(X, Y) v=0$. We prove this by induction on the length $l(m)$ of the word $m$. If $l(m)=0$ then $m=1$ and we have

$$
w m(X, Y) v=w v=\operatorname{Tr}(v \cdot w)=-\operatorname{Tr}([X, Y])=0 .
$$

Assume we proved the claim for all words of length $<l$ and take a word of the form $m(x, y)=m_{1}(x, y) y x m_{2}(x, y)$ with $l\left(m_{1}\right)+l\left(m_{2}\right)+2=l$. Then, we have

$$
\begin{aligned}
w m(X, Y) & =w m_{1}(X, Y) Y X m_{2}(X, Y) \\
& =w m_{1}(X, Y)([Y, X]+X Y) m_{2}(X, Y) \\
& =\left(w m_{1}(X, Y) v\right) \cdot w m_{2}(X, Y)+w m_{1}(X, Y) X Y m_{2}(X, Y) \\
& =w m_{1}(X, Y) X Y m_{2}(X, Y)
\end{aligned}
$$

where we used the induction hypotheses in the last equality (the bracketed term vanishes). Hence we can reorder the terms in $m(x, y)$ if necessary and have that $w m(X, Y)=w X^{l_{1}} Y^{l_{2}}$ with $l_{1}+l_{2}=l$ and $l_{1}$ the number of occurrences of $x$ in $m(x, y)$. Hence, we have to prove the claim for $X^{l_{1}} Y^{l_{2}}$.

$$
\begin{aligned}
w X^{l_{1}} Y^{l_{2}} v & =\operatorname{Tr}\left(X^{l_{1}} Y^{l_{2}} v w\right)=-\operatorname{Tr}\left(X^{l_{1}} Y^{l_{2}}[X, Y]\right) \\
& =-\operatorname{Tr}\left(\left[X^{l_{1}} Y^{l_{2}}, X\right] Y\right)=-\operatorname{Tr}\left(X^{l_{1}}\left[Y^{l_{2}}, X\right] Y\right) \\
& =-\sum_{i=0}^{l_{2}-1} \operatorname{Tr}\left(X^{l_{1}} Y^{i}[Y, X] Y^{l_{2}-i}\right)=-\sum_{i=0}^{l_{2}-1} \operatorname{Tr}\left(Y^{l_{2}-i} X^{l_{1}} Y^{i}[Y, X]\right. \\
& =-\sum_{i=0}^{l_{2}-1} \operatorname{Tr}\left(Y^{l_{2}-i} X^{l_{1}} Y^{i} v \cdot w=-\sum_{i=0}^{l_{2}-1} w Y^{m_{2}-i} X^{l_{1}} Y^{i} v\right.
\end{aligned}
$$

But we have seen that $w Y^{l_{2}-i} X^{l_{1}} Y^{i}=w X^{l_{1}} Y^{l_{2}}$ hence the above implies that $w X^{l_{1}} Y^{l_{2}} v=-l_{2} w X^{l_{1}} Y^{l_{2}} v$. But then $w X^{l_{1}} Y^{l_{2}} v=0$, proving the claim.

Consequently, $w \mathbb{C}\langle X, Y\rangle v=0$ and by the cyclicity condition we have $w \mathbb{C}^{n}=0$ hence $w=0$. Finally, as $v . w+[X, Y]=0$ this implies that $[X, Y]=0$ and we can identify the fiber $\chi^{-1}(0)$ with $C_{n}^{c}$ (identifying $H_{n}$ with the closed submanifold of $Q_{n}$ where $w=0$ ), finishing the proof of proposition 8.1. Recalling that the Hilbert scheme $H i l b_{n} \mathbb{C}^{2}$ is the orbit space of the $G L_{n}$ action on $C_{n}^{c}$ we have the situation


We will construct a slice for the $G L_{n}$-action on $C_{n}^{c}$. First, consider the orbit $\operatorname{map} G L_{n} \xrightarrow{o} H_{n}$ for a point $(v, X, Y) \in H_{n}^{c}$ defined by $g \mapsto g .(v, X, Y)$. The differential of this map in the point $\mathbb{\tau}_{n}$ is given by the linear map

$$
d o_{\mathbb{1}_{n}}: M_{n}(\mathbb{C}) \longrightarrow H_{n} \quad \text { where } \quad d o_{\rrbracket_{n}}(m)=(m v,[m, X],[m, Y])
$$

as one verifies as in the previous section. Observe that this differential is injective whenever $(v, X, Y) \in H_{n}^{c}$. Indeed, if $m$ satisfies $m v=0$ and $[m, X]=0=[m, Y]$ then the nullspace $\operatorname{Ker} m$ is a subspace of $\mathbb{C}^{n}$ containing $v$ and stable under $X$ and $Y$ so must be $\mathbb{C}^{n}$ whence $m=0$. Define an Hermitian inner product on the vectorspace $H_{n}$ by the rule

$$
\left\langle\left(v_{1}, X_{1}, Y_{1}\right),\left(v_{2}, X_{2}, Y_{2}\right)\right\rangle=\operatorname{Tr}\left(v_{1} \bar{v}_{2}^{t r}+X_{1} \bar{X}_{2}^{t r}+Y_{1} \bar{Y}_{2}^{t r}\right) .
$$

The subspace of the tangentspace $H_{n}$ in $(v, X, Y)$ orthogonal to $I m d o_{\rrbracket_{n}}$ is then the subspace of triples $(a, C, D)$ such that

$$
\begin{aligned}
0 & =\operatorname{Tr}\left(m v \bar{a}^{t r}\right)+\operatorname{Tr}\left([m, X] \bar{C}^{t r}\right)+\operatorname{Tr}\left([m, Y] \bar{D}^{t r}\right) \\
& =\operatorname{Tr}\left(m\left(v \bar{a}^{t r}+\left[X, \bar{C}^{t r}\right]+\left[Y, \bar{D}^{t r}\right]\right)\right.
\end{aligned}
$$

Again using the non-degeneracy of the trace map on $n \times n$ matrices, we have

$$
\left(I m d o_{\rrbracket_{n}}\right)^{\perp}=\left\{(a, C, D) \in H_{n} \mid v \bar{a}^{t r}+\left[X, \bar{C}^{t r}\right]+\left[Y, \bar{D}^{t r}\right]=0\right\}
$$

which is of dimension $n^{2}+n$ (using injectivity of $d o_{\chi_{n}}$ ). Reasoning as in the foregoing section we obtain the slice.

Proposition 8.2. Let $(v, X, Y) \in H_{n}^{c}$, then the $n^{2}+n$-dimensional manifold

$$
S=\left\{(v+a, X+C, Y+D) \mid v \bar{a}^{t r}+\left[X, \bar{C}^{t r}\right]+\left[Y, \bar{D}^{t r}\right]=0\right\}
$$

is a slice for the action of $G L_{n}$ on $H_{n}$ in $(v, X, Y)$.
Now let $(v, X, Y) \in C_{n}^{c}$, that is, assume that $[X, Y]=0$. We have seen that $C_{n}^{c}$ is a submanifold of dimension $n^{2}+2 n$. Thus, the tangentspace to $C_{n}^{c}$ in $p=(v, X, Y)$ is of dimension $n^{2}+2 n$. This is the subspace $(a, C, D) \in H_{n}$ such that $[X+\epsilon C, Y+\epsilon D]=0$, that is

$$
T_{p} C_{n}^{c}=\{(a, C, D) \mid[X, D]+[C, Y]=0\}
$$

Observe that as $C_{n}^{c}$ is $G L_{n}$-stable the image of $d o_{\mathbb{1}_{n}}$ is contained in the tangentspace (can be checked immediately using the Jacobi identity and $[X, Y]=0$ ). Therefore, the orthogonal complement of Im $d o_{\mathbb{1}_{n}}$ in the tangentspace is of dimension $n^{2}+2 n-n^{2}=2 n$.

Proposition 8.3. Let $(v, X, Y) \in C_{n}^{c}$, then the $2 n$-dimensional manifold $S^{\prime}$

$$
\left\{(v+a, X+C, Y+D) \mid v \bar{a}^{t r}+\left[X, \bar{C}^{t r}\right]+\left[Y, \bar{D}^{t r}\right]=0 \text { and }[X, D]+[C, Y]=0\right\}
$$

is a slice for the $G L_{n}$-action on $C_{n}^{c}$.
Because the dimension of the slice is independent of the point in $C_{n}^{c}$ this concludes the proof of the following result.

