# NONTRIVIAL LINEAR PROJECTIONS ON THE GRASSMANNIAN $\mathrm{Gr}_{3}\left(\mathbb{C}^{6}\right)$ 

YANHE HUANG, GEORGE PETROULAKIS, FRANK SOTTILE, AND IGOR ZELENKO


#### Abstract

A typical linear projection of the Grassmannian in its Plücker embedding is injective, unless its image is a projective space. A notable exception are self-adjoint linear projections, which have even degree. We consider linear projections of $\mathrm{Gr}_{3} \mathbb{C}^{6}$ with low-dimensional centers of projection. When the center has dimension less than five, we show that the projection has degree 1 . When the center has dimension five and the projection has degree greater than 1 , we show that it is self-adjoint.


## 1. Introduction

Consider a linear ordinary differential operator (ODO) of order $n$

$$
\begin{equation*}
L x(t)=x^{(n)}(t)+a_{n-1}(t) x^{(n-1)}(t)+\cdots+a_{0}(t) x(t) \tag{1.1}
\end{equation*}
$$

where $a_{0}, \ldots, a_{n-1}$ are complex-valued continuous functions on an interval $I \subset \mathbb{R}$. Let $V_{L}$ be the space of complex-valued solutions of the homogeneous equation $L x=0$.

The Wronskian of $m$ smooth functions $f_{1}(t), \ldots, f_{m}(t)$ on $I$ is the determinant

$$
\operatorname{Wr}\left(f_{1}(t), f_{2}(t), \ldots, f_{m}(t)\right):=\operatorname{det}\left(\begin{array}{cccc}
f_{1}(t) & f_{2}(t) & \cdots & f_{m}(t) \\
f_{1}^{\prime}(t) & f_{2}^{\prime}(t) & \cdots & f_{m}^{\prime}(t) \\
\vdots & \vdots & \ddots & \vdots \\
f_{1}^{(m-1)}(t) & f_{2}^{(m-1)}(t) & \cdots & f_{m}^{(m-1)}(t)
\end{array}\right)
$$

The Wronskian $\operatorname{Wr}\left(f_{1}(t), \ldots, f_{m}(t)\right)$ is not identically zero when $f_{1}(t), \ldots, f_{m}(t)$ form a basis of an $m$-dimensional subspace $\Lambda$ in $V_{L}$. If $g_{1}(t), \ldots, g_{m}(t)$ is another basis, then

$$
\mathrm{Wr}\left(g_{1}(t), g_{2}(t), \ldots, g_{m}(t)\right)=c \operatorname{Wr}\left(f_{1}(t), f_{2}(t), \ldots, f_{m}(t)\right),
$$

where $c$ is the determinant of the transition matrix between the bases. Therefore, the one-dimensional linear subspace of $C^{\infty}(I)$ spanned by the Wronskian $\operatorname{Wr}\left(f_{1}(t), \ldots, f_{m}(t)\right)$ depends only upon $\Lambda$. This element of the projective space $\mathbb{P} C^{\infty}(I)$ is called the Wronskian of the subspace $\Lambda$. This defines the Wronski map $\mathrm{Wr}_{L, m}$ from the Grassmannian $\mathrm{Gr}_{m} V_{L}$ of $m$-dimensional subspaces of $V_{L}$ to $\mathbb{P} C^{\infty}(I)$.

[^0]For complex algebraic varieties $X, Y$ of the same dimension and a dominant map $F: X \rightarrow Y$, the number of points in a preimage $F^{-1}(y)$ for $y \in Y$ is constant over an open dense subset of $Y$. This constant number is the degree of the map $F$ [6].

Consider this for the Wronski map $\mathrm{Wr}_{L, m}$ when the image of $\mathrm{Gr}_{m} V_{L}$ has the same dimension as $\mathrm{Gr}_{m} V_{L}$. For generic linear ODO $L$ of order $n$ and any $m \in\{2, \ldots, n-1\}$ the Wronski map $\mathrm{Wr}_{L, m}$ is injective (see Remark (1.1) and so $\mathrm{Wr}_{L, m}$ has degree 1. For any $L$, is it injective when $m=1$ or $m=n-1$. We are interested in the following question.
Question 1. Under what conditions on a linear ODO L of order $n$ and on $1<m<n-1$ does the Wronski map $\mathrm{Wr}_{L, m}$ have degree greater than 1?

The classical Wronski map is when $V$ is the space of polynomials of degree $n-1$. This corresponds to the ODO $L_{0} x(t)=x^{(n)}(t)$. Work of Schubert in 1886 [9], combined with a result of Eisenbud and Harris in 1983 [4] shows that the Wronski map $\mathrm{Wr}_{L_{0}, m}$ has degree

$$
\begin{equation*}
\frac{1!2!\cdots(n-m-1)!\cdot(m(n-m))!}{m!(m+1)!\cdots(n-1)!} \tag{1.2}
\end{equation*}
$$

The degree exceeds 1 except in the trivial cases of $m=1$ or $m=n-1$.
Three of us addressed Question 1 in a previous paper [7]. The operator

$$
\begin{equation*}
L^{*} x(t):=(-1)^{n} x^{(n)}(t)+\sum_{i=1}^{n-1}(-1)^{i}\left(a_{i} x\right)^{(i)}(t) \tag{1.3}
\end{equation*}
$$

is (formally) adjoint to the operator $L$ (1.1). An operator $L$ is a (formally) self-adjoint differential operator if $L^{*}=L$. When $L$ is self-adjoint, its order $n$ is even.

Two linear ODOs $L$ and $\widetilde{L}$ on $I$ are equivalent if there exists a smooth nonvanishing function $\mu$ on $I$ such that

$$
\widetilde{L} x=\frac{1}{\mu} L(\mu x) .
$$

We paraphrase two results from [7]. For both, $L$ is a linear ODO of order $n$.
Theorem 2.9 of [7] If $L$ is equivalent to a self-adjoint operator and $n=2 m$, then the Wronski map $\mathrm{Wr}_{L, m}$ has even degree.

Corollary 1.8 of [7] If the Wronski map $\mathrm{Wr}_{L, m}$ has degree 2, then $n=2 m$ and $L$ is equivalent to a self-adjoint linear operator.

The proof of [7, Thm. 2.9] is based on two observations. First, if $L$ is equivalent to a self-adjoint operator then the space $V_{L}$ is endowed with a canonical (up to a nonzero scaling) symplectic structure $\sigma_{L}$. Second, if $\Lambda^{\angle}$ is the skew-orthogonal complement of an $m$-dimensional subspace $\Lambda$ of $V_{L}$ with respect to the form $\sigma_{L}$, then

$$
\begin{equation*}
\mathrm{Wr}_{L, m}\left(\Lambda^{\llcorner }\right)=\mathrm{Wr}_{L, m}(\Lambda) \tag{1.4}
\end{equation*}
$$

so that the Wronskian is preserved under taking skew-orthogonal complement.
From (1.2) it follows that for the ODO $L_{0} x(t)=x^{(n)}(t)$ with $n \geq 5$ and $m \notin\{1, n-1\}$ the Wronski map $\mathrm{Wr}_{L_{0}, m}$ has degree greater than 2 . Thus $n=2 m$ is not necessary for the degree of the Wronski map to exceed 1.

Question 2. When $n=2 m$, does the statement of [7, Cor. 1.8] generalize as follows: If the Wronski map $\mathrm{Wr}_{L, m}$ of a 2 m-th order linear ODO L has degree greater than 2 , is $L$ equivalent to a self-adjoint operator?

We address a generalization of Question 2. The Grassmannian $\mathrm{Gr}_{m} V_{L}$ is a subvariety of Plücker space $\mathbb{P} \bigwedge^{m} V_{L}$. Given a linear subspace $\mathbb{P} Z \subset \mathbb{P} \bigwedge^{m} V_{L}(Z$ is a linear subspace of $\left.\bigwedge^{m} V_{L}\right)$, the linear projection with center $\mathbb{P} Z$ is the map $\mathbb{P} \bigwedge^{m} V_{L} \backslash \mathbb{P} Z \rightarrow \mathbb{P}\left(\bigwedge^{m} V_{L}\right) / Z$ induced by the map $\bigwedge^{m} V_{L} \rightarrow\left(\bigwedge^{m} V_{L}\right) / Z$. When $\mathbb{P} Z$ is disjoint from the Grassmannian, it induces the linear projection $\pi_{Z}: \mathrm{Gr}_{m} V_{L} \rightarrow \mathbb{P}\left(\bigwedge^{m} V_{L}\right) / Z$.

Proposition 2.3 of [7] identifies the Wronski map with a linear projection. We explain that. Given a basis $f_{1}, \ldots, f_{n}$ for $V_{L}$, let $f_{1}^{*}, \ldots f_{n}^{*} \in V_{L}^{*}$ be its dual basis and set

$$
\begin{equation*}
c(t):=\sum_{i=1}^{n} f_{i}(t) f_{i}^{*} \in V_{L}^{*}, \text { for } t \in I \tag{1.5}
\end{equation*}
$$

Fix $m \in\{1, \ldots, n-1\}$ and define the following subspace of $\bigwedge^{m} V_{L}^{*}$,

$$
\begin{equation*}
X_{L}:=\left\langle c(t) \wedge c^{\prime}(t) \wedge \cdots \wedge c^{(m-1)}(t) \mid t \in I\right\rangle \tag{1.6}
\end{equation*}
$$

where

$$
c^{(j)}(t)=\sum_{i=1}^{n} f_{i}^{(j)}(t) f_{i}^{*} \in V_{L}^{*}
$$

By [7, Prop. 2.3], the Wronski map takes values in the space $X_{L}^{*}$ dual to $X_{L}$, which is $\left(\bigwedge^{m} V\right) / X_{L}^{\perp}$, where

$$
X_{L}^{\perp}=\left\{w \in \bigwedge^{m} V_{L} \mid \omega(v)=0 \quad \forall v \in X_{L}\right\}
$$

is the annihilator of $X_{L}$ (for details see [7, pp. 755-6]).
Remark 1.1. For generic linear ODO $L, X_{L}=\bigwedge^{m} V_{L}^{*}$, which implies that the Wronski map is injective (this is a consequence of Proposition 3.1 below, as $X_{L}^{\perp}=0$ ).
Remark 1.2. As a consequence of [7, Sect. 2.3], a linear ODO of order $2 m$ is self-adjoint if and only if there exists a symplectic form $\sigma$ on $V_{L}^{*}$ such that

$$
X_{L}^{\perp} \supseteq \mathbb{C} \sigma \wedge \bigwedge^{m-2} V_{L}
$$

Moreover, the canonical symplectic form on $V_{L}$ is induced by the form $\sigma$ through the identification of $V_{L}$ with $V_{L}^{*}$ via $\sigma$. This inclusion implies that

$$
\operatorname{dim} X_{L}^{\perp} \geq \operatorname{dim} \bigwedge^{m-2} V_{L}^{*}=\binom{2 m}{m-2}
$$

with equality for a generic self-adjoint linear ODO of order $2 m$. When $m=3$ and $L$ is self-adjoint, the minimal possible dimension of $X_{L}^{\perp}$ is 6 .

Let $V$ be an even-dimensional complex vector space and $1<m<\operatorname{dim} V$. A linear subspace $Z \subset \bigwedge^{m} V$ is self-adjoint if there exists a symplectic form $\sigma$ on $V^{*}$ such that

$$
Z \supseteq \mathbb{C} \sigma \wedge \wedge^{m-2} V
$$

We state our main results.

Theorem 3.6. When $m=2$ and $n=4$, if $\mathbb{P} Z$ is a linear subspace disjoint from $\mathrm{Gr}_{2} \mathbb{C}^{4}$, then $Z$ is self-adjoint.

When $m=3$ and $n=6$, we consider centers $Z$ of projective dimensions four or five.
Proposition 1.3. Suppose that $m=3$ and $n=6$. Let $Z \subset \bigwedge^{3} \mathbb{C}^{6}$ be a linear subspace with $\mathbb{P} Z$ disjoint from $\mathrm{Gr}_{3} \mathbb{C}^{6}$.
(1) [Corollary 3.17 If $\operatorname{dim} \mathbb{P} Z \leq 4$, then $\pi_{Z}$ has degree 1 .
(2) [Theorem 3.18] If $\operatorname{dim} \mathbb{P} Z=5$, then $\pi_{Z}$ has degree greater than 1 if and only if $Z$ is self-adjoint.

We deduce our main results concerning Question 2.
Theorem 1.4. Let $L$ be a linear $O D O$ of order 4. Then the degree of the Wronski map $\mathrm{Wr}_{L, 2}$ exceeds 1 if and only if $L$ is equivalent to a self-adjoint linear ODO.

Theorem 1.5. Let $L$ be a linear ODO of order 6. Then the following statements hold.
(1) If $\operatorname{dim} X_{L}^{\perp} \leq 5$, then the degree of the Wronski map $\operatorname{Wr}_{L, 3}$ is equal to 1 .
(2) If $\operatorname{dim} X_{L}^{\perp}=6$, then the degree of the Wronski map $\mathrm{Wr}_{L, 3}$ exceeds 1 if and only if $L$ is equivalent to a self-adjoint linear ODO.

In the next section, we discuss an application of Theorems 1.4 and 1.5 to pole placement in linear systems theory. We prove our main results in Section 3 ,

## 2. Application to Pole Placement for Constant Output Feedback

For a background on linear systems theory, see [2]. A state-space realization of a (strictly proper) $m$-input $p$-output linear system is a triple $\Sigma=(A, B, C)$ of matrices of sizes $N \times N, N \times m$, and $p \times N$. This defines a system of first order constant coefficient linear differential equations,

$$
\begin{equation*}
\dot{x}=A x+B u \quad \text { and } \quad y=C x \tag{2.1}
\end{equation*}
$$

where $x \in \mathbb{C}^{N}, u \in \mathbb{C}^{m}$, and $y \in \mathbb{C}^{p}$ are functions of $t \in \mathbb{C}$ (and $\dot{x}=\frac{d}{d t} x$ ). Applying Laplace transform $(u(t) \mapsto \hat{u}(s))$ and assuming that $x(0)=0$, we eliminate $\widehat{x}$ to obtain

$$
\widehat{y}(s)=C(s I-A)^{-1} B \widehat{u}(s)=G(s) \widehat{u}(s),
$$

where $G(s):=C(s I-A)^{-1} B$ is the transfer function of (2.1). This $p \times m$ matrix of rational functions has poles at the eigenvalues of $A$.

A linear system may be controlled with output feedback, setting $u=K y$, where $K$ is a constant $m \times p$ matrix. Substitution in (2.1) and elimination gives the closed loop system,

$$
\dot{x}=(A+B K C) x
$$

whose transfer function has poles at the zeroes of the characteristic polynomial

$$
\begin{equation*}
P_{\Sigma}(K)=P_{\Sigma}:=\operatorname{det}(s I-(A+B K C)) \tag{2.2}
\end{equation*}
$$

The map $K \mapsto P_{\Sigma}(K)$ is called the pole placement map. Given a system (2.1) with state-space realization $\Sigma$ and poles $z=\left\{z_{1}, \ldots, z_{N}\right\} \subset \mathbb{C}$, the pole placement problem asks for a matrix $K$ such that $P_{\Sigma}(K)$ vanishes at the points of $z$. This is only possible for
general $z$ if $N \leq m p$ [1]. We are interested when $N \geq m p$ and the pole placement map is a nontrivial branched cover of its image.

Using the injection $\operatorname{Mat}_{m \times p} \mathbb{C} \rightarrow \mathrm{Gr}_{p} \mathbb{C}^{m+p}$ where $K$ is sent to the column space of the matrix $\binom{K}{I_{p}}$, standard manipulations show that the pole placement map is a linear projection of $\operatorname{Gr}_{p} \mathbb{C}^{m+p}$. The map that sends $s \in \mathbb{P}^{1}$ to the column space of $\binom{I_{m}}{G(s)}$ defines the Hermann-Martin curve $\gamma_{\Sigma}: \mathbb{P}^{1} \rightarrow \operatorname{Gr}_{m} \mathbb{C}^{m+p}$ [8]. Its degree is the McMillan degree, which is the minimal number $N$ in a state-space realization giving the transfer function $G(s)$. Such a minimal representation is observable and controllable [2].

If $X_{\Sigma} \subset \Lambda^{m} \mathbb{C}^{m+p}$ is the linear span of the image of the curve $\gamma_{\Sigma}$ and $Z:=X_{\Sigma}^{\perp}$ is its annihilator in $\bigwedge^{p} \mathbb{C}^{m+p}$, then the pole placement map $P_{\Sigma}$ is the linear projection $\pi_{Z}$, and we may identify the quotient $X_{\Sigma}^{*}=\left(\bigwedge^{p} \mathbb{C}^{m+p}\right) / Z$ with the space of polynomials of degree at most $N$. The pole placement map is proper if $\emptyset \neq \mathbb{P} Z$ is disjoint from the Grassmannian $\mathrm{Gr}_{p} \mathbb{C}^{m+p}$. This terminology is not standard in systems theory.