Theorem 8.4. The Hilbert scheme Hilb $\mathbb{C}^{2}$ of $n$ points in $\mathbb{C}^{2}$ is a manifold of dimension $2 n$.

## 8b. The Hilbert game.

The determination of the slice is easier in a triple $(v, X, Y)$ in standard $\sigma$ - form for a Hilbert $n$-stair $\sigma$. In fact the description of $O_{n}(\sigma)$ is a slice for the action of $G L_{n}$ on $H_{n}(\sigma)$. It is an interesting exercise to determine the covering of $H i l b_{n} \mathbb{C}^{2}$ by the $2 n$-dimensional submanifolds $H_{i l b} \mathbb{C}^{2}(\sigma)=$ $H i l b_{n} \mathbb{C}^{2} \cap O_{n}(\sigma)$ of $O_{n}(\sigma)=\mathbb{C}^{n^{2}+n}$. For example, consider Hilb $\mathbb{C}^{2}(\bigcirc)$. Because

$$
\left.\left[\begin{array}{ll}
0 & a \\
1 & b
\end{array}\right],\left[\begin{array}{ll}
c & d \\
e & f
\end{array}\right]\right]=\left[\begin{array}{cc}
a e-d & a f-a c-b d \\
c+b e-f & d-a e
\end{array}\right]
$$

this subset can be identified with $\mathbb{C}^{4}$ using the equalities $d=a r$ and $f=$ $c+b e$. Similarly, one has identifications


However, for $\mathrm{Hilb}_{3} \mathbb{C}^{2}(\square \bigcirc)$ the description is more complicated. Observe that some of these intersections may be empty. For example, consider the Hilbert 5-stair

then the associated series of words is $\{1, x, y, x y, y x\}$ whence $\sigma(v, X, Y)=0$ whenever $[X, Y]=0$. Hence all Hilbert stairs $\sigma$ containing $\sigma_{5}$ (that is, if we recover $\sigma_{5}$ after removing certain rows and columns) satisfy $H i l b_{n} \mathbb{C}^{2}(\sigma)=$ $\emptyset$.

Call a Hilbert $n$-stair $\sigma$ a forbidden position if $H i l b_{n} \mathbb{C}^{2}(\sigma)$ is empty. A forbidden position of minimal size is called a blocking position. Examples of blocking positions are $\sigma_{5}$ above and the Hilbert 6 -stairs

and


Determine all blocking positions for small $n$ (up to color changes). Consider the following two person game on an $n$-stair. Left and right take turns where left places bLack stones and right white stones according to
the Hilbert $n$-stair rules. The person unable to move or forced to move to a forbidden position is declared the looser. Determine the values of positions for small $n$ following the rules of combinatorial game theory as explained in J.H. Conway's "On Numbers and Games". For example,


## 8c. Connectedness.

We have shown that $H i l b_{n} \mathbb{C}^{2}$ is a manifold of dimension $2 n$. A priori it may have many connected components (all of dimension $2 n$ ). We will now show that $H i l b_{n} \mathbb{C}^{2}$ is connected. As there is clearly a component of $H i l b_{n} \mathbb{C}^{2}$ of which points in general position correspond to $n$ distinct points in $\mathbb{C}^{2}$, this result implies that any fat $n$-point of $\mathbb{C}^{2}$ can be deformed into $n$ distinct simple points. In the next section we will show that a similar result does not hold for $H i l b_{n} \mathbb{C}^{m}$ with $m, n$ sufficiently large.

Recall that the symmetric power $S^{n} \mathbb{C}^{1}$ parametrizes sets of $n$-points on the line $\mathbb{C}^{1}$ and can be identified with $\mathbb{C}^{n}$. Consider the map

$$
\operatorname{Hilb}_{n} \mathbb{C}^{2} \xrightarrow{\pi} S^{n} \mathbb{C}^{1}
$$

defined by mapping a cyclic triple $(v, X, Y) \in C_{n}^{c}$ with $[X, Y]=0$ in the orbit corresponding to the point of $\operatorname{Hilb}_{n} \mathbb{C}^{2}$ to the set $\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}$ of eigenvalues of $X$. Observe that this map does not depend on the point chosen in the orbit. Let $\Delta$ be the big diagonal in $S^{n} \mathbb{C}^{1}$, that is, $S^{n} \mathbb{C}^{1}-\Delta$ is the space of all sets of $n$ distinct points from $\mathbb{C}^{1}$. Clearly, $S^{n} \mathbb{C}^{1}-\Delta$ is a connected $n$-dimensional manifold. We claim that

$$
\pi^{-1}\left(S^{n} \mathbb{C}^{1}-\Delta\right) \simeq\left(S^{n} \mathbb{C}^{1}-\Delta\right) \times \mathbb{C}^{n}
$$

and hence is connected. Indeed, take a matrix $X$ with $n$ distinct eigenvalues $\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}$. We may diagonalize $X$. But then, as

$$
\left.\left[\begin{array}{ccc}
\lambda_{1} & & \\
& \ddots & \\
& & \lambda_{n}
\end{array}\right],\left[\begin{array}{ccc}
y_{11} & \ldots & y_{1 n} \\
\vdots & & \vdots \\
y_{n 1} & \ldots & y_{n n}
\end{array}\right]\right]=\left[\begin{array}{ccc}
\left(\lambda_{1}-\lambda_{1}\right) y_{11} & \ldots & \left(\lambda_{1}-\lambda_{n}\right) y_{1 n} \\
\vdots & & \vdots \\
\left(\lambda_{n}-\lambda_{1}\right) y_{n 1} & \ldots & \left(\lambda_{n}-\lambda_{n}\right) y_{n n}
\end{array}\right]
$$

we see that also $Y$ must be a diagonal matrix with entries $\left(\mu_{1}, \ldots, \mu_{n}\right) \in \mathbb{C}^{n}$ where $\mu_{i}=y_{i i}$. But then the cyclicity condition implies that all coordinates
of $v$ must be non-zero. Now, the stabilizer subgroup of the commuting (diagonal) matrix-pair $(X, Y)$ is the maximal torus $T_{n}=\mathbb{C}^{*} \times \ldots \times \mathbb{C}^{*}$ of diagonal invertible $n \times n$ matrices. Using its action we may assume that all coordinates of $v$ are equal to 1 . That is, the points in $\pi^{-1}\left(\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}\right)$ with $\lambda_{i} \neq \lambda_{j}$ have unique representants of the form

$$
\left(\left[\begin{array}{c}
1 \\
1 \\
\vdots \\
1
\end{array}\right],\left[\begin{array}{llll}
\lambda_{1} & & & \\
& \lambda_{2} & & \\
& & \ddots & \\
& & & \lambda_{n}
\end{array}\right],\left[\begin{array}{llll}
\mu_{1} & & & \\
& \mu_{2} & & \\
& & \ddots & \\
& & & \mu_{n}
\end{array}\right]\right)
$$

that is $\pi^{-1}\left(\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}\right.$ can be identified with $\mathbb{C}^{n}$, proving the claim. Next, we claim that all the fibers of $\pi$ have dimension at most $n$.

Let $\left\{\lambda_{1}, \ldots, \lambda_{n}\right\} \in S^{n} \mathbb{C}^{1}$ then there are only finitely many $X$ in Jordan normalform with eigenvalues $\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}$. Fix such an $X$, then the subset $T(X)$ of cyclic triples $(X, Y . v) \in C_{n}^{c}$ has dimension at most $n+\operatorname{dim} C(X)$ where $C(X)$ is the centralizer of $X$ in $M_{n}(\mathbb{C})$, taht is, $C(X)=\left\{Y \in M_{n}(\mathbb{C}) \mid\right.$ $X Y=Y X\}$. The stabilizer subgroup $\operatorname{Stab}(X)=\left\{g \in G L_{n} \mid g X g^{-1}=X\right\}$ is an open subset of the vectorspace $C(X)$ and acts freely on the subset $T(X)$ because the action of $G L_{n}$ on $C_{n}^{c}$ has trivial stabilizers. But then, the orbitspace for the $\operatorname{Stab}(X)$-action on $T(X)$ has dimension at most $n+$ $\operatorname{dim} C(X)-\operatorname{dim} \operatorname{Stab}(X)=n$. As we only have to consider finitely many $X$ this proves the claim.