Consider the following change of coordinates in the state, input, and output spaces

$$
\begin{equation*}
x=R \widetilde{x}, \quad u=Q \widetilde{y}+W \widetilde{u}, \quad \text { and } \quad y=T \widetilde{y} \tag{2.3}
\end{equation*}
$$

where $R, W$, and $T$ are invertible matrices and $Q$ is a $m \times p$ matrix. The transformation of the space $\mathbb{C}^{N} \times \mathbb{C}^{m} \times \mathbb{C}^{p}$ given by (2.3) is a state-feedback transformation. Substituting (2.3) into (2.1), we obtain a new state-space realization in $(\widetilde{x}, \widetilde{u}, \widetilde{y})$,

$$
\dot{\tilde{x}}=\widetilde{A} \widetilde{x}+\widetilde{B} \widetilde{u} \quad \text { and } \quad \widetilde{y}=\widetilde{C} \widetilde{x}
$$

given by the triple of matrices $\widetilde{\Sigma}=(\widetilde{A}, \widetilde{B}, \widetilde{C})$, where

$$
\begin{equation*}
\widetilde{A}=R^{-1}\left(A+B Q T^{-1} C\right) R, \quad \widetilde{B}=R^{-1} B W, \quad \text { and } \quad \widetilde{C}=T^{-1} C R \tag{2.4}
\end{equation*}
$$

Two realizations are state-feedback equivalent if one is a state-feedback transformation of the other. The following is standard.
Proposition 2.1. Equivalent state-space realizations have equivalent Hermann-Martin curves, where the equivalence is induced by an element of GL $\left(\mathbb{C}^{m+p}\right)$.

A state-space realization (2.1) is symmetric [5] if $A^{T}=A$ and $C=B^{T}$.
Proposition 2.2. [7, Sect. 3.2] For a controllable and observable linear system with statespace realization $\Sigma$ (2.1), the corresponding center $Z$ is self-adjoint if and only if the realization $\Sigma$ is state-feedback equivalent to a symmetric realization.

The degree of the pole placement map of a symmetric state-space realization is at least 2, because $P_{\Sigma}\left(K^{T}\right)=P_{\Sigma}(K)$. The following corollaries are consequences of Theorem 3.6, of Corollary 3.17, and of Theorem 3.18.
Corollary 2.3. If a controllable and observable linear system with $m=p=2$ has a proper pole placement map, then any state-space realization (2.1) is state-feedback equivalent to a symmetric realization.
Corollary 2.4. Suppose that $\Sigma$ is a state-space realization (2.1) of a controllable and observable linear system with $m=p=3$ whose pole placement map is proper and has degree greater than 1. If the center $Z$ of the pole placement map has dimension at most six, then $\operatorname{dim} Z=6$, and $\Sigma$ is state-feedback equivalent to a symmetric realization.

## 3. Linear projections of the Grassmannian

For a finite-dimensional vector space $W$, let $W^{*}$ be its linear dual. Write $\mathbb{P} W$ for its projective space of one-dimensional linear subspaces. Then $\mathbb{P} W^{*}$ is identified with the set of hyperplanes in $W$. For a vector subspace $Z \subset W, \mathbb{P} Z$ is a linear subspace of $\mathbb{P} W$. We will often write $Z$ for $\mathbb{P} Z$, and $\alpha$ for a nonzero vector in $W$, for the linear subspace $\langle\alpha\rangle$, and for the corresponding point of $\mathbb{P} W$. Context will determine which we intend.

Let $m, n$ be positive integers with $m<n$ and let $V$ be an $n$-dimensional complex vector space. For a proper linear subspace $Z \subsetneq \mathbb{P} \bigwedge^{m} V$, the projection with center $Z$,

$$
\begin{equation*}
\mathbb{P} \bigwedge^{m} V \backslash Z \longrightarrow \mathbb{P}\left(\bigwedge^{m} V\right) / Z \tag{3.1}
\end{equation*}
$$

is induced by the quotient map $\bigwedge^{m} V \rightarrow\left(\bigwedge^{m} V\right) / Z$. This projection is a rational map on $\mathbb{P} \bigwedge^{m} V$ as it is not defined on $Z$.

The Grassmannian $\mathrm{Gr}_{m} V$ of $m$-dimensional subspaces of $V$ is embedded into $\mathbb{P} \bigwedge^{m} V$ via the Plücker embedding which sends an $m$-dimensional space $\Lambda$ with basis $v_{1}, \ldots, v_{m}$ to the span of its Plücker vector $v_{1} \wedge \cdots \wedge v_{m}$, written $\Lambda$. Elements of $\Lambda^{m} V$ representing points of $\operatorname{Gr}_{m} V$ are decomposable. Whether we intend $\Lambda \in \operatorname{Gr}_{m} V$ to be a point of $\mathbb{P} \bigwedge^{m} V$ or a linear subspace of $V$ will often be determined by context.

Let $Z \subset \mathbb{P} \bigwedge^{m} V$ be a linear subspace disjoint from $\operatorname{Gr}_{m} V$. Write $\pi_{Z}$ for the restriction of the corresponding linear projection (3.1) to $\mathrm{Gr}_{m} V$. In 7] such a linear projection was called a generalized Wronski map, a terminology motivated by the following result.

Proposition 3.1 ([7, Prop. 2.3]). The Wronski map $\mathrm{Wr}_{L, m}$ of an nth order linear ODO $L$ is the projection $\pi_{Z}$ with center $Z=X_{L}^{\perp}$, where $X_{L}$ is defined by (1.6).
Remark 3.2. Note that $X_{L}^{\perp}$ is disjoint from the Grassmannian $\operatorname{Gr}_{m} V_{L}$. This is because Wronskians are not identically zero and the formulation (1.5).

Assume that $\operatorname{dim} V=2 m$. A 2-form $\sigma \in \bigwedge^{2} V$ is an element of the tensor space $V \otimes V$. It is a linear map $V^{*} \rightarrow V$ which is given by contraction, $\left.v \mapsto v\right\lrcorner \sigma$. The rank of $\sigma$ is its rank as a linear map, and this is an even integer. When $\sigma$ has rank $2 m$, it is a symplectic form on $V^{*}$. Then corresponding map $V^{*} \rightarrow V$ is an isomorphism and $\sigma$ induces a symplectic form $\sigma^{*} \in \Lambda^{2} V^{*}$ on $V$. The skew-orthogonal complement to $\Lambda \in \operatorname{Gr}_{m} V$ is the linear subspace

$$
\Lambda^{\angle}:=\left\{w \in V \mid \sigma^{*}(w, v)=0 \quad \forall v \in \Lambda\right\} .
$$

This also has dimension $m$, so $\Lambda^{\angle} \in \mathrm{Gr}_{m} V$.
A linear subspace $Z \subset \mathbb{P} \bigwedge^{m} V$ is self-adjoint if there exists a symplectic form $\sigma$ on $V^{*}$ such that

$$
\begin{equation*}
Z \supseteq \mathbb{P}\left(\mathbb{C} \sigma \wedge \bigwedge^{m-2} V\right) \tag{3.2}
\end{equation*}
$$

By [7, Cor. 1.5],

$$
\begin{equation*}
\pi_{Z}\left(\Lambda^{\llcorner }\right)=\pi_{Z}(\Lambda), \quad \forall \Lambda \in \operatorname{Gr}_{m} V \tag{3.3}
\end{equation*}
$$

Thus when $Z$ is self-adjoint, the degree of $\pi_{Z}$ is even and hence exceeds 1 . We address the converse: Does degree of $\pi_{Z}$ exceeding 1 imply that the center $Z$ is self-adjoint?
3.1. Projection from a point. Let $\pi_{\omega}$ be the linear projection with center $\omega \in \mathbb{P} \bigwedge^{m} V$.

Lemma 3.3. Suppose that $Z \subset \mathbb{P} \bigwedge^{m} V$ is a linear subspace disjoint from the Grassmannian $\operatorname{Gr}_{m} V$. For $\Lambda, \Lambda^{\prime} \in \operatorname{Gr}_{m} V$, we have $\pi_{Z}(\Lambda)=\pi_{Z}\left(\Lambda^{\prime}\right)$ if and only if there exists a point $\omega \in Z$ such that $\pi_{\omega}(\Lambda)=\pi_{\omega}\left(\Lambda^{\prime}\right)$ if and only if $Z$ meets the line $\left\langle\Lambda, \Lambda^{\prime}\right\rangle$ in $\mathbb{P} \Lambda^{m} V$ containing the points $\Lambda, \Lambda^{\prime}$.

Proof. If $\pi_{\omega}(\Lambda)=\pi_{\omega}\left(\Lambda^{\prime}\right)$, then for any subspace $Z$ containing $\omega, \pi_{Z}(\Lambda)=\pi_{Z}\left(\Lambda^{\prime}\right)$. For the other direction, suppose that $\pi_{Z}(\Lambda)=\pi_{Z}\left(\Lambda^{\prime}\right)$ with $\Lambda \neq \Lambda^{\prime}$ in $\operatorname{Gr}_{m} V$. Then the line $\left\langle\Lambda, \Lambda^{\prime}\right\rangle$ they span meets $Z$. If $\omega \in\left\langle\Lambda, \Lambda^{\prime}\right\rangle \cap Z$, then $\pi_{\omega}(\Lambda)=\pi_{\omega}\left(\Lambda^{\prime}\right)$.

For a center $Z \subset \mathbb{P} \bigwedge^{m} V$ disjoint from the Grassmannian $\operatorname{Gr}_{m} V$, define

$$
\begin{equation*}
\mathcal{S}_{Z}:=\left\{\Lambda \in \operatorname{Gr}_{m}(V) \mid \exists \Lambda^{\prime} \neq \Lambda \text { such that } \pi_{Z}(\Lambda)=\pi_{Z}\left(\Lambda^{\prime}\right)\right\} \tag{3.4}
\end{equation*}
$$

and for $\omega \in \mathbb{P} \bigwedge^{m} V$, similarly define $\mathcal{S}_{\omega}$. Lemma 3.3 is equivalent to

$$
\begin{equation*}
\mathcal{S}_{Z}=\bigcup_{\omega \in Z} \mathcal{S}_{\omega} \tag{3.5}
\end{equation*}
$$

Remark 3.4. Lemma 3.3 motivates our approach to study the degree of the map $\pi_{Z}$. First, for each $\omega \in Z$, describe all $\Lambda \in \operatorname{Gr}_{m} V$ such that there exist $\Lambda^{\prime} \neq \Lambda$ in $\operatorname{Gr}_{m} V$ with $\pi_{\omega}(\Lambda)=\pi_{\omega}\left(\Lambda^{\prime}\right)$. Then take a union of all such $\Lambda$ for $\omega \in Z$. If this union does not contain an open dense set of $\mathrm{Gr}_{m} V$ then $\pi_{Z}$ has degree 1 .

The group $\mathrm{GL}(V)$ of invertible linear transformations on $V$ acts on $\mathrm{Gr}_{m} V$ and $\mathbb{P} \bigwedge^{m} V$, and for $\omega \in \mathbb{P} \bigwedge^{m} V, \Lambda, \Lambda^{\prime} \in \operatorname{Gr}_{m} V$, and $g \in \operatorname{GL}(V)$, we have

$$
\pi_{\omega}(\Lambda)=\pi_{\omega}\left(\Lambda^{\prime}\right) \quad \text { if and only if } \quad \pi_{g \cdot \omega}(g \cdot \Lambda)=\pi_{g \cdot \omega}\left(g \cdot \Lambda^{\prime}\right)
$$

Therefore, to find the set of pairs $\Lambda, \Lambda^{\prime} \in \operatorname{Gr}_{m}(V)$ with the same image under $\pi_{\omega}$ it is enough to find this set for one representative of the GL( $V$ )-orbit of $\omega$.

Remark 3.5. Suppose that $\operatorname{dim} V=4$. The Grassmannian $\mathrm{Gr}_{2} V \subset \mathbb{P} \bigwedge^{2} V \simeq \mathbb{P}^{5}$ is a quadratic hypersurface. Thus, if $\omega \in \mathbb{P} \bigwedge^{2} V \backslash \mathrm{Gr}_{2} V$, then $\pi_{\omega}: \mathrm{Gr}_{2} V \rightarrow \mathbb{P}\left(\bigwedge^{2} V\right) / \omega \simeq \mathbb{P}^{4}$ has degree two. In particular, $\mathcal{S}_{\omega} \subset \mathrm{Gr}_{2} V$ is dense and therefore has dimension four.

This will be relevant in Section 3.2, where we show that for $\omega \in \mathbb{P} \bigwedge{ }^{3} \mathbb{C}^{6} \backslash \operatorname{Gr}_{3} \mathbb{C}^{6}$, either $\mathcal{S}_{\omega}$ is either zero-dimensional, empty, or four-dimensional, and the last case may be understood to be a consequence of the projection map on $\mathrm{Gr}_{2} V$.

This degree two projection $\mathrm{Gr}_{2} V \rightarrow \mathbb{P}^{4}$ is intrinsically related to symplectic structures.
Theorem 3.6. When $\operatorname{dim} V=4$, any $\sigma \in \mathbb{P} \wedge^{2} V \backslash \mathrm{Gr}_{2} V$ is a symplectic form on $V^{*}$. For $\Lambda, \Lambda^{\prime} \in \operatorname{Gr}_{2} V$ with $\Lambda \neq \Lambda^{\prime}$, we have that $\pi_{\sigma}(\Lambda)=\pi_{\sigma}\left(\Lambda^{\prime}\right)$ if and only if $\Lambda^{\prime}=\Lambda^{\angle}$, the skew-orthogonal complement of $\Lambda$ with respect to the symplectic form $\sigma^{*}$.
3.2. Projection from a point when $m=3$ and $n=6$. Assume that $\operatorname{dim} V=6$. When convenient, we identify $V$ with $\mathbb{C}^{6}$ with the standard basis $\left\{e_{1}, \ldots, e_{6}\right\}$ and let $\left\{e_{1}^{*}, \ldots, e_{6}^{*}\right\}$ be the dual basis for $V^{*}$. Following Remark 3.4, we first study the action of $\operatorname{GL}(V)$ on $\mathbb{P} \bigwedge^{3} V$. The orbits under this action were described by Segre in 1918 [10]. For $i, j, k$, write $e_{i j k}$ for $e_{i} \wedge e_{j} \wedge e_{k}$ and $e_{i j}$ for $e_{i} \wedge e_{j}$. Then $e_{123}$ is the Plücker vector of $\left\langle e_{1}, e_{2}, e_{3}\right\rangle$.

Theorem 3.7 (Segre [10], see also [3]). The action of $\operatorname{GL}(V)$ on $\mathbb{P} \bigwedge^{3} V$ has four orbits $O_{0}, O_{1}, O_{5}$, and $O_{10}$, where $O_{i}$ has codimension i. A normal form for an element $\omega_{i} \in O_{i}$ of each orbit is as follows.
(1) $\omega_{0}=e_{123}+e_{456}$, a point on the line between $e_{123}$ and $e_{456}$.
(2) $\omega_{1}=e_{126}-e_{153}+e_{234}$, a general point in the tangent space to $\mathrm{Gr}_{3} V$ at $e_{123}$.
(3) $\omega_{5}=e_{1} \wedge\left(e_{23}+e_{45}\right)$, a point on the line between $e_{123}$ and $e_{145}$.
(4) $\omega_{10}=e_{123}$, a point on the Grassmannian $\mathrm{Gr}_{3} V$.

Remark 3.8. For a 3-plane $\Lambda \in \mathrm{Gr}_{3} V$, the tangent space $T_{\Lambda} \mathrm{Gr}_{3} V$ to the Grassmannian is $\operatorname{Hom}(\Lambda, V / \Lambda)$. A general point of $T_{\Lambda} \mathrm{Gr}_{3} V$ corresponds to an isomorphism $\Lambda \xrightarrow{\sim} V / \Lambda$. The normal form in Theorem $3.7(2)$ is the point of $T_{e_{123}} \mathrm{Gr}_{3} V$ corresponding to the isomorphism that sends $e_{i}$ to $e_{i+3} \bmod \left\langle e_{1}, e_{2}, e_{3}\right\rangle$. It is the tangent vector at $t=0$ to the curve

$$
\begin{equation*}
\Lambda(t)=e_{1}(t) \wedge e_{2}(t) \wedge e_{3}(t) \tag{3.6}
\end{equation*}
$$

where $e_{i}(t)=e_{i}+t e_{i+3}$ for $i=1,2,3$.
Remark 3.9. The tangent variety $\mathcal{T} X$ of a projective variety $X \subset \mathbb{P}^{N}$ is the union of all lines tangent to $X$. The orbits from Theorem 3.7 are described geometrically as follows.
(1) The orbit $O_{0}$ is the complement of the tangent variety $\mathcal{T} \mathrm{Gr}_{3} V$ of $\mathrm{Gr}_{3} V \subset \mathbb{P} \bigwedge^{3} V$.
(2) Let $\mathcal{T}_{1}$ be the union of all lines in $\mathbb{P} \bigwedge^{3} V$ connecting two points in $\mathrm{Gr}_{3} V$ whose corresponding subspaces in $V$ have nonzero intersection. Then

$$
\mathrm{Gr}_{3} V \subset \mathcal{T}_{1} \subset \mathcal{T} \mathrm{Gr}_{3} V
$$

and $O_{1}$ is the complement of $\mathcal{T}_{1}$ in $\mathcal{T} \mathrm{Gr}_{3} V$.
(3) The orbit $O_{5}$ is the complement of $\mathrm{Gr}_{3} V$ in $\mathcal{T}_{1}$.
(4) The orbit $O_{10}$ is $\mathrm{Gr}_{3} V$.