The diagonal $\Delta$ has dimension $n-1$ in $S^{n} \mathbb{C}^{1}$ and hence by the foregoing we know that the dimension of $\pi^{-1}(\Delta)$ is at most $2 n-1$. Let $H$ be the connected component of $H i l b_{n} \mathbb{C}^{2}$ containing teh connected subset $\pi^{-1}\left(S^{n} \mathbb{C}^{1}-\Delta\right)$. If $\pi^{-1}(\Delta)$ were not entirely contained in $H$, then $H_{i l b} \mathbb{C}^{2}$ would have a component of dimension less than $2 n$, which we proved not to be the case. Thus $H=H i l b_{n} \mathbb{C}^{2}$ and we have proved :

Theorem 8.5. The Hilbert scheme Hilb $\mathbb{C}^{2}$ of n points in $\mathbb{C}^{2}$ is a connected manifold of dimension $2 n$.

Let $(v, X, Y)$ be a cyclic triple representing a point in $h \in H i l b_{n} \mathbb{C}^{2}$, that is $[X, Y]=0$. We claim that $X$ and $Y$ are simultaneously upper triangularizable. Let $\lambda$ be an eigenvalue of $X$ and consider the eigenspace $V_{\lambda}=\{w \in$ $\left.\mathbb{C}^{n} \mid X w=\lambda w\right\}$. Then, $Y V_{\lambda} \subset V_{\lambda}$ as $X Y w=Y(X w)=Y(\lambda w)=\lambda Y w$. Let $Y_{\lambda}$ be the matrix representing the action of $Y$ on $V_{\lambda}$ then up to basechange in $V_{\lambda}$ we may assume that $Y_{\lambda}$ is in Jordan normal form, but then $X$ and $Y$ have at least one common eigenvector $w \in V_{\lambda}$ such that $Y w=\mu w$. Consider a new basis $\left\{f_{1}, \ldots, f_{n}\right\}$ with $f_{n}=w$, then in this basis the matrixpair $(X, Y)$
has blockform


But then $\left(X_{1}, Y_{1}\right)$ is a commuting matrixpair in $M_{n-1}(\mathbb{C})$ and by induction we may assume that they are simultaneously upper triangularizable, proving the claim. Hence, we have a cyclic triple in the orbit of $(v, X, Y)$ of the form

$$
\left(\left[\begin{array}{c}
w_{1} \\
\vdots \\
w_{n}
\end{array}\right],\left[\begin{array}{ccc}
\lambda_{1} & \ldots & * \\
& \ddots & \vdots \\
& & \lambda_{n}
\end{array}\right],\left[\begin{array}{ccc}
\mu_{1} & \ldots & * \\
& \ddots & \vdots \\
& & \mu_{n}
\end{array}\right]\right)
$$

The Hilbert-Chow map Hilb $\mathbb{C}^{2} \xrightarrow{H} S^{n} \mathbb{C}^{2}$ is defined by sending the point $h$ representing the orbit of $(v, X, Y)$ to the set $\left\{\left(\lambda_{1}, \mu_{1}\right), \ldots,\left(\lambda_{n}, \nu_{n}\right)\right\}$ of $n$ points in $\mathbb{C}^{2}$. If $\left(\lambda_{i}, \mu_{i}\right) \neq\left(\lambda_{j}, \mu_{j}\right)$ for $i<j$ and consider the $2 \times 2$ minors of $X$ and $Y$

$$
X_{i j}=\left[\begin{array}{cc}
\lambda_{i} & x_{i j} \\
0 & \lambda_{j}
\end{array}\right] \quad Y_{i j}=\left[\begin{array}{cc}
\mu_{i} & y_{i j} \\
0 & \mu_{j}
\end{array}\right]
$$

Assume that $\lambda_{i} \neq \lambda_{j}$, conjugating with the matrix $g=\left[\begin{array}{cc}1 & \frac{x_{i j}}{\lambda_{i}-\lambda_{j}} \\ 0 & 1\end{array}\right]$ gives

$$
\left[\begin{array}{cc}
\lambda_{i} & 0 \\
0 & \lambda_{j}
\end{array}\right] \quad\left[\begin{array}{cc}
\mu_{i} & y_{i j}^{\prime} \\
0 & \mu_{j}
\end{array}\right]
$$

but then the commutation relation forces $y_{i j}^{\prime}=0$. Repeating this argument and possibly permuting the base vectors we may assume that the commuting matrix pair $(X, Y)$ can be brought into block form

and


Here, $X_{i}\left(\operatorname{resp} . Y_{i}\right)$ is an upper triangular $m_{i} \times m_{i}$ matrix with single eigenvalue $\lambda_{i}\left(\right.$ resp. $\left.\mu_{i}\right)$ where $p_{i}$ is the multiplicity with which $p_{i}=\left(\lambda_{i}, \mu_{i}\right)$ appears in the $n$-set of points. The image under the Hilbert-Chow map is then the formal sum

$$
H(v, X, Y)=m_{1}\left[p_{1}\right]+\ldots m_{k}\left[p_{k}\right] \in S^{n} \mathbb{C}^{2}
$$

Another way to phrase the above block decomposition is in terms of the $n$-dimensional algebra $\mathbb{C}[X, Y]=\frac{\mathbb{C}[x, y]}{I}$ determined by $h \in H i l b_{n} \mathbb{C}^{2}$. As $\mathbb{C}[X, Y]$ is the $\mathbb{C}$-subalgebra of $M_{n}(\mathbb{C})$ generated by the commuting matrixpair $(X, Y)$ we have a decomposition

$$
\mathbb{C}[X, Y]=\mathbb{C}\left[X_{1}, Y_{1}\right] \oplus \ldots \oplus \mathbb{C}\left[X_{k}, Y-k\right]=\frac{\mathbb{C}[x, y]}{I_{1}} \oplus \ldots \oplus \frac{\mathbb{C}[x, y]}{I_{k}}
$$

where $I_{j}$ is an ideal of $\mathbb{C}[x, y]$ of codimension $m_{j}$ concentrated in the point $p_{j}=\left(\lambda_{j}, \mu_{j}\right)$. This means that $p_{j}$ is the only point of $\mathbb{C}^{2}$ where every polynomial $f \in I_{j}$ vanishes. Let us draw some consequences from this decomposition. If all multiplicities $m_{j}$ are one, that is if $H(v, X, Y)$ does not lie on the diagonal $\Delta$ in $S^{n} \mathbb{C}^{2}$, then $X$ and $Y$ are simultaneously diagonalizable. But then, as the stabilizer subgroup of the commuting matrix-pair $(X, Y)$ is equal to the maximal torus $T_{n}=\mathbb{C}^{*} \times \ldots \times \mathbb{C}^{*}$ of diagonal elements in $G L_{n}$ we see that there is precisely one orbit in $C_{n}^{c}$ lying over this point represented by the cyclic triple

$$
\left(\left[\begin{array}{c}
1 \\
\vdots \\
1
\end{array}\right],\left[\begin{array}{lll}
\lambda_{1} & & \\
& \ddots & \\
& & \lambda_{n}
\end{array}\right],\left[\begin{array}{lll}
\mu_{1} & & \\
& \ddots & \\
& & \mu_{n}
\end{array}\right]\right)
$$

Hence, over the open subset $S^{n} \mathbb{C}^{2}-\Delta$ the Hilbert-Chow map $H$ is a one-to-one correspondence. As we have seen that $H i l b_{n} \mathbb{C}^{2}$ is a manifold, the Hilbert-Chow map is a resolution of singularities of $S^{n} \mathbb{C}^{2}$.

When $\delta=m_{1}\left[p_{1}\right]+\ldots+m_{k}\left[p_{k}\right] \in \Delta$, then any point in the fiber of the Hilbert-Chow map is determined by a $k$-tuple $\left(h_{1}, \ldots, h_{k}\right)$ where $h_{i} \in$ Hilb $_{m_{i}} \mathbb{C}^{2}$ and concentrated in $p_{i}$, that is, the image of $h_{i}$ under the Hilbert-Chow map $\mathrm{Hilb}_{m_{i}} \mathbb{C}^{2} \longrightarrow S^{m_{i}} \mathbb{C}^{2}$ is $m_{i}\left[p_{i}\right]$. The parallel translations in $\mathbb{C}^{2}$ give a natural one-to-one correspondence between $H^{-1}\left(m_{i}[(0,0)]\right)$ and $H^{-1}\left(m_{i}\left[p_{i}\right]\right)$. Hence, it is important to study the subset $H^{-1}(n[(0,0)])$ of $H_{i l b} \mathbb{C}^{2}$. Its points correspond to codimension $n$ ideals $I$ of $\mathbb{C}[x, y]$ concentrated in $(0,0)$. The $n$-dimensional algebra $\mathbb{C}[x, y] / I$ has a unique underlying point $(0,0)$, that is, it is a local algebra of dimension $n$. Hence, the algebras $\mathbb{C}[x, y] / I$ are examples of fat points of multiplicity $n$, that is, commutative local algebras of dimension $n$.