We describe $\mathcal{S}_{\omega}$ for $\omega \in \mathbb{P} \bigwedge^{3} V \backslash \operatorname{Gr}_{3} V$.
Proposition 3.10. Let $\omega \in \mathbb{P} \bigwedge^{3} V \backslash \operatorname{Gr}_{3} V$. Then
(1) If $\omega \in O_{0}$, then $\mathcal{S}_{\omega}$ is finite and $\operatorname{dim} \mathcal{S}_{\omega}=0$
(2) If $\omega \in O_{1}$, then $\pi_{\omega}$ is injective, so that $\mathcal{S}_{\omega}=\emptyset$.
(3) If $\omega \in O_{5}$, then $\operatorname{dim} \mathcal{S}_{\omega}=4$.

Proof. By Lemma 3.3, $\Lambda \in \mathcal{S}_{\omega}$ if and only if there is a $\Lambda^{\prime} \in \operatorname{Gr}_{3} V$ with $\Lambda \neq \Lambda^{\prime}$ such that $\omega \in\left\langle\Lambda, \Lambda^{\prime}\right\rangle$, the line in $\mathbb{P} \Lambda^{2} V$ spanned by the Plücker vectors of $\Lambda$ and $\Lambda^{\prime}$.

Let $\Lambda \neq \Lambda^{\prime}$ be distinct 3-planes in $\operatorname{Gr}_{3} V$ and $\omega \in\left\langle\Lambda, \Lambda^{\prime}\right\rangle \backslash \operatorname{Gr}_{3} V$. By Remark 3.9(2), $\omega \notin O_{1}$, which proves (2). We argue by the dimension of $\Lambda \cap \Lambda^{\prime}$. If $\operatorname{dim} \Lambda \cap \Lambda^{\prime}=0$, then $\omega \in O_{0}$, by Theorem 3.7(1). Since $\operatorname{dim} \mathrm{Gr}_{3} V=9$ and $\operatorname{dim} \mathbb{P} \wedge^{3} V=19$, dimensioncounting shows that for a point $\omega \in O_{0}, \mathcal{S}_{\omega}$ is zero-dimensional and hence finite, proving (1). If $\operatorname{dim} \Lambda \cap \Lambda^{\prime}=1$, then $\omega \in O_{5}$, by Theorem 3.7(3). Statement (3) is Lemma 3.14 below. If $\operatorname{dim} \Lambda \cap \Lambda^{\prime}=2$, then $\left\langle\Lambda, \Lambda^{\prime}\right\rangle \subset \mathrm{Gr}_{3} V$.

An element $\omega \in \bigwedge^{3} V$ defines two linear maps

$$
\begin{array}{rlrl}
\wedge \omega: & V & \lrcorner \wedge^{4} V \\
& v & \lrcorner \omega: & V^{*} \longrightarrow \bigwedge^{2} V \\
& v \longmapsto v\lrcorner \omega
\end{array}
$$

Lemma 3.11. If $\omega \in O_{5}$, then both $\wedge \omega$ and $\lrcorner \omega$ have one-dimensional kernels.
Proof. Computations using the normal form of $\omega \in O_{5}$ given by Theorem 3.7(3) show that the kernel of $\wedge \omega_{5}$ is $\left\langle e_{1}\right\rangle$ and the kernel of $\lrcorner \omega_{5}$ is $\left\langle e_{6}^{*}\right\rangle$.

For $\omega \in O_{5}$, write $\alpha_{\omega} \in \mathbb{P} V$ for the kernel of $\wedge \omega$ and $A_{\omega} \in \mathbb{P} V^{*}$ for the kernel of $\lrcorner \omega$. We regard $\alpha_{\omega}$ as a 1-dimensional linear subspace of $V$ and $A_{\omega}$ as a hyperplane in $V$.
Corollary 3.12. Let $\omega \in O_{5}$. Then $\alpha_{\omega} \subset A_{\omega}$, $A_{\omega}$ is the smallest subspace $W$ of $V$ such that $\omega \in \Lambda^{3} W$, and if $\pi_{\omega}(\Lambda)=\pi_{\omega}\left(\Lambda^{\prime}\right)$ for $\Lambda \neq \Lambda^{\prime} \in \operatorname{Gr}_{3} V$, then $\alpha_{\omega}=\Lambda \cap \Lambda^{\prime}$ and $A_{\omega}=\left\langle\Lambda, \Lambda^{\prime}\right\rangle$ (their span in $V$ ). Finally, there is an indecomposable 2-form $\sigma \in \Lambda^{2} A_{\omega}$ such that $\omega=\alpha_{\omega} \wedge \sigma$, with $\alpha_{\omega}$ and $\sigma$ well-defined up to scalars.
Proof. By the normal form of Theorem 3.7(3) and the proof of Proposition 3.10, $\alpha_{\omega}=$ $\Lambda \cap \Lambda^{\prime}$, so that $\left\langle\Lambda, \Lambda^{\prime}\right\rangle$ is a hyperplane in $V$. Since $\omega, \Lambda, \Lambda^{\prime}$ are collinear in $\mathbb{P} \Lambda^{3} V, \omega \in \Lambda^{3} A_{\omega}$. For any four dimensional subspace $W$ of $V, \bigwedge^{3} W \subset \operatorname{Gr}_{3} V$, which shows the minimality of $A_{\omega}$. The last statement follows from these identifications and Theorem 3.7(3).

By Corollary 3.12, if $\omega \in O_{5}$, then $\omega \in \mathbb{C} \alpha_{\omega} \wedge \bigwedge^{2} A_{\omega} \simeq \bigwedge^{2}\left(A_{\omega} / \alpha_{\omega}\right)$. Notice that $\Lambda \mapsto \Lambda / \alpha_{\omega}$ identifies the Schubert variety

$$
\begin{equation*}
\Omega_{\omega}:=\left\{\Lambda \in \operatorname{Gr}_{3} V \mid \alpha_{\omega} \in \Lambda \subset A_{\omega}\right\} \tag{3.7}
\end{equation*}
$$

with $\operatorname{Gr}_{2}\left(A_{\omega} / \alpha_{\omega}\right) \simeq \operatorname{Gr}_{2} \mathbb{C}^{4}$.
Let $\operatorname{Fl}(1,5 ; V) \subset \mathbb{P} V \times \mathbb{P} V^{*}$ be the flag variety whose points are pairs $(\alpha, A)$ with $\alpha \subset A$; the one-dimensional linear subspace $\alpha$ lies in the hyperplane $A$. The projection of $\mathrm{Fl}(1,5 ; V)$ to each projective space factor is a $\mathbb{P}^{4}$ bundle. Let $L \rightarrow \mathrm{Fl}(1,5 ; V)$ be the subbundle of $\mathbb{P} \bigwedge^{3} V \times \operatorname{Fl}(1,5 ; V)$ whose fiber over $(\alpha, A)$ is $\mathbb{P}\left(\alpha \wedge \bigwedge^{2} A\right) \simeq \mathbb{P}^{5}$. The Schubert variety $\Omega_{\omega}$ (3.7) depends only upon the flag $\alpha_{\omega} \subset A_{\omega}$ and it lies in $\mathbb{P}\left(\alpha_{\omega} \wedge \bigwedge^{2} A_{\omega}\right)$. Write $\Omega(\alpha, A)$ for the Schubert variety corresponding to the flag $\alpha \subset A$. A consequence of this definition and Corollary 3.12 is the following.
Corollary 3.13. For $\omega \in O_{5}$, the map $\omega \mapsto\left(\alpha_{\omega}, A_{\omega}\right) \in \mathrm{Fl}(1,5 ; V)$ realizes $O_{5}$ as a bundle over $\mathrm{Fl}(1,5 ; V)$, which is a dense open subset of $L$. The points in the fiber above $(\alpha, A)$ consist of points in $\mathbb{P}\left(\alpha \wedge \bigwedge^{2} A\right)$ in the complement of $\Omega(\alpha, A)$.

Lemma 3.14. For $\omega \in O_{5}, \mathcal{S}_{\omega}$ is a dense subset of $\Omega_{\omega}$ and therefore has dimension four.
Proof. In the proof of Corollary 3.12, we observed that if $\Lambda \neq \Lambda^{\prime}$ are 3-planes in $\operatorname{Gr}_{3} V$ with $\pi_{\omega}(\Lambda)=\pi_{\omega}\left(\Lambda^{\prime}\right)$, then $\alpha_{\omega} \subset \Lambda \subset A_{\omega}$. This implies that $\mathcal{S}_{\omega} \subset \Omega_{\omega}$.

Consider the restriction of $\pi_{\omega}$ to $\Omega_{\omega} \subset \operatorname{Gr}_{3} V$. Both $\omega$ and $\Omega_{\omega}$ lie in $\mathbb{P}\left(\alpha_{\omega} \wedge \bigwedge^{2} A_{\omega}\right)$, which is identified with $\mathbb{P} \bigwedge^{2}\left(A_{\omega} / \alpha_{\omega}\right)$. Write $\omega=\alpha_{\omega} \wedge \sigma$ with $\sigma \in \bigwedge^{2}\left(A_{\omega} / \alpha_{\omega}\right)$. Identifying $\Omega_{\omega}$ with $\operatorname{Gr}_{2}\left(A_{\omega} / \alpha_{\omega}\right)$, the map $\pi_{\omega}$ on $\Omega_{\omega}$ becomes $\pi_{\sigma}$, which has degree 2, by Remark 3.5, This completes the proof.

Theorem 3.6 and the proof of Lemma 3.14 imply the following corollary.
Corollary 3.15. Let $\omega=\alpha \wedge \sigma \in O_{5}$. If $\Lambda \neq \Lambda^{\prime}$ are 3-planes then $\pi_{\omega}(\Lambda)=\pi_{\omega}\left(\Lambda^{\prime}\right)$ if and only if $\Lambda, \Lambda^{\prime} \in \Omega_{\omega}$ and $\Lambda^{\prime} / \alpha=(\Lambda / \alpha)^{L_{\sigma}}$.
3.3. The center has dimension less than five. By Proposition 3.10 and (3.5), if $Z \subset \mathbb{P} \bigwedge^{3} V$ is a linear subspace that does not meet the Grassmannian $\mathrm{Gr}_{3} V$, then

$$
\begin{equation*}
\mathcal{S}_{Z}=\bigcup_{\omega \in Z \cap O_{0}} \mathcal{S}_{\omega} \cup \bigcup_{\omega \in Z \cap O_{5}} \mathcal{S}_{\omega} \tag{3.8}
\end{equation*}
$$

and it follows that

$$
\begin{equation*}
\operatorname{dim} \mathcal{S}_{Z} \leq \max \left\{\operatorname{dim}\left(Z \cap O_{0}\right), \operatorname{dim}\left(Z \cap O_{5}\right)+4\right\} \tag{3.9}
\end{equation*}
$$

Since $\operatorname{dim} \operatorname{Gr}_{3} V=9$, the last relation implies the following result.
Theorem 3.16. If $Z \subset \mathbb{P} \bigwedge^{3} V$ is a linear subspace that does not meet the Grassmannian $\mathrm{Gr}_{3} V$, $\operatorname{dim} Z<9$, and $\operatorname{dim} Z \cap O_{5} \leq 4$, then $\pi_{Z}$ has degree 1 on $\mathrm{Gr}_{3} V$.

Proof. From the assumptions and (3.9), we have that $\operatorname{dim} \mathcal{S}_{Z} \leq 8$. Thus $\operatorname{Gr}_{3} V \backslash \mathcal{S}_{Z}$ contains a nonempty Zariski open set and therefore $\pi_{Z}$ has degree 1 .

Corollary 3.17. If $Z$ does not meet the Grassmannian $\operatorname{Gr}_{3} V$ and $\operatorname{dim} Z \leq 4$, then $\pi_{Z}$ has degree 1.
3.4. Five-dimensional center. Let $Z \subset \mathbb{P} \bigwedge^{3} V$ be a linear subspace such that the following three conditions hold,
(i) $\operatorname{dim} Z=5$,
(ii) $\operatorname{dim} Z \cap O_{5} \geq 5$, which together with (i) is equivalent to $\operatorname{dim} Z \cap O_{5}=5$, and
(iii) $Z$ does not meet the Grassmannian $\operatorname{Gr}_{3} V$, so that $Z \subset O_{5}$.

We establish the following result.
Theorem 3.18. If $Z \subset \mathbb{P} \bigwedge^{3} V$ is a linear subspace that does not meet the Grassmannian $\operatorname{Gr}_{3} V, \operatorname{dim} Z=5$, and the degree of $\pi_{Z}$ exceeds 1 , then $Z$ is self-adjoint.

The hypotheses imply that $Z \subset O_{5}$. We begin with a lemma about lines in $O_{5}$. For this, $\omega_{i}, \sigma_{i}, \rho_{i}, \alpha_{i}, v_{i}, w_{i}$ for $i=1,2$, and $v$ are vectors and not points in projective space.

Lemma 3.19. Assume that $\omega_{1}, \omega_{2} \in O_{5}$ and the line they span lies in $O_{5}$. If $\omega_{i}=\alpha_{i} \wedge \sigma_{i}$ for $i=1,2$ as in Corollary 3.12, then one of the following cases holds.
(1) $\left\langle\alpha_{1}\right\rangle=\left\langle\alpha_{2}\right\rangle$.
(2) $\alpha_{1}$ and $\alpha_{2}$ are linearly independent and $\left\langle\sigma_{1}\right\rangle \equiv\left\langle\sigma_{2}\right\rangle \bmod \left\langle\alpha_{1}, \alpha_{2}\right\rangle$. There is a 2-form $\sigma \in \bigwedge^{2} V$ such that, up to a scalar factor, $\omega_{i}=\alpha_{i} \wedge \sigma$ for $i=1,2$.
(3) There exist $v, w_{1}, w_{2}, v_{1}, v_{2} \in V$ where $\alpha_{1}, \alpha_{2}, v, v_{1}, v_{2}$ are linearly independent with $\left\langle v, v_{1}, v_{2}\right\rangle=\left\langle v, v_{1}, w_{1}\right\rangle=\left\langle v, v_{2}, w_{2}\right\rangle$ such that
$\omega_{1}=\alpha_{1} \wedge\left(\alpha_{2} \wedge w_{1}+v \wedge v_{1}\right) \quad$ and $\quad \omega_{2}=\alpha_{2} \wedge\left(\alpha_{1} \wedge w_{2}+v \wedge v_{2}\right)$.
Proof. Suppose that (1) does not hold, so that $\alpha_{1}$ and $\alpha_{2}$ are linearly independent. Let us suppose that $\alpha_{1}=e_{1}$ and $\alpha_{2}=e_{2}$. Let $U:=\left\langle e_{1}, e_{2}\right\rangle$ and $W=\left\langle e_{3}, \ldots, e_{6}\right\rangle \simeq \mathbb{C}^{4}$, which are transversal. We express $\sigma_{1}, \sigma_{2}$ in terms of $e_{2}$ and $e_{1}$ respectively. We have

$$
\begin{align*}
& \omega_{1}=e_{1} \wedge \sigma_{1}=e_{1} \wedge\left(e_{2} \wedge w_{1}+\rho_{1}\right)  \tag{3.10}\\
& \omega_{2}=e_{2} \wedge \sigma_{2}=e_{2} \wedge\left(e_{1} \wedge w_{2}+\rho_{2}\right)
\end{align*}
$$

where $w_{1}, w_{2} \in W$, and $\rho_{1}, \rho_{2} \in \bigwedge^{2} W$ are the terms in $\sigma_{1}, \sigma_{2}$ that do not contain $e_{2}$ and $e_{1}$ respectively. For $i=1,2$, since $\sigma_{i}$ is indecomposable, neither $\rho_{i}$ nor $w_{i} \wedge \rho_{i}$ is zero.