## Week 9

## Reductivity of $G L_{n}$.

In this section we will put the examples considered before in a general framework. To begin, in each of the examples we have a linear action of the basechange group $G L_{n}$ on a finite dimensional $\mathbb{C}$-vectorspace $V$, that is $g(v+w)=g \cdot v+g, w, \mathbb{1}_{n} \cdot v=v$ and $\left(g g^{\prime}\right) \cdot v=g \cdot\left(g^{\prime} \cdot v\right)$ for all $g, g^{\prime} \in G L_{n}$ and $v, w \in V$

| Jordan forms | $\left.\begin{array}{l}V=M_{n}(\mathbb{C}) \\ \\ \text { Dynamical systems } \\ \text { D.m }\end{array}\right)$ |
| :--- | :--- |
|  | $V=M_{n \times m}(\mathbb{C}) \oplus M_{n}(\mathbb{C}) \oplus M_{p \times n}(\mathbb{C})$ |
|  | $g \cdot(A, B, C)=\left(g A, g B g^{-1}, C g^{-1}\right)$ |
| Hilbert schemes | $V=\mathbb{C}^{n} \oplus M_{n}(\mathbb{C}) \oplus M_{n}(\mathbb{C})$ |
|  | $g \cdot(v, X, Y)=\left(g v, g X g^{-1}, g Y g^{-1}\right)$ |

Our first objective will be to control all linear actions possible. We will call actions on $V$ and $W$ isomorphic if there is a linear isomorphism $V \xrightarrow{\phi} W$ which is $G L_{n}$-equivariant, that is $g . \phi(v)=\phi(g . v)$ for all $g \in G L_{n}$ and $v \in V$.

## Problem 7.

Describe all linear actions of $G L_{n}$ on finite dimensional vectorspaces $V$ up to isomorphism.

Let $G$ be a group and $V$ a finite dimensional $\mathbb{C}$-vectorspace on which $G$ acts linearly. We say that $V$ is a $G$-representation. If $V$ and $W$ are $G$ representations, then so are $V \oplus W$ and $V \otimes W$ where the actions are given by

$$
g \cdot(v, w)=(g \cdot v, g \cdot w) \quad \text { and } \quad g \cdot(v \otimes w)=g \cdot v \otimes g \cdot w .
$$

for all $g \in G$ and all $v \in V, w \in W$. A $G$-subrepresentation of $V$ is a subspace $W$ which is left stable under the action of $G$. A $G$-representation $V$ is said to be simple if $V$ contains no proper $G$-subrepresentations and is said to be completely reducible if $V$ is the direct sum of simple $G$-subrepresentations.

## 9a. Haar measures.

Our first aim is to prove that $G L_{n}$ is a reductive group, that is, all $G L_{n}$ representations are completely reducible. The method of proof is based on the 'averaging over the group'-idea used to prove that finite groups are reductive. We will briefly sketch it. Let $G$ be a finite group and $W$ a $G$ subrepresentation of a $G$-representation $V$. Let $\phi: V \longrightarrow W$ be a linear projection obtained from extending a basis of $W$ to one of $V$. Then, we consider the averaged linear map

$$
\pi: V \longrightarrow W \quad \text { where } \quad v \mapsto \sum_{g \in G} g\left(\phi\left(g^{-1} \cdot v\right)\right)
$$

This map is $G$-equivariant and restricted to $W$ it is multiplication by $\#(G)$. The kernel $K$ of $\pi$ is a subspace of $V$, stable under the action of $G$ and complementary to $W$. That is, $V=W \oplus K$ is a decomposition of $G$ representations. Continuing gives a complete decomposition of $V$.

We will replace the sum by an integral and the finite group by the compact subgroup of unitary matrices

$$
U_{n}=\left\{A \in G L_{n} \mid A \cdot \bar{A}^{t r}=I_{n}\right\}
$$

Clearly, $U_{n}$ is a subgroup of $G L_{n}$ and we claim that it is a compact Lie group, that is a real compact differentiable manifold (a $C^{\infty}$-manifold) with a differentiable groupstructure. Because $U_{n}$ is a group it suffices to verify the manifold property in a neighborhood of the unit element $e=\mathbb{1}_{n}$. Let $H e r m_{n}$ be the $\mathbb{R}$-vectorspace of Hermitian $n \times n$ matrices $\operatorname{Herm}_{n}=\{H \in$ $\left.M_{n}(\mathbb{C}) \mid H=\bar{H}^{t r}\right\}$ and consider the map

$$
f: G L_{n} \longrightarrow \operatorname{Herm}_{n} \quad \text { defined by } \quad A \longrightarrow A \bar{A}^{t r}
$$

Calculating the differential with the $\epsilon$-method gives that $\operatorname{Im} d f_{\mathbb{T}_{n}}(X)=$ $X+\bar{X}^{t r}$ and as any Hermitian matrix $H$ can be written as $\frac{1}{2}\left(H+\bar{H}^{t r}\right)$ this differential is surjective. By the implicit function theorem (over $\mathbb{R}$ ) we deduce that the fiber $f^{-1}\left(\mathbb{D}_{n}\right)=U_{n}$ is a real manifold of $\mathbb{R}$-dimension $2 n^{2}-\operatorname{dim}$ Herm $_{n}=n^{2}$. Finally, $U_{n}$ is compact as it is closed in the $\mathbb{R}$ topology on $M_{n}$ and bounded as the norms of all entries in a fixed row (or column) add up to one.

Compact Lie groups have a Haar measure allowing to integrate complex valued continuous functions in an invariant way. For example $U_{1}=\{c \in C \mid$ $c \bar{c}=1\}$ is the unit circle $S^{1}=\left\{e^{i x} \mid 0 \leq x<2 \pi\right\}$. We define a complex valued linear map $\int_{U_{1}} \cdot(g) d g$ on the space of all continuous functions $f: U_{1} \longrightarrow \mathbb{C}$ by

$$
\int_{U_{1}} f(g) d g=\frac{1}{2 \pi} \int_{0}^{2 \pi} f\left(e^{i x}\right) d x
$$

We see that this map is normalized, that is $\int_{U_{1}} d g=1$ and is left and right invariant, that is for any $h=e^{i y} \in U_{1}$

$$
\int_{U_{1}} f(g h) d g=\frac{1}{2 \pi} \int_{0}^{2 \pi} f\left(e^{x+y}\right) d x=\frac{1}{2 \pi} \int_{0}^{2 \pi} f\left(e^{x^{\prime}}\right) d x^{\prime}=\int_{U_{1}} f(g) d g
$$

and similarly for multiplication by $h$ on the left. For the compact group $G=U_{1} \times U_{1}$, that is the torus group $S^{1} \times S^{1}$

we can take as the normalized invariant integral $\int_{U_{1} \times U_{1}} f(g) d g=$ $\frac{1}{4 \pi^{2}} \int_{0}^{2 \pi} \int_{0}^{2 \pi} f\left(e^{i x}, e^{i y}\right) d x d y$. In general, a Haar measure on a compact Lie group is a linear functional $\int_{G} f(g) d g$ which is normalized $\int_{G} d g=1$ and is left and right invariant $\int_{G} f(g h) d g=\int_{G} f(g) d g=\int_{G} f(h g) d g$. Assuming its existence we can prove :

Proposition 9.1. If G is a compact Lie group, then every $G$-representation is completely reducible.

Let $W$ be a subspace of a finite dimensional $G$-representation $V$ which is invariant under the $G$-action. Extending a basis of $W$ to one of $V$ gives a linear projection $V \xrightarrow{\pi} W$ which is the identity on $W$. For $v \in V$ we have complex valued coordinate maps

$$
G \longrightarrow W=\mathbb{C}^{k} \quad \text { defined by } \quad g \mapsto g \cdot\left(\pi\left(g^{-1} \cdot v\right)\right.
$$

Integrating these coordinate maps defines a map $\phi: V \longrightarrow W$

$$
\phi(v)=\int_{G} g \cdot \pi\left(g^{-1} \cdot v\right) d g
$$

which is linear and the identity on $W$. Moreover, $\phi$ commutes with the $G$-action as for every $h \in G$

$$
\begin{array}{cc}
\phi(h \cdot v) & =\int_{G} g \cdot \pi\left(g^{-1} \cdot h \cdot v\right) d g=h \cdot \int_{G} h^{-1} g \cdot \pi\left(g^{-1} h \cdot v\right) d g \\
\stackrel{*}{=} & h \cdot \int_{G} g \cdot \pi\left(g^{-1} \cdot v\right) d g=h \cdot \phi(v)
\end{array}
$$

where the starred equality uses invariance of the Haar measure on $G$. But then, $V=W \oplus \operatorname{Ker} \phi$ is a decomposition as $G$-representations. Continuing whenever one of the components has a non-trivial $G$-subrepresentation, we arrive after a finite number of steps at a decomposition of $V$ into simple $G$-representations proving proposition 9.1.

## 9b. Cartan decomposition.

Next, we want to move from the (real) compact Lie group to the associated (complex) algebraic group. In particular, from the (real) torus $T_{n}(\mathbb{R})=U_{1} \times$ $\ldots \times U_{1}$ to the (complex) torus $T_{n}=\mathbb{C}^{*} \times \ldots \times \mathbb{C}^{*}$ and from the unitary group $U_{n}$ to the basechange group $G L_{n}$. Let us first consider the torus case. A polynomial $f(t) \in \mathbb{C}[t]$ has only finitely many zeroes so if $f\left(U_{1}\right)=0, f$ must be the zero polynomial whence $f\left(\mathbb{C}^{*}\right)=0$. A similar result holds for $n$-dimensional tori. Let $f\left(x_{1}, \ldots, x_{n}\right)$ be a polynomial in $\mathbb{C}\left[x_{1}, \ldots, x_{n}\right]$ and write it as a polynomial in $x_{n}$ with coefficients in $\mathbb{C}\left[x_{1}, \ldots, x_{n-1}\right]$
$f\left(x_{1}, \ldots, x_{n}\right)=f_{0}\left(x_{1}, \ldots, x_{n-1}\right) x_{n}^{k}+\ldots+f_{k}\left(x_{1}, \ldots, x_{n-1}\right) x_{n}+f_{k+1}\left(x_{1}, \ldots, x_{n-1}\right)$
Assume that $f\left(T_{n}(\mathbb{R})\right)=0$ and fix $\left(u_{1}, \ldots, u_{n-1}\right) \in T_{n-1}(\mathbb{R})$, then $f\left(u_{1}, \ldots, u_{n-1}, x_{n}\right)$ is a polynomial vanishing on $U_{1}$, whence all the coefficients $f_{i}\left(u_{1}, \ldots, u_{n-1}\right)$ must be zero. Hence, the $f_{i}$ are polynomials such that $f_{i}\left(T_{n-1}(\mathbb{R})\right)=0$ and by induction on $n$ we may assume that then $f_{i}\left(T_{n-1}\right)=0$, whence $f\left(T_{n}\right)=0$.

Assume now that $V$ is a $T_{n}$-representation. Assume that $V$ has a decomposition $V=W \oplus W^{\prime}$ as a $T_{n}(\mathbb{R})$-representation. Consider the normalizer subgroup

$$
N=N_{T_{n}}(W)=\left\{c=\left(c_{1}, \ldots, c_{n}\right) \in T_{n} \mid c . W \subset W\right\}
$$

Extending a basis $\left\{w_{1}, \ldots, w_{l}\right\}$ of $W$ to $V$ we see that this condition can be expressed by the fact that certain certain coordinates of $c . w_{i}$ (which are polynomial functions in the $c_{j}$ ) must be zero. Hence, $N$ can be identified as the subset of points of $T_{n}$ which are simultaneous zeroes of a set of polynomials $\left\{f_{a}\right\}$. Because $W$ is a $T_{n}(\mathbb{R})$-representation we have $T_{n}(\mathbb{R}) \subset N$ and hence $f_{a}\left(T_{n}(\mathbb{R})\right)=0$ for all $f_{a}$, whence $f_{a}\left(T_{n}\right)=0$. That is, $N=T_{n}$ and so $W$ is also a $T_{n}$-representation. That is, a decomposition of $V$ in simple
$T_{n}(\mathbb{R})$-representations is also a decomposition of simple $T_{n}$-representations. Hence, the complex torus $T_{n}=\mathbb{C}^{*} \times \ldots \times \mathbb{C}^{*}$ is a reductive group.

To prove a similar result for $G L_{n}$ we need a polynomial density result for $U_{n}$ with respect to $G L_{n}$. This follows from the Cartan decomposition asserting that

$$
G L_{n}=U_{n} T_{n} U_{n}
$$

where $T_{n}$ is the maximal torus of diagonal matrices in $G L_{n}$. Indeed, for $g \in G L_{n}, g \bar{g}^{t r}$ is an Hermitian matrix and hence diagonalizable by unitary matrices (this follows from the Gramm-Schmidt procedure of orthonormal bases in linear algebra). So, there is a $u \in U_{n}$

$$
u^{-1} g \bar{g}^{t r} u=\left[\begin{array}{ccc}
\alpha_{1} & & \\
& \ddots & \\
& & \alpha_{n}
\end{array}\right]=\underbrace{u^{-1} g u}_{p} \cdot \underbrace{u^{-1} \bar{g}^{t r} u}_{\bar{p}^{t r}}
$$

Because $\alpha_{i}=\sum_{j=1}^{n}\left\|p_{i j}\right\|^{2}, \alpha_{i}>0 \in \mathbb{R}$. Let $\beta_{i}=\sqrt{\alpha_{i}}$ and consider the diagonal matrix

$$
d=\left[\begin{array}{lll}
\beta_{1} & & \\
& \ddots & \\
& & \beta_{n}
\end{array}\right]
$$

Clearly, $g=u d\left(d^{-1} u^{-1} g\right)$ and $v=d^{-1} u^{-1} g$ is a unitary matrix as

$$
\begin{aligned}
v \bar{v}^{t r} & =\left(d^{-1} u^{-1} g\right) \cdot\left(\bar{g}^{t r} u d^{-1}\right)=d^{-1}\left(u^{-1} g \bar{g}^{t r} u\right) d^{-1} \\
& =d^{-1} d^{2} d^{-1}=\mathbb{1}_{n}
\end{aligned}
$$

proving the Cartan decomposition. Let $f=f\left(x_{11}, x_{12}, \ldots, x_{n n}\right)$ be a polynomial function in the matrix entries of $M_{n}(\mathbb{C})$ such that $f\left(U_{n}\right)=0$. Then $f^{\prime}=f \mid T_{n}$ is a polynomial such that $f^{\prime}\left(T_{n}(\mathbb{R})\right)=0$ whence $f\left(T_{n}\right)=0$. By continuity of the matrix-multiplication in $G L_{n}$ and the Cartan decomposition, it follows that $f\left(G L_{n}\right)=0$. Using this polynomial denseness of $U_{n}$ in $G L_{n}$ we can repeat the above $T_{n}$ argument verbatim and conclude :

Theorem 9.2. The n-dimensional torus $T_{n}$ and the basechange group $G L_{n}$ are reductive groups. That is, every representation $V$ admits a decomposition as a direct sum of simple subrepresentations $S_{i}$

$$
V=S_{1} \oplus \ldots \oplus S_{k}
$$

Moreover, such a decomposition is unique up to permutation of the factors.
The last statement follows because any $G$-equivariant map $S \xrightarrow{\phi} S^{\prime}$ between two simple $G$-representations is either the zero map or an isomorphism (the kernel Ker $\phi$ is a $G$-subrepresentation of $S$ ).