Let $\lambda, \mu \in \mathbb{C}$ be nonzero. Since $\lambda \omega_{1}+\mu \omega_{2} \in O_{5}$, it has the form $\alpha \wedge \sigma$, where $0 \neq \alpha \in V$ is defined up to a scalar by $\alpha \wedge\left(\lambda \omega_{1}+\mu \omega_{2}\right)=0$. Let us write $\alpha=a e_{1}+b e_{2}+v$, where $v \in W$. The vector $v$ and the coefficients $a$ and $b$ are functions of $\lambda$ and $\mu$, up to a common scalar, and at least one of $a, b$, and $v$ is nonzero. We use (3.10) to rewrite $\alpha \wedge\left(\lambda \omega_{1}+\mu \omega_{2}\right)=0$ as

$$
\left(a e_{1}+b e_{2}+v\right) \wedge\left(\lambda e_{12} \wedge w_{1}+\lambda e_{1} \wedge \rho_{1}-\mu e_{12} \wedge w_{2}+\mu e_{2} \wedge \rho_{2}\right)=0
$$

Recall that $e_{12}=e_{1} \wedge e_{2}$. Expanding gives

$$
\begin{equation*}
e_{12} \wedge\left(\mu a \rho_{2}-\lambda b \rho_{1}+v \wedge\left(\lambda w_{1}-\mu w_{2}\right)\right)-\lambda e_{1} \wedge v \wedge \rho_{1}-\mu e_{2} \wedge v \wedge \rho_{2}=0 \tag{3.11}
\end{equation*}
$$

These summands lie in $e_{12} \wedge \bigwedge^{2} W, e_{1} \wedge \bigwedge^{3} W$, and $e_{2} \wedge \bigwedge^{3} W$, respectively, and are therefore linearly independent. This gives the following three equations,

$$
\begin{align*}
\mu a \rho_{2}-\lambda b \rho_{1} & =v \wedge\left(\mu w_{2}-\lambda w_{1}\right),  \tag{3.12}\\
v \wedge \rho_{1} & =0, \quad \text { and }  \tag{3.13}\\
v \wedge \rho_{2} & =0 \tag{3.14}
\end{align*}
$$

The last two are linear equations for $v \in W$. Note that each $\rho_{i}$ is either decomposable (lies in $\operatorname{Gr}_{2} W$ ) or indecomposable, corresponding to having rank 2 or rank 4. If either $\rho_{1}$ or $\rho_{2}$ is indecomposable and hence of rank 4 , then $v=0$ is the only solution.

Suppose first that $v=0$ is a solution to (3.13) and (3.14). Then (3.12) implies that $\left\langle\rho_{1}\right\rangle=\left\langle\rho_{2}\right\rangle$. (We cannot have $a b=0$, for then (3.12) and $(a, b) \neq(0,0)$ implies that one of $\rho_{1}$ or $\rho_{2}$ is zero.) Scaling $\omega_{1}$ and $\omega_{2}$ if necessary, $\rho_{1}=\rho_{2}=\rho$, and using (3.10) we may set $\sigma=e_{2} \wedge w_{1}+e_{1} \wedge w_{2}+\rho$. Then Case (2) holds.

Suppose that (3.13) and (3.14) admit a nonzero solution, $v$. Thus $\rho_{1}$ and $\rho_{2}$ are each decomposable, and they have the form $\rho_{i}=v_{i} \wedge v$, for nonzero $v_{1}, v_{2} \in W$. Then

$$
\begin{equation*}
\lambda \omega_{1}+\mu \omega_{2}=e_{12} \wedge\left(\lambda w_{1}-\mu w_{2}\right)+\left(\lambda e_{1} \wedge v_{1}+\mu e_{2} \wedge v_{2}\right) \wedge v \tag{3.15}
\end{equation*}
$$

This is indecomposable for $(\lambda, \mu) \neq(0,0)$.
Suppose that $\left\langle\rho_{1}\right\rangle=\left\langle\rho_{2}\right\rangle$, which corresponds to a 2-plane $H \subset W$. Then (3.12) for all $\lambda, \mu$ implies that $w_{1}, w_{2} \in H$. In particular, $\rho_{1}=v^{\prime} \wedge w_{1}$, for some $v^{\prime} \in H$. But then $\sigma_{1}=\left(e_{1}+v^{\prime}\right) \wedge w_{1}$, which contradicts its being indecomposable.

Now suppose that $\rho_{1}$ and $\rho_{2}$ are linearly independent. If $H_{i} \in \mathrm{Gr}_{2} W$ is the 2-plane corresponding to $\rho_{i}$, then $\langle v\rangle=H_{1} \cap H_{2}$, and thus $v$ is independent of $\lambda, \mu$ (up to a scalar), and we also see that $v, v_{1}, v_{2}$ are linearly independent. We establish Case (3) by showing that $\left\langle v, v_{1}, v_{2}\right\rangle=\left\langle v, v_{1}, w_{1}\right\rangle=\left\langle v, v_{2}, w_{2}\right\rangle$.

Consider the 2 -forms $\mu a \rho_{2}-\lambda b \rho_{1}$ for all $\lambda, \mu$. If these are all 0 , then $a=b=0$ as $\rho_{1}$ and $\rho_{2}$ are linearly independent. Then (3.12) implies that $v, w_{1}, w_{2}$ are proportional, which implies that $\sigma_{1}$ and $\sigma_{2}$ are decomposable, a contradiction.

Thus, for general $\lambda$, $\mu$, the 2 -form $\mu a \rho_{2}-\lambda b \rho_{1} \in \Lambda^{2}\left\langle v, v_{1}, v_{2}\right\rangle$ is nonzero. By (3.12), for all $\lambda, \mu$ we have that $\mu w_{2}-\lambda w_{1} \in\left\langle v, v_{1}, v_{2}\right\rangle$. Since $\sigma_{1}$ is indecomposable, $w_{1}$ is independent of $v, v_{1}$, and the same holds for $w_{2}, v, v_{2}$, which completes the proof.

A line in $O_{5}$ has type $(i)$ if it satisfies condition $(i)$ of Lemma 3.19.

Corollary 3.20. Let $\ell \subset O_{5}$ be a line. If $\ell$ has type (1), then $\alpha_{\omega}$ is the same point in $\mathbb{P} V$ for every $\omega \in \ell$. If $\ell$ has type (3), then $A_{\omega} \in \mathbb{P} V^{*}$ is the same hyperplane for every $\omega \in \ell$.

Proof. The claim about lines of type (1) follows from their definition and Lemma3.19(1). Suppose $\ell$ has type (3). Recall that for $\omega \in O_{5}, A_{\omega}$ is the unique hyperplane of $V$ with $\omega \in \bigwedge^{3} A_{\omega}$. By the normal form for points on a line of type (3) from Lemma 3.19(3), we see that $A_{\omega}=\left\langle\alpha_{1}, \alpha_{2}, v, v_{1}, v_{2}\right\rangle$ for all $\omega \in \ell$.

Now let us define

$$
\begin{align*}
& E_{Z}:=\left\{\alpha_{\omega} \in \mathbb{P} V \mid \text { for } \omega \in Z\right\}, \text { and } \\
& F_{Z}:=\left\{A_{\omega} \in \mathbb{P} V^{*} \mid \text { for } \omega \in Z\right\} . \tag{3.16}
\end{align*}
$$

Lemma 3.21. If $Z$ is a linear subspace of $\mathbb{P} \bigwedge^{3} V$ of dimension five with $Z \subset O_{5}$ such that the degree of $\pi_{Z}$ exceeds 1 , then $E_{Z}=\mathbb{P} V$ and $F_{Z}=\mathbb{P} V^{*}$.

The proof we give uses the following fact about maps between projective spaces.
Proposition 3.22. If $\phi: \mathbb{P}^{r} \rightarrow \mathbb{P}^{r}$ is a nonconstant map, then it is onto.
Proof. Suppose that $\phi\left(\mathbb{P}^{r}\right) \neq \mathbb{P}^{r}$. Since the image is closed, we may compose $\phi$ with the linear projection from a point $x \notin \phi\left(\mathbb{P}^{r}\right)$, obtaining a map $\psi: \mathbb{P}^{r} \rightarrow \mathbb{P}^{r-1}$. This is given by $r$ homogeneous forms $f_{1}, \ldots, f_{r}$ of the same degree $d$ with no common zeroes; for $z \in \mathbb{P}^{r}$, $\psi(z)=\left[f_{1}(z), \ldots, f_{r}(z)\right]$. We must have $d>0$, as $\phi$ and hence $\psi$ is nonconstant. This contradicts $f_{1}, \ldots, f_{r}$ having no common zeroes, as $r$ forms of degree $d$ define a subvariety in $\mathbb{P}^{r}$ of codimension at most $r$.

Proof of Lemma 3.21. Recall the map $O_{5} \rightarrow \operatorname{Fl}(1,5 ; V)$ that sends $\omega$ to the flag $\alpha_{\omega} \subset A_{\omega}$. Then $E_{Z}$ is the image of $Z$ under the further map to $\mathbb{P} V$ and $F_{Z}$ is its image under the map to $\mathbb{P} V^{*}$. As $Z, \mathbb{P} V$, and $\mathbb{P} V^{*}$ are all projective spaces of dimension five, for each of $\mathbb{P} V$ and $\mathbb{P} V^{*}$, the image of $Z$ is either a point, or the map is surjective.