## Week 10

## $G L_{n}$-representations.

Reductivity of $G L_{n}$ reduces our problem to that of classifying all simple $G L_{n}$-representations $V$. Again, we will first consider this problem for the $n$ dimensional complex torus $T_{n}$. We claim that the simple $T_{n}$-representations are all one dimensional and are classified by the lattice $\mathbb{Z}^{n}$. For example, for $T_{2}$ we get the discrete set


## 10a. Characters.

Let $V$ be a simple $T_{n}$-representation of dimension $m$ and $h=\left(r_{1}, \ldots, r_{n}\right) \in$ $T_{n}(\mathbb{R})$ with all $r_{i}$ roots of unity. The subgroup $<h>$ of $T_{n}(\mathbb{R})$ is a finite Abelian group so as an $<h>$ - representation $V=\oplus V_{\lambda_{i}}$ where $V_{\lambda_{i}}$ is the eigenspace $\left\{v \in V \mid h . v=\lambda_{i} v\right\}$. $T_{n}$ being Abelian, each of the $V_{\lambda_{i}}$ is a $T_{n}{ }^{-}$ subrepresentation. By simplicity, there is only one non-zero eigenspace, that is $V=V_{\lambda}$. Varying $h$ we see that the subgroup $\mu \times \ldots \times \mu$ of $T_{n}(\mathbb{R})$ consisting of elements having all its entries roots of unities acts diagonally on $V$. But then so does its closure, which is $T_{n}(\mathbb{R})$ so $V$ is a direct sum of one-dimensional $T_{n}(\mathbb{R})$-subrepresentations, a contradiction unless $m=1$. Hence, all simple $T_{n}$-representations are one dimensional. For $V$ a one-dimensional simple $T_{n}$-representation, the action determines (and is determined by) a groupmorphism $T_{n} \xrightarrow{\chi} \mathbb{C}^{*}$. The only groupmorphisms $\mathbb{C}^{*} \longrightarrow \mathbb{C}^{*}$ are easily seen to be the maps $x \mapsto x^{\nu}$ for some $\nu \in \mathbb{Z}$. Hence $V$
is determined by the character

$$
\chi: T_{n} \longrightarrow \mathbb{C}^{*} \quad \text { given by } \quad\left(x_{1}, \ldots, x_{n}\right) \mapsto x_{1}^{a_{1}} \ldots x_{n}^{a_{n}}
$$

for some $a=\left(a_{1}, \ldots, a_{n}\right) \in \mathbb{Z}^{n}$, proving the claim. Traditionally, one writes the character group $X\left(T_{n}\right)=\mathbb{Z}^{n}$ additively and denotes the character corresponding to the standard basis vector $e_{i}$ by $\epsilon_{i}: T_{n} \longrightarrow \mathbb{C}^{*}$ defined by $\left(x_{1}, \ldots, x_{n}\right) \mapsto x_{i}$.

Proposition 10.1. Every finite dimensional $T_{n}$-representation $V$ is the direct sum of its eigenspaces

$$
V_{\chi}=\left\{v \in V \mid x . v=\chi(x) v \text { for all } x=\left(x_{1}, \ldots, x_{n}\right) \in T_{n}\right\}
$$

where $\chi=\sum_{i} a_{i} \epsilon_{i}$ is a character in $X\left(T_{n}\right)=\mathbb{Z}^{n}$.
On $X\left(T_{n}\right)=\mathbb{Z}^{n}$ we put an ordering defined as follows. Let $\lambda=\sum a_{i} \epsilon_{i}$ and $\mu=\sum b_{i} \epsilon_{i}$, then

$$
\lambda \leq \mu \Leftrightarrow \sum_{i=1}^{n} a_{i}=\sum_{i=1}^{n} b_{i} \quad \text { and } \sum_{i=1}^{k} a_{i} \leq \sum_{i=1}^{k} b_{i} \text { for all } 1 \leq k \leq n
$$

For example, for $T_{2}$ the characters smaller than the circled characters all lie on the indicated halflines


Let $V$ be a finite dimensional $G L_{n}$-representation. As a $T_{n}$-representation (using the diagonal embedding $T_{n} \hookrightarrow G L_{n}$ ) $V$ decomposes as a direct sum of weight spaces

$$
V=\oplus_{\lambda \in X\left(T_{n}\right)} V_{\lambda}
$$

We will now investigate the restriction imposed on the $V_{\lambda}$ by the $G L_{n}$ action. To begin, consider the permutation matrixes $P_{\sigma}$ for $\sigma \in S_{n}$, having a 1 at all positions $(i, \sigma(i))$ and zeroes elsewhere. There is a natural action of the symmetric group $S_{n}$ on $T_{n}$ and on $X\left(T_{n}\right)$ defined by $\sigma . x=\sigma \cdot\left(x_{1}, \ldots, x_{n}\right)=\left(x_{\sigma(1)}, \ldots, x_{\sigma(n)}\right)$ and $\sigma \cdot \lambda=\sigma \cdot\left(\sum a_{i} \epsilon_{i}\right)=\sum a_{i} \epsilon_{\sigma(i)}$. If $v \in V_{\lambda}$ and $x=\operatorname{diag}\left(x_{1}, \ldots, x_{n}\right) \in T_{n} \longleftrightarrow G L_{n}$, we have

$$
\begin{aligned}
x \cdot\left(P_{\sigma} \cdot v\right) & =P_{\sigma} \cdot\left(P_{\sigma}^{-1} x P_{\sigma}\right) \cdot v=P_{\sigma} \cdot(\sigma \cdot x) \cdot v \\
& =P_{\sigma} \cdot(\sigma \lambda(x) v)=\sigma \lambda(x) P_{\sigma} \cdot v
\end{aligned}
$$

that is, $P_{\sigma} . v \in V_{\sigma . \lambda}$. So we have that $\operatorname{dim} V_{\lambda}=\operatorname{dim} V_{\sigma \lambda}$ for all $\lambda \in X\left(T_{n}\right)$ and all $\sigma \in S_{n}$. Further, consider the elementary matrices

$$
x_{i j}(t)=\left[\begin{array}{ccccccc}
1 & & & & & & \\
& \ddots & & & & & \\
& & 1 & \ldots & t & & \\
& & & \ddots & \vdots & & \\
& & & & 1 & & \\
& & & & & \ddots & \\
& & & & & & 1
\end{array}\right]
$$

having 1's on the diagonal, $t \in \mathbb{C}$ at place $(i, j)$ and zeroes elsewhere. One calculates that for every $x=\operatorname{diag}\left(x_{1}, \ldots, x_{n}\right) \in T_{n} \longleftrightarrow G L_{n}$ we have

$$
x x_{i j}(t) x^{-1}=x_{i j}\left(\left(x_{i}-x_{j}\right) t\right)=x_{i j}\left(\left(\epsilon_{i}-\epsilon_{j}\right)(x) . t\right)
$$

If $v \in V_{\lambda}$, then as the map $t \mapsto x_{i j}(t) . v$ is a polynomial map in $t$ we can write with respect to a basis $\left\{w_{1}, \ldots, w_{l}\right\}$ of $V$ that

$$
x_{i j}(t) . v=\sum_{a=1}^{l}\left(\sum_{b \geq 0} c_{a b} t^{b}\right) w_{a}=\sum_{b \geq 0} t^{b} v_{b} \quad \text { with } \quad v_{b}=\sum_{a=1}^{l} c_{a b} w_{a}
$$

As $x_{i j}(0)=\mathbb{1}_{n}$ we have $v_{0}=v$ and we claim that $v_{b} \in V_{\lambda+b\left(\epsilon_{i}-\epsilon_{j}\right)}$. Indeed, for all $x=\operatorname{diag}\left(x_{1}, \ldots, x_{n}\right) \in T_{n} \longrightarrow G L_{n}$ and $t \in \mathbb{C}$ we have

$$
\begin{array}{rlc}
x \cdot x_{i j}(t) \cdot v & = & \left(x x_{i j}(t) x^{-1}\right) \cdot x \cdot v=x_{i j}\left(\left(\epsilon_{i}-\epsilon_{j}\right)(x) \cdot t\right) \cdot(\lambda(x) v) \\
& = & \lambda(x) \sum_{b \geq 0}\left(\left(\epsilon_{i}-\epsilon_{j}\right)(x) \cdot t\right)^{b} v_{b}=\lambda(x) \sum_{b \geq 0} t^{b}\left(\left(\epsilon_{i}-\epsilon_{j}\right)(x)\right)^{b} v_{b} \\
& = & \sum_{b \geq 0} t^{b}\left(\lambda+b\left(\epsilon_{i}-\epsilon_{j}\right)\right)(x) v_{b}
\end{array}
$$

and on the other hand we have the equalities

$$
x \cdot x_{i j}(t) \cdot v=x \cdot\left(\sum_{b \geq 0} t^{b} v_{b}\right)=\sum_{b \geq 0} t^{b}\left(x \cdot v_{b}\right)
$$