By Corollary 3.12, if $\Lambda \in \mathcal{S}_{\omega}$ for $\omega \in O_{5}$, then $\alpha_{\omega} \subset \Lambda \subset A_{\omega}$. If $E_{Z}$ is a point $\alpha$, then $\mathcal{S}_{Z} \subset\left\{\Lambda \in \operatorname{Gr}_{3} V \mid \alpha \subset \Lambda\right\}$, which is a proper subvariety of $\mathrm{Gr}_{3} V$, and thus $\pi_{Z}$ has degree 1. Similarly, if $F_{Z}$ is a point, then $\pi_{Z}$ has degree 1.

We have another technical lemma.
Lemma 3.23. Given $k+1$ linearly independent elements $\left\{\alpha_{i}\right\}_{i=1}^{k+1}$ in $V$, if $\rho \in \Lambda^{2} V$ satisfies

$$
\begin{equation*}
\rho \equiv 0 \quad \bmod \left\langle\alpha_{i}, \alpha_{k+1}\right\rangle, \quad \forall i \in\{1, \ldots, k\} \tag{3.17}
\end{equation*}
$$

then up to a nonzero constant,

$$
\rho \equiv\left\{\begin{array}{rrr}
\alpha_{1} \wedge \alpha_{2} & \bmod \alpha_{k+1} & k=2  \tag{3.18}\\
0 & \bmod \alpha_{k+1} & k>2
\end{array}\right.
$$

Proof. From (3.17) it follows that for any $i$ there exist $\beta_{i}, \gamma_{i}$ in $V$ such that

$$
\rho=\alpha_{i} \wedge \beta_{i}+\alpha_{k+1} \wedge \gamma_{i}
$$

Therefore for any $1 \leq i \neq j \leq k$

$$
\begin{equation*}
\alpha_{i} \wedge \beta_{i}-\alpha_{j} \wedge \beta_{j}+\alpha_{k+1} \wedge\left(\gamma_{i}-\gamma_{j}\right)=0 \tag{3.19}
\end{equation*}
$$

Since $\alpha_{i}, \alpha_{j}, \alpha_{k+1}$ are linearly independent, by the classical Cartan lemma we have

$$
\begin{equation*}
\beta_{i} \in\left\langle\alpha_{i}, \alpha_{j}, \alpha_{k+1}\right\rangle \tag{3.20}
\end{equation*}
$$

If $k>2$, then for any $i \in\{1, \ldots, k\}$, as there is more than one choice of $j \in\{1, \ldots, k\} \backslash$ $\{i\}$ in (3.20), we obtain that

$$
\begin{equation*}
\beta_{i} \in\left\langle\alpha_{i}, \alpha_{k+1}\right\rangle \tag{3.21}
\end{equation*}
$$

which implies that $\rho \equiv 0 \bmod \alpha_{k+1}$.
If $k=2$ then again by (3.20) and (3.19), we have that

$$
\beta_{1}=c \alpha_{2} \quad \bmod \left\langle\alpha_{1}, \alpha_{3}\right\rangle, \quad \beta_{2}=-c \alpha_{1} \bmod \left\langle\alpha_{2}, \alpha_{3}\right\rangle,
$$

for some constant $c$, which completes the proof.
With these lemmas in place, we give the proof of Theorem 3.18,
Proof of Theorem 3.18. For this proof, $Z \subset \bigwedge^{3} V$ is a linear subspace of dimension six and $\mathbb{P} Z$ is its image in $\mathbb{P} \bigwedge^{3} V$. By (3.2), to show that $Z$ is self-adjoint, we must produce a form $\sigma \in \bigwedge^{2} V$ such that $Z=V \wedge \mathbb{C} \sigma$.

By Lemma 3.21, the maps from $\mathbb{P} Z$ to each of $\mathbb{P} V$ and $\mathbb{P} V^{*}$ are surjective. Thus we may choose a basis $\left\{\omega_{i}\right\}_{i=1}^{6}$ for $Z$ whose images in each of $\mathbb{P} V$ and $\mathbb{P} V^{*}$ are linearly independent. For each $i=1, \ldots, 6$, write $\omega_{i}=\alpha_{i} \wedge \sigma_{i}$, so that $\alpha_{i}$ is the image of $\omega_{i}$ in $\mathbb{P} V$ and let $A_{i}$ be its image in $\mathbb{P} V^{*}$. Then $\left\{\alpha_{i} \mid i=1, \ldots, 6\right\}$ form a basis for $V$ and $\left\{A_{i} \mid i=1, \ldots, 6\right\}$ form a basis for $V^{*}$. These vectors $\omega_{i}, \sigma_{i}$, and $\alpha_{i}$ are only defined up to scalar multiples, so we may freely replace any by a scalar multiple.

By Corollary 3.20, no line $\left\langle\omega_{i}, \omega_{j}\right\rangle$ for $i \neq j$ has type (1) or (3), as $\alpha_{i}$ and $\alpha_{j}$ are independent and $A_{i} \neq A_{j}$. Therefore, they all have type (2). By Lemma 3.19)(2), there exists $\sigma \in \bigwedge^{2} V$ such that $\alpha_{i} \wedge \sigma_{i}=\alpha_{i} \wedge \sigma$ for $i=1,2$. Applying Lemma 3.19(2) to $\left\langle\omega_{1}, \omega_{3}\right\rangle$ and to $\left\langle\omega_{2}, \omega_{3}\right\rangle$, after replacing $\sigma$ and $\sigma_{3}$ (and possibly $\alpha_{1}, \alpha_{2}, \alpha_{3}$ ) by scalar multiples,

$$
\sigma-\sigma_{3} \equiv 0 \quad \bmod \left\langle\alpha_{i}, \alpha_{3}\right\rangle, \text { for } i=1,2
$$

By Lemma 3.23 for $\rho=\sigma-\sigma_{3}$ and $k=2$, we have

$$
\sigma-\sigma_{3} \equiv c \alpha_{1} \wedge \alpha_{2} \quad \bmod \left\langle\alpha_{3}\right\rangle
$$

for some constant $c$. Consequently, there exists $\beta \in V$ such that

$$
\sigma-c \alpha_{1} \wedge \alpha_{2}=\sigma_{3}+\alpha_{3} \wedge \beta
$$

Setting $\widetilde{\sigma}:=\sigma-c \alpha_{1} \wedge \alpha_{2}$ we get

$$
\begin{equation*}
\alpha_{i} \wedge \sigma_{i}=\alpha_{i} \wedge \tilde{\sigma}, \text { for } \quad i=1,2,3 \tag{3.22}
\end{equation*}
$$

Since the lines between $\omega_{4}=\alpha_{4} \wedge \sigma_{4}$ and $\omega_{i}$ for $i=1,2,3$ have type (2), Lemma 3.19(2) implies that after multiplying by scalars, we have

$$
\begin{equation*}
\widetilde{\sigma} \equiv \sigma_{4} \bmod \left\langle\alpha_{i}, \alpha_{4}\right\rangle, \text { for } \quad i=1,2,3 \tag{3.23}
\end{equation*}
$$

Then, by Lemma 3.23 with $\rho=\widetilde{\sigma}-\sigma_{4}$ and $k=3$ we have

$$
\tilde{\sigma} \equiv \sigma_{4} \bmod \alpha_{4},
$$

which implies that in addition to (3.22) we have $\alpha_{4} \wedge \sigma_{4}=\alpha_{4} \wedge \widetilde{\sigma}$. The same arguments applied to $\alpha_{5}$ and $\alpha_{6}$ imply that for all $1 \leq i \leq 6$, we have $\omega_{i}=\alpha_{i} \wedge \widetilde{\sigma}$. As $\alpha_{1}, \ldots, \alpha_{6}$ form a basis for $V$, we have that $Z=V \wedge \mathbb{C} \tilde{\sigma}$, which implies that it is self-adjoint, and completes the proof of Theorem 3.18.

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Yanhe Huang, Department of Mathematics, University of California, Berkeley, 970 Evans Hall, Berkeley, CA 94720-3840 USA

E-mail address: yanhe_huang@berkeley.edu
George Petroulakis, Athens, Greece
E-mail address: Georgios.Petroulakis.1@city.ac.uk
Frank Sottile, Department of Mathematics, Texas A\&M University, College Station, Texas 77843, USA

E-mail address: sottile@math.tamu.edu
URL: http://www.math.tamu.edu/~ sottile
Igor Zelenko, Department of Mathematics, Texas A\&M University, College Station, TEXAS 77843, USA

E-mail address: zelenko@math.tamu.edu
URL: http://www.math.tamu.edu/~zelenko


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