So, comparing terms we establish our claim as

$$
x \cdot v_{b}=\left(\lambda+b\left(\epsilon_{i}-\epsilon_{j}\right)\right)(x) v_{b} \quad \text { that is, } \quad v_{b} \in V_{\lambda+b\left(\epsilon_{i}-\epsilon_{j}\right)} .
$$

## 10b. Borel subgroups.

Let $\Delta_{n}$ be the subgroup of $G L_{n}$ consisting of upper triangular matrices with 1's on the diagonal and let $\mathbf{\nabla}_{n}$ be the corresponding subgroup of lower triangular matrices in $G L_{n}$. Consider the product map

$$
\boldsymbol{\nabla}_{n} \times T_{n} \times \mathbf{\Delta}_{n} \xrightarrow{\text { prod }} G L_{n} \quad(l, d, u) \mapsto l d u
$$

The differential of prod in the unit element $\left(\mathbb{1}_{n}, \mathbb{1}_{n}, \mathbb{1}_{n}\right)$ can be computed by the $\epsilon$-method expanding

$$
\left(\mathbb{1}_{n}+\epsilon L\right)\left(\mathbb{1}_{n}+\epsilon D\right)\left(\mathbb{1}_{n}+\epsilon U\right)=\mathbb{1}_{n}+\epsilon(L+D+U)
$$

with $L$ (resp. $U$ ) a strictly lower (resp. upper) triangular matrix and $D$ a diagonal matrix. Hence, $d \operatorname{prod}_{\left(\mathbb{1}_{n}, \mathbb{1}_{n}, \mathbb{1}_{n}\right)}$ is an isomorphism and so the product map has a dense image in $G L_{n}$. Therefore,

$$
V=V_{l_{1}} \oplus \ldots \oplus V_{l_{s}} \quad V_{l}=\oplus\left\{V_{\lambda} \mid \lambda=\sum a_{i} \epsilon_{i} \text { with } \sum_{i} a_{i}=l\right\}
$$

as $\Delta_{n}$ and $\nabla_{n}$ are generated by the respective $x_{i j}(t)$ and by the above equality we have that for any $v \in V_{l}$ that $x_{i j}(t) v \in V_{l}$. Hence, the non-zero weight spaces $V_{\lambda}$ of a $G L_{n}$-representation $V$ lie in a finite number of hyperplanes where they form a configuration stable under the action of $S_{n}$. For $T_{2}$ the configuration of non-zero weight spaces is symmetric with respect to the main diagonal and constrained to the indicated boxes


The dashed circled weights are the highest weights with respect to the ordering on $X\left(T_{2}\right)$. For general $n$ we similarly have for each component $V_{l}$ a highest weight $V_{\lambda}$ with respect to the ordering on $X\left(T_{n}\right)$. An element $v \in V_{\lambda}$ is then called a highest weight vector of $V$ (or of the component $V_{l}$ ). Let $v_{b}$ be the vectors in $V_{\lambda+b\left(\epsilon_{i}-\epsilon_{j}\right)}$ introduced before, then again using the dense map $\boldsymbol{\nabla}_{n} \times T_{n} \times \mathbf{\Delta}_{n} \longrightarrow G L_{n}$ we see that the subspace of $V$ spanned by $v=v_{0}$ and the $v_{b}$ is a simple $G L_{n}$-subrepresentation of $V$. We now have all the necessary ingredients to finish the proof of :

Theorem 10.2. There is a one-to-one correspondence between

1. isomorphism classes of simple $G L_{n}$-representations, and
2. $\lambda=\sum_{i} a_{i} \epsilon_{i} \in X\left(T_{n}\right)$ which are highest weights, that is,

$$
a_{1} \geq a_{2} \geq \ldots \geq a_{n}
$$

We have already seen that a highest weight vector generates a simple component. Observe that a highest weight vector spans a one-dimensional
subspace which is stable under the action of the Borel subgroup $B_{n}$ of upper triangular elements of $G L_{n}$. Further, if $V$ is a simple $G L_{n}$-representation then using the $S_{n}$-action we know that $V$ has a highest weight vector. Remains to prove that two $G L_{n}$-representations spanned by a highest weight vector of the same weight are isomorphic and that every $\lambda=\sum_{i} a_{i} \epsilon_{i}$ with $a_{1} \geq \ldots \geq a_{n}$ occurs as a highest weight for a $G L_{n}$-representation. First, let $V$ and $V^{\prime}$ be simple $G L_{n}$-representations with highest weight vectors $v$ and $v^{\prime}$ of weight $\lambda$. Then, $W=V \oplus V^{\prime}$ is a $G L_{n}$-representation and $w=\left(v, v^{\prime}\right)$ a vector of highest weight $\lambda$ so generates a simple component $W^{\prime}$ of $W$ with $W_{\lambda}^{\prime}=\mathbb{C} w . V \cap W^{\prime}$ is a $G L_{n}$-subrepresentation of $V$ so is either 0 or $V$. As $W_{\lambda}^{\prime}=\mathbb{C} w \not \subset V_{\lambda}=\mathbb{C}(v, 0)$ this intersection must be zero. Then, the composition

$$
W^{\prime} \hookrightarrow V \oplus V^{\prime} \xrightarrow{p r_{1}} \sim V
$$

is into and gives therefore an isomorphism $W \simeq V$. Repeating the argument with $V^{\prime}$ instead of $V$ we deduce that $V \simeq V^{\prime}$.

As for the existence, let $\lambda=\sum_{i} a_{i} \epsilon_{i}$ with $a_{1} \geq \ldots \geq a_{n}$ and consider $\omega_{i}=\epsilon_{1}+\ldots+\epsilon_{i}$, then we have

$$
\lambda=b_{1} \omega_{1}+\ldots b_{n} \omega_{n}
$$

with $b_{n}=a_{n} \in \mathbb{Z}$ and all $b_{i}=a_{i}-a_{i+1} \in \mathbb{N}$ for $1 \leq i \leq n-1$. Recall the construction of exterior product $\wedge^{i} \mathbb{C}^{n}$ for $1 \leq i \leq n$ that is, the subspace of $\mathbb{C}^{n} \otimes \ldots \otimes \mathbb{C}^{n}$ ( $i$ terms) spanned by the anti-symmetric tensors which is clearly a $G L_{n}$ - subrepresentation of $\mathbb{C}^{n} \otimes \ldots \otimes \mathbb{C}^{n}$ with the diagonal action on the tensorproduct, $g \cdot\left(c_{1} \otimes \ldots \otimes c_{i}\right)=\left(g \cdot c_{1}\right) \otimes \ldots \otimes\left(g \cdot c_{i}\right)$. Consider the $G L_{n}$-representation

$$
V=\left(\wedge^{1} \mathbb{C}^{n}\right)^{\otimes b_{1}} \otimes\left(\wedge^{2} \mathbb{C}^{2}\right)^{\otimes b_{2}} \otimes \ldots \otimes\left(\wedge^{n} \mathbb{C}^{n}\right)^{\otimes b_{n}}
$$

where $\left(\wedge^{i} \mathbb{C}^{n}\right)^{\otimes k}$ is the $k$-fold tensorproduct of $\wedge^{i} \mathbb{C}^{n}$ for $k \in \mathbb{N}_{+},\left(\wedge^{i} \mathbb{C}^{n}\right)^{0}=$ $\mathbb{C}_{\text {triv }}$ the trivial one-dimensional $G L n$-representation and $\left(\wedge^{n} \mathbb{C}^{n}\right)^{\otimes m}$ is the one-dimensional $G L_{n}$-representation defined by $g \mapsto(\operatorname{det} g)^{m}$ (note that $m<$ 0 is possible). If $\left\{e_{1}, \ldots, e_{n}\right\}$ is the standard basis of $\mathbb{C}^{n}$, then $e_{1} \wedge e_{2} \wedge \ldots \wedge e_{i}$ is stable under the action of the Borel subgroup $B_{n}$ and has weight $\omega_{i}$ as clearly

$$
\operatorname{diag}\left(x_{1}, \ldots, x_{n}\right) \cdot\left(e_{1} \wedge e_{2} \wedge \ldots \wedge e_{i}\right)=\prod_{j=1}^{i} x_{j}\left(e_{1} \wedge e_{2} \wedge \ldots \wedge e_{i}\right)
$$

But then, the following vector $v$ is stable under the Borel $B_{n}$ and has weight $\lambda$

$$
v= \begin{cases}e_{1}^{\otimes b_{1}} \otimes\left(e_{1} \wedge e_{2}\right)^{\otimes b_{2}} \otimes \ldots \otimes\left(e_{1} \wedge \ldots \wedge e_{n}\right)^{\otimes b_{n}} & \text { when } b_{n} \in \mathbb{N}, \text { and } \\ e_{1}^{\otimes b_{1}} \otimes\left(e_{1} \wedge e_{2}\right)^{\otimes b_{2}} \otimes \ldots \otimes\left(f_{1} \wedge \ldots \wedge f_{n}\right)^{\otimes-b_{n}} & \text { when } b_{n}<0\end{cases}
$$

where $\left\{f_{1}, \ldots, f_{n}\right\}$ is the dual basis, that is the standard basis of $\mathbb{C}^{n *}$. Hence, $v$ generates a simple $G L_{n}$-representation of highest weight $\lambda$. This concludes the proof of theorem 10.2 . That is, we can identify the set $\Omega_{G L_{n}}$ of isomorphism classes of simple $G L_{n}$-representations with the subset of $X\left(T_{n}\right)$

$$
\Omega_{G L_{n}}=\mathbb{N} \omega_{1} \oplus \mathbb{N} \omega_{2} \oplus \ldots \oplus \mathbb{N} \omega_{n-1} \oplus \mathbb{Z} \omega_{n}
$$

For example, the simple $G L_{2}$-representations are in one-to-one correspondence to the subset of $\mathbb{Z}^{2}=X\left(T_{2}\right)$


We will denote the simple $G L_{n}$-representation of highest weight $\lambda$ by $S_{\lambda}$. For example,

$$
\mathbb{C}^{n}=S_{(1,0, \ldots, 0)} \quad \text { and } \quad \mathbb{C}^{n *}=S_{(0, \ldots, 0,-1)}
$$

Consider the action of $G L_{n}$ on $M_{n}(\mathbb{C})$ by conjugation. If $e_{i j}$ is the matrix having a 1 at place $(i, j)$ and zeroes elsewhere, then we have

$$
M_{n}(\mathbb{C})=M_{n}(\mathbb{C})_{0} \oplus \oplus_{i \neq j} M_{n}(\mathbb{C})_{\epsilon_{i}-\epsilon_{j}} \quad \text { with } M_{n}(\mathbb{C})_{\epsilon_{i}-\epsilon_{j}}=\mathbb{C} e_{i j}
$$

and $M_{n}(\mathbb{C})_{0}$ is the space of diagonal matrices. The weightspaces left invariant by the $B_{n}$-action are $\mathbb{C} e_{1 n}$ and $\mathbb{C} \mathbb{1}_{n}$. The first generates the simple representation with highest weight $(1,0, \ldots, 0,-1)$, the second is the trivial representation (highest weight $(0, \ldots, 0)$. That is, as $G L_{n}$-representations

$$
M_{n}(\mathbb{C})=M_{n}(\mathbb{C})_{(0, \ldots, 0)} \oplus M_{n}(\mathbb{C})_{(1,0, \ldots, 0,-1)}=\mathbb{C} \mathbb{1}_{n} \oplus M_{n}^{0}(\mathbb{C})
$$

where $M_{n}^{0}(\mathbb{C})$ is the space of all trace zero matrices. The weights $\epsilon_{i}-\epsilon_{j}$ are called the roots of $G L_{n}$. The roots $\epsilon_{i}-\epsilon_{j}$ with $i<j$ are called positive. This allows a reformulation of the ordering in $X\left(T_{n}\right):$ for $\lambda, \mu \in X\left(T_{n}\right)$ we have $\lambda \leq \mu$ if and only if $\mu-\lambda$ can be written as a sum of positive roots.

## 10c. $G L_{n}$-varieties.

Combining reductivity of $G L_{n}$ with the combinatorial description of the isomorphism classes of simple $G L_{n}$-representations allows us to determine all
linear actions of $G L_{n}$ on a finite dimensional vectorspace $V$. To formalize the notion of $G L_{n}$-variety we consider the induced action of $G L_{n}$ on the polynomial functions $f$ on $V$. Considering the diagram

we see that this action is defined by the rule $g \cdot f(v)=f\left(g^{-1} \cdot v\right)$. Alternatively, if the dimension of $V$ is $l$, then the ring of all polynomial maps on $V$ is $\mathbb{C}[V]=\mathbb{C}\left[x_{1}, \ldots, x_{l}\right]$ where the linear forms $\mathbb{C} x_{1}+\ldots+\mathbb{C} x_{l}$ can be identified with the dual vectorspace $V^{*} . V^{*}$ is a $G L_{n}$-representation with action $g . \phi$ for $V \xrightarrow{\phi} \mathbb{C}$ in $\mathbb{V}^{*}$ defined by $(g . \phi)(v)=\phi\left(g^{-1} . v\right)$. Hence, the action of $G L_{n}$ on the polynomial ring coincides with the natural action of $G L_{n}$ on the symmetric algebra of $V^{*}$

$$
S\left(V^{*}\right)=\oplus_{i=0}^{\infty} S^{i} V^{*} \simeq \mathbb{C}\left[x_{1}, \ldots, x_{l}\right]
$$

where $S^{i} V^{*}$ is the subspace of symmetric tensors in $V^{*} \otimes \ldots \otimes V^{*}(i$ terms). The action of $G L_{n}$ on $\mathbb{C}[V]$ is locally finite, that is, any finite set of polynomials $\left\{f_{1}, \ldots, f_{a}\right\} \subset \mathbb{C}[V]$ is contained in a finite dimensional $G L_{n}$ subrepresentation $W$ of $\mathbb{C}[V]$. Indeed, let $k$ be such that all $f_{i}$ have total degree at most $k$, then they are all contained in the finite dimensional $G L_{n}$ subrepresentation $\oplus_{i=0}^{k} S^{i} V^{*}$.

Recall that the Hilbert basis theorem asserts that any ideal $I \triangleleft \mathbb{C}[V]$ is finitely generated. We say that an ideal $I \triangleleft \mathbb{C}[V]$ is $G L_{n}$-stable if $g . I \subset I$ for all $g \in G L_{n}$.

Definition 10.3. A $G L_{n}$-variety is a couple $(V, I)$ where

- $V$ is a finite dimensional $G L_{n}$-representation, and
- $I$ is a $G L_{n}$-stable ideal of $\mathbb{C}[V]=S\left(V^{*}\right)$.

Examples of $G L_{n}$-varieties.
Jordan forms

$$
\begin{aligned}
& V=M_{n}(\mathbb{C}) \\
& g \cdot m=g m g^{-1} \\
& \mathbb{C}[V]=\mathbb{C}\left[m_{11}, \ldots, m_{n n}\right] \\
& I=0
\end{aligned}
$$

Examples of $G L_{n}$-varieties.
Dynamical systems

$$
\begin{aligned}
& V=M_{n \times m}(\mathbb{C}) \oplus M_{n}(\mathbb{C}) \oplus M_{p \times n}(\mathbb{C}) \\
& g \cdot(A, B, C)=\left(g A, g B g^{-1}, C g^{-1}\right) \\
& \mathbb{C}[V]=\mathbb{C}\left[a_{11}, \ldots, a_{n m}, b_{11}, \ldots, b_{n n}, c_{11}, \ldots, c_{p n}\right] \\
& I=0
\end{aligned}
$$

Hilbert schemes $\quad V=\mathbb{C}^{n} \oplus M_{n}(\mathbb{C}) \oplus M_{n}(\mathbb{C})$
$g .(v, X, Y)=\left(g v, g X g^{-1}, g Y g^{-1}\right)$
$\mathbb{C}[V]=\mathbb{C}\left[v_{1}, \ldots, v_{n}, x_{11}, \ldots, x_{n n}, y_{11}, \ldots, y_{n n}\right]$ $I=\left(\sum_{k} x_{i k} y_{k j} \mid \forall i, j\right)$